Topology of Numbers

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Preface

This book provides an introduction to Number Theory from a point of view that is more geometric than is usual for the subject, inspired by the idea that pictures are often a great aid to understanding. The title of the book, *Topology of Numbers*, is intended to express this visual slant, where we are using the term “Topology” with its general meaning of “the spatial arrangement and interlinking of the components of a system”.

A central geometric theme of the book is a certain two-dimensional figure known as the Farey diagram, discovered by Adolf Hurwitz in 1894, which displays certain relationships between rational numbers beyond just their usual distribution along the one-dimensional real number line. Among the many things the diagram elucidates that will be explored in the book are Pythagorean triples, the Euclidean algorithm, Pell’s equation, continued fractions, Farey sequences, and two-by-two matrices with integer entries and determinant $\pm 1$.

But most importantly for this book, the Farey diagram can be used to study quadratic forms $Q(x, y) = ax^2 + bxy + cy^2$ in two variables with integer coefficients via John Conway’s marvelous idea of the topograph of such a form. The origins of the wonderfully subtle theory of quadratic forms $Q(x, y)$ can be traced back to ancient times. A big boost was provided in the 1600s by Fermat, but it was only in the period 1750-1800 that Euler, Lagrange, Legendre, and especially Gauss were able to uncover the main features of the theory.

The principal goal of the book is to present an accessible introduction to this theory from a geometric viewpoint that complements the usual purely algebraic approach. Prerequisites for reading the book are fairly minimal, hardly going beyond high school mathematics for the most part. One topic that often forms a significant part of elementary number theory courses is congruences modulo an integer $n$. It would be helpful if the reader has already seen and used these a little, but we will not develop congruence theory as a separate topic and will instead just use congruences as the need arises, proving whatever nontrivial facts are required including several of the basic ones that form part of a standard introductory number theory course. Among these is quadratic reciprocity, where we give Eisenstein’s classical proof since it involves some geometry.

The high point of the basic theory of quadratic forms $Q(x, y)$ is the class group first constructed by Gauss. This can be defined purely in terms of quadratic forms, which is how it was first presented, or by means of Kronecker’s notion of ideals introduced some 75 years after Gauss’s work. For subsequent developments and generalizations the viewpoint of ideals has proven to be central to all of modern algebra. In this book we present both approaches to the class group, first the older version just in terms of forms, then the later version using ideals.
Items still to be added to the book:

— More exercises, particularly in the later chapters.

— A new section in Chapter 6 going into more detail about rational points on quadratic curves, including Legendre’s theorem on solvability of the Diophantine equation $ax^2 + by^2 + cz^2 = 0$. This will be applied to give a second proof, at least in the case of fundamental discriminants, of Gauss’s theorem that the principal genus of forms of a given discriminant consists of the elements in the class group that are squares of other elements. The proof currently included relies on Dirichlet’s theorem on primes in arithmetic progressions, which is not proved in the book.
1. Pythagorean Triples

As an introduction to the sorts of questions that we will be studying, let us consider right triangles whose sides all have integer lengths. The most familiar example is the \((3, 4, 5)\) right triangle, but there are many others as well, such as the \((5, 12, 13)\) right triangle. Thus we are looking for triples \((a, b, c)\) of positive integers such that \[a^2 + b^2 = c^2.\] Such triples are called Pythagorean triples because of the connection with the Pythagorean Theorem. Our goal will be a formula that gives them all. The ancient Greeks knew such a formula, and even before the Greeks the ancient Babylonians must have known a lot about Pythagorean triples because one of their clay tablets from nearly 4000 years ago has been found which gives a list of 15 different Pythagorean triples, the largest of which is \((12709, 13500, 18541)\). (Actually the tablet only gives the numbers \(a\) and \(c\) from each triple \((a, b, c)\) for some unknown reason, but it is easy to compute \(b\) from \(a\) and \(c\).)

There is an easy way to create infinitely many Pythagorean triples from a given one just by multiplying each of its three numbers by an arbitrary number \(n\). For example, from \((3, 4, 5)\) we get \((6, 8, 10)\), \((9, 12, 15)\), \((12, 16, 20)\), and so on. This process produces right triangles that are all similar to each other, so in a sense they are not essentially different triples. In our search for Pythagorean triples there is thus no harm in restricting our attention to triples \((a, b, c)\) whose three numbers have no common factor. Such triples are called primitive. The large Babylonian triple mentioned above is primitive, since the prime factorization of 13500 is \(2^23^35^3\) but the other two numbers in the triple are not divisible by 2, 3, or 5.

A fact worth noting in passing is that if two of the three numbers in a Pythagorean triple \((a, b, c)\) have a common factor \(n\), then \(n\) is also a factor of the third number. This follows easily from the equation \[a^2 + b^2 = c^2,\] since for example if \(n\) divides \(a\) and \(b\) then \(n^2\) divides \(a^2\) and \(b^2\), so \(n^2\) divides their sum \(c^2\), hence \(n\) divides \(c\). Another case is that \(n\) divides \(a\) and \(c\). Then \(n^2\) divides \(a^2\) and \(c^2\) so \(n^2\) divides their difference \(c^2 - a^2 = b^2\), hence \(n\) divides \(b\). In the remaining case that \(n\) divides \(b\) and \(c\) the argument is similar.

A consequence of this divisibility fact is that primitive Pythagorean triples can also be characterized as the ones for which no two of the three numbers have a common factor.

If \((a, b, c)\) is a Pythagorean triple, then we can divide the equation \(a^2 + b^2 = c^2\) by \(c^2\) to get an equivalent equation \[(\frac{a}{c})^2 + (\frac{b}{c})^2 = 1.\] This equation is saying that the point \((x, y) = (\frac{a}{c}, \frac{b}{c})\) is on the unit circle \(x^2 + y^2 = 1\) in the \(xy\)-plane. The coordinates \(\frac{a}{c}\) and \(\frac{b}{c}\) are rational numbers, so each Pythagorean triple gives a rational point on the circle, i.e., a point whose coordinates are both rational. Notice that multiplying
each of $a$, $b$, and $c$ by the same integer $n$ yields the same point $(x, y)$ on the circle. Going in the other direction, given a rational point on the circle, we can find a common denominator for its two coordinates so that it has the form $(\frac{a}{c}, \frac{b}{c})$ and hence gives a Pythagorean triple $(a, b, c)$. We can assume this triple is primitive by canceling any common factor of $a$, $b$, and $c$, and this doesn’t change the point $(\frac{a}{c}, \frac{b}{c})$. The two fractions $\frac{a}{c}$ and $\frac{b}{c}$ must then be in lowest terms since we observed earlier that if two of $a$, $b$, $c$ have a common factor, then all three have a common factor.

From the preceding observations we can conclude that the problem of finding all Pythagorean triples is equivalent to finding all rational points on the unit circle $x^2 + y^2 = 1$. More specifically, there is an exact one-to-one correspondence between primitive Pythagorean triples and rational points on the unit circle that lie in the interior of the first quadrant (since we want all of $a, b, c, x, y$ to be positive).

In order to find all the rational points on the circle $x^2 + y^2 = 1$ we will use a construction that starts with one rational point and creates many more rational points from this one starting point. The four obvious rational points on the circle are the intersections of the circle with the coordinate axes, which are the points $(\pm 1, 0)$ and $(0, \pm 1)$. It doesn’t really matter which one we choose as the starting point, so let’s choose $(0, 1)$. Now consider a line which intersects the circle in this point $(0, 1)$ and some other point $P$, as in the figure at the right. If the line has slope $m$, its equation will be $y = mx + 1$. If we denote the point where the line intersects the $x$-axis by $(r, 0)$, then $m = -1/r$ so the equation for the line can be rewritten as $y = 1 - \frac{x}{r}$. To find the coordinates of the point $P$ in terms of $r$ we substitute $y = 1 - \frac{x}{r}$ into the equation $x^2 + y^2 = 1$ and solve for $x$:

$$
\begin{align*}
x^2 + \left(1 - \frac{x}{r}\right)^2 &= 1 \\
x^2 + 1 - \frac{2x}{r} + \frac{x^2}{r^2} &= 1 \\
\left(1 + \frac{1}{r^2}\right)x^2 - \frac{2x}{r} &= 0 \\
\left(\frac{r^2 + 1}{r^2}\right)x^2 &= \frac{2x}{r} \\
x &= \frac{2r}{r^2 + 1} \quad \text{or} \quad x = 0
\end{align*}
$$

Now we plug $x = \frac{2r}{r^2 + 1}$ into the formula $y = 1 - \frac{x}{r}$. This gives:

$$y = 1 - \frac{x}{r} = -\frac{1}{r}\left(\frac{2r}{r^2 + 1}\right) + 1 = -\frac{2}{r^2 + 1} + 1 = \frac{r^2 - 1}{r^2 + 1}$$
Summarizing, the coordinates \((x, y)\) of the point \(P\) are given by the following formula:

\[
(x, y) = \left(\frac{2r}{r^2 + 1}, \frac{r^2 - 1}{r^2 + 1}\right)
\]

Note that when \(x = 0\) there are two points \((0, \pm 1)\) on the circle. The point \((0, -1)\) comes from the value \(r = 0\), while if we let \(r\) approach \(\pm \infty\) then the point \(P\) approaches \((0, 1)\), as we can see either from the picture or from the formula for \((x, y)\).

If \(r\) is a rational number, then the formula for \((x, y)\) shows that both \(x\) and \(y\) are rational, so we have a rational point on the circle. Conversely, if both coordinates \(x\) and \(y\) of the point \(P\) on the circle are rational, then the slope \(m\) of the line must be rational, hence \(r\) must also be rational since \(r = -1/m\). We could also solve the equation \(y = 1 - \frac{x}{r}\) for \(r\) to get \(r = \frac{x}{1-x}\), showing again that \(r\) will be rational if \(x\) and \(y\) are rational (and \(y\) is not 1). The conclusion of all this is that, starting from the initial rational point \((0, 1)\) we have found formulas that give all the other rational points on the circle.

Since there are infinitely many choices for the rational number \(r\), there are infinitely many rational points on the circle. But we can say something much stronger than this: Every arc of the circle, no matter how small, contains infinitely many rational points. This is because every arc on the circle corresponds to an interval of \(r\)-values on the \(x\)-axis, and every interval in the \(x\)-axis contains infinitely many rational numbers. Since every arc on the circle contains infinitely many rational points, we can say that the rational points are dense in the circle, meaning that for every point on the circle there is an infinite sequence of rational points approaching the given point.

Now we can go back and find formulas for Pythagorean triples. If we set the rational number \(r\) equal to \(p/q\) with \(p\) and \(q\) integers having no common factor, then the formulas for \(x\) and \(y\) become:

\[
x = \frac{2\left(\frac{p}{q}\right)}{\frac{p^2}{q^2} + 1} = \frac{2pq}{p^2 + q^2}
\]

\[
y = \frac{\frac{p^2}{q^2} - 1}{\frac{p^2}{q^2} + 1} = \frac{p^2 - q^2}{p^2 + q^2}
\]

Our final formulas for Pythagorean triples are then:

\[
(a, b, c) = (2pq, p^2 - q^2, p^2 + q^2)
\]

Here are a few examples with small values of \(p\) and \(q\):
Suppose both odd, their sum and difference are both even and we can write they were both even they would have the common factor of 2. Because primitive if $p > q$ then this occurs only when $p$ and $q$ have no common divisor and $p > q$, such that $p$ and $q$ have no common factor and are of opposite parity (one even and the other odd).

<table>
<thead>
<tr>
<th>$(p, q)$</th>
<th>$(x, y)$</th>
<th>$(a, b, c)$</th>
</tr>
</thead>
<tbody>
<tr>
<td>(2, 1)</td>
<td>(4/5, 3/5)</td>
<td>(4, 3, 5)</td>
</tr>
<tr>
<td>(3, 1)*</td>
<td>(6/10, 8/10)*</td>
<td>(6, 8, 10)*</td>
</tr>
<tr>
<td>(3, 2)</td>
<td>(12/13, 5/13)</td>
<td>(12, 5, 13)</td>
</tr>
<tr>
<td>(4, 1)</td>
<td>(8/17, 15/17)</td>
<td>(8, 15, 17)</td>
</tr>
<tr>
<td>(4, 3)</td>
<td>(24/25, 7/25)</td>
<td>(24, 7, 25)</td>
</tr>
<tr>
<td>(5, 1)*</td>
<td>(10/26, 24/26)*</td>
<td>(10, 24, 26)*</td>
</tr>
<tr>
<td>(5, 2)</td>
<td>(20/29, 21/29)</td>
<td>(20, 21, 29)</td>
</tr>
<tr>
<td>(5, 3)*</td>
<td>(30/34, 16/34)*</td>
<td>(30, 16, 34)*</td>
</tr>
<tr>
<td>(5, 4)</td>
<td>(40/41, 9/41)</td>
<td>(40, 9, 41)</td>
</tr>
<tr>
<td>(6, 1)</td>
<td>(12/37, 35/37)</td>
<td>(12, 35, 37)</td>
</tr>
<tr>
<td>(6, 5)</td>
<td>(60/61, 11/61)</td>
<td>(60, 11, 61)</td>
</tr>
<tr>
<td>(7, 1)*</td>
<td>(14/50, 48/50)*</td>
<td>(14, 48, 50)*</td>
</tr>
<tr>
<td>(7, 2)</td>
<td>(28/53, 45/53)</td>
<td>(28, 45, 53)</td>
</tr>
<tr>
<td>(7, 3)*</td>
<td>(42/58, 40/58)*</td>
<td>(42, 40, 58)*</td>
</tr>
<tr>
<td>(7, 4)</td>
<td>(56/65, 33/65)</td>
<td>(56, 33, 65)</td>
</tr>
<tr>
<td>(7, 5)*</td>
<td>(70/74, 24/74)*</td>
<td>(70, 24, 74)*</td>
</tr>
<tr>
<td>(7, 6)</td>
<td>(84/85, 13/85)</td>
<td>(84, 13, 85)</td>
</tr>
</tbody>
</table>

The starred entries are the ones with nonprimitive Pythagorean triples. Notice that this occurs only when $p$ and $q$ are both odd, so that not only is $2pq$ even, but also both $p^2 - q^2$ and $p^2 + q^2$ are even, so all three of $a$, $b$, and $c$ are divisible by 2. The primitive versions of the nonprimitive entries in the table occur higher in the table, but with $a$ and $b$ switched. This is a general phenomenon, as we will see in the course of proving the following basic result:

**Proposition.** Up to interchanging $a$ and $b$, all primitive Pythagorean triples $(a, b, c)$ are obtained from the formula $(a, b, c) = (2pq, p^2 - q^2, p^2 + q^2)$ where $p$ and $q$ are positive integers, $p > q$, such that $p$ and $q$ have no common factor and are of opposite parity (one even and the other odd).

**Proof:** We need to investigate when the formula $(a, b, c) = (2pq, p^2 - q^2, p^2 + q^2)$ gives a primitive triple, assuming that $p$ and $q$ have no common divisor and $p > q$.

**Case 1:** Suppose $p$ and $q$ have opposite parity. If all three of $2pq$, $p^2 - q^2$, and $p^2 + q^2$ have a common divisor $d > 1$ then $d$ would have to be odd since $p^2 - q^2$ and $p^2 + q^2$ are odd when $p$ and $q$ have opposite parity. Furthermore, since $d$ is a divisor of both $p^2 - q^2$ and $p^2 + q^2$ it must divide their sum $(p^2 + q^2) + (p^2 - q^2) = 2p^2$ and also their difference $(p^2 + q^2) - (p^2 - q^2) = 2q^2$. However, since $d$ is odd it would then have to divide $p^2$ and $q^2$, forcing $p$ and $q$ to have a common factor (since any prime factor of $d$ would have to divide $p$ and $q$). This contradicts the assumption that $p$ and $q$ had no common factors, so we conclude that $(2pq, p^2 - q^2, p^2 + q^2)$ is primitive if $p$ and $q$ have opposite parity.

**Case 2:** Suppose $p$ and $q$ have the same parity, hence they are both odd since if they were both even they would have the common factor of 2. Because $p$ and $q$ are both odd, their sum and difference are both even and we can write $p + q = 2P$ and
$p - q = 2Q$ for some integers $P$ and $Q$. Any common factor of $P$ and $Q$ would have to divide $P + Q = \frac{p + q}{2} + \frac{p - q}{2} = p$ and $P - Q = \frac{p + q}{2} - \frac{p - q}{2} = q$, so $P$ and $Q$ have no common factors. In terms of $P$ and $Q$ our Pythagorean triple becomes

$$(a, b, c) = (2pq, p^2 - q^2, p^2 + q^2)$$

After canceling the factor of 2 we get a new Pythagorean triple, with the first two coordinates switched, and this one is primitive by Case 1 since $P$ and $Q$ can’t both be odd, because if they were, then $p = P + Q$ and $q = P - Q$ would both be even, which is impossible since they have no common factor.

From Cases 1 and 2 we can conclude that if we allow ourselves to switch the first two coordinates, then we get all primitive Pythagorean triples from the formula by restricting $p$ and $q$ to be of opposite parity and to have no common factors.

2. Rational Points on Quadratic Curves

The same technique we used to find the rational points on the circle $x^2 + y^2 = 1$ can also be used to find all the rational points on other quadratic curves $Ax^2 + Bxy + Cy^2 + Dx + Ey = F$ with integer or rational coefficients $A, B, C, D, E, F$, provided that we can find a single rational point $(x_0, y_0)$ on the curve to start the process. For example, the circle $x^2 + y^2 = 2$ contains the rational points $(\pm 1, \pm 1)$ and we can use one of these as an initial point. Taking the point $(1, 1)$, we would consider lines $y - 1 = m(x - 1)$ of slope $m$ passing through this point. Solving this equation for $y$ and plugging into the equation $x^2 + y^2 = 2$ would produce a quadratic equation $ax^2 + bx + c = 0$ whose coefficients are polynomials in the variable $m$, so these coefficients would be rational whenever $m$ is rational.

From the quadratic formula $x = \frac{-b \pm \sqrt{b^2 - 4ac}}{2a}$ we see that the sum of the two roots is $-b/a$, a rational number if $m$ is rational, so if one root is rational then the other root will be rational as well. The initial point $(1, 1)$ on the curve $x^2 + y^2 = 2$ gives $x = 1$ as one rational root of the equation $ax^2 + bx + c = 0$, so for each rational value of $m$ the other root $x$ will be rational as well. Then the equation $y - 1 = m(x - 1)$ implies that $y$ will also be rational, and hence we obtain a rational point $(x, y)$ on the curve for each rational value of $m$. Conversely, if $x$ and $y$ are both rational then obviously $m = (y - 1)/(x - 1)$ will be rational. Thus one obtains a dense set of rational points on the circle $x^2 + y^2 = 2$, since $m$ can be any rational number. An exercise at the end of this chapter is to work out the formulas explicitly.
If instead of \( x^2 + y^2 = 2 \) we consider the circle \( x^2 + y^2 = 3 \) then there aren’t any obvious rational points. In fact this circle contains no rational points at all. For if there were a rational point, this would yield a solution of the equation \( a^2 + b^2 = 3c^2 \) by integers \( a, b, \) and \( c \). We can assume \( a, b, \) and \( c \) have no common factor. Then \( a \) and \( b \) can’t both be even, otherwise the left side of the equation would be even, forcing \( c \) to be even, so \( a, b, \) and \( c \) would have a common factor of 2. To complete the argument we look at the equation modulo 4. (This means that we consider the remainders obtained after division by 4.) The square of an even number has the form \((2n)^2 = 4n^2\), which is 0 modulo 4, while the square of an odd number has the form \((2n + 1)^2 = 4n^2 + 4n + 1\), which is 1 modulo 4. Thus, modulo 4, the left side of the equation is either 0 + 1, 1 + 0, or 1 + 1 since \( a \) and \( b \) are not both even. So the left side is either 1 or 2 modulo 4. However, the right side is either 3 \cdot 0 or 3 \cdot 1 modulo 4. We conclude that there can be no integer solutions of \( a^2 + b^2 = 3c^2 \).

The technique we just used to show that \( a^2 + b^2 = 3c^2 \) has no integer solutions can be used in many other situations as well. The underlying reasoning is that if an equation with integer coefficients has an integer solution, then this gives a solution modulo \( n \) for all numbers \( n \). For solutions modulo \( n \) there are only a finite number of possibilities to check, although for large \( n \) this is a large finite number. If one can find a single value of \( n \) for which there is no solution modulo \( n \), then the original equation has no integer solutions. However, this implication is not reversible, as it is possible for an equation to have solutions modulo \( n \) for every number \( n \) and still have no actual integer solutions. A concrete example is the equation \( 2x^2 + 7y^2 = 1 \). This obviously has no integer solutions, yet it does have solutions modulo \( n \) for each \( n \), although this is certainly not obvious. Note that the ellipse \( 2x^2 + 7y^2 = 1 \) does contain rational points such as \((1/3, 1/3)\) and \((3/5, 1/5)\). These can in fact be used to show that \( 2x^2 + 7y^2 = 1 \) has solutions modulo \( n \) for each \( n \), as we will show in Section 6.3 of Chapter 6 when we study congruences in more detail.

In Chapter 6 we will also find a complete answer to the question of when the circle \( x^2 + y^2 = n \) contains rational points. It turns out that there are rational points on this circle only when there are integer points on it, and we will see that the existence of integer solutions of \( x^2 + y^2 = n \) depends heavily on the prime factorization of \( n \). Namely, we will show that \( x^2 + y^2 = n \) has integer solutions exactly when each prime factor of \( n \) of the form \( 4k + 3 \) occurs to an even power in the prime factorization of \( n \).
3. Rational Points on a Sphere

As another application of the same idea, we can find all the rational points on the sphere \( x^2 + y^2 + z^2 = 1 \), the triples \((x, y, z)\) of rational numbers that satisfy this equation. To do this we consider a line from the north pole \((0, 0, 1)\) to a point \((u, v, 0)\) in the \(xy\)-plane. This line intersects the sphere at some point \((x, y, z)\), and we want to find formulas expressing \(x\), \(y\), and \(z\) in terms of \(u\) and \(v\). To do this we use the following figure:

Suppose we look at the vertical plane containing the triangle \(ONQ\). From our earlier analysis of rational points on a circle of radius 1 we know that if the segment \(OQ\) has length \(|OQ| = r\), then \(|OP'| = \frac{2r}{r^2 + 1}\) and \(|PP'| = \frac{r^2 - 1}{r^2 + 1}\). From the right triangle \(OBQ\) we see that \(u^2 + v^2 = r^2\) since \(u = |OB|\) and \(v = |BQ|\). The triangle \(OBQ\) is similar to the triangle \(OAP'\). Since the length of \(OP'\) is \(\frac{2}{r^2 + 1}\) times the length of \(OQ\) we conclude from similar triangles that

\[
x = |OA| = \frac{2}{r^2 + 1}|OB| = \frac{2}{r^2 + 1} \cdot u = \frac{2u}{u^2 + v^2 + 1}
\]

and

\[
y = |AP'| = \frac{2}{r^2 + 1}|BQ| = \frac{2}{r^2 + 1} \cdot v = \frac{2v}{u^2 + v^2 + 1}
\]

Also we have

\[
z = |PP'| = \frac{r^2 - 1}{r^2 + 1} = \frac{u^2 + v^2 - 1}{u^2 + v^2 + 1}
\]
Summarizing, we have expressed $x$, $y$, and $z$ in terms of $u$ and $v$ by the formulas

$$x = \frac{2u}{u^2 + v^2 + 1}, \quad y = \frac{2v}{u^2 + v^2 + 1}, \quad z = \frac{u^2 + v^2 - 1}{u^2 + v^2 + 1}.$$

These formulas imply that we get a rational point $(x, y, z)$ on the sphere $x^2 + y^2 + z^2 = 1$ for each pair of rational numbers $(u, v)$. We get all rational points on the sphere in this way (except for the north pole $(0, 0, 1)$, of course) since it is possible to express $u$ and $v$ in terms of $x$, $y$, and $z$ by the formulas

$$u = \frac{x}{1-z}, \quad v = \frac{y}{1-z},$$

which one can easily verify by substituting into the previous formulas.

Here is a short table giving a few rational points on the sphere and the corresponding integer solutions of the equation $a^2 + b^2 + c^2 = d^2$:

<table>
<thead>
<tr>
<th>$(u, v)$</th>
<th>$(x, y, z)$</th>
<th>$(a, b, c, d)$</th>
</tr>
</thead>
<tbody>
<tr>
<td>$(1, 1)$</td>
<td>$(2/3, 2/3, 1/3)$</td>
<td>$(2, 2, 1, 3)$</td>
</tr>
<tr>
<td>$(2, 2)$</td>
<td>$(4/9, 4/9, 7/9)$</td>
<td>$(4, 4, 7, 9)$</td>
</tr>
<tr>
<td>$(1, 3)$</td>
<td>$(2/11, 6/11, 9/11)$</td>
<td>$(2, 6, 9, 11)$</td>
</tr>
<tr>
<td>$(2, 3)$</td>
<td>$(2/7, 3/7, 6/7)$</td>
<td>$(2, 3, 6, 7)$</td>
</tr>
<tr>
<td>$(1, 4)$</td>
<td>$(1/9, 4/9, 8/9)$</td>
<td>$(1, 4, 8, 9)$</td>
</tr>
</tbody>
</table>

As with rational points on the circle $x^2 + y^2 = 1$, rational points on the sphere $x^2 + y^2 + z^2 = 1$ are dense, so there are lots of them scattered all over the sphere.

In linear algebra courses one is often called upon to create unit vectors $(x, y, z)$ by taking a given vector and rescaling it to have length 1 by dividing it by its length. For example, the vector $(1, 1, 1)$ has length $\sqrt{3}$ so the corresponding unit vector is $(1/\sqrt{3}, 1/\sqrt{3}, 1/\sqrt{3})$. It is rare that this process produces unit vectors having rational coordinates, but we now have a method for creating as many rational unit vectors as we like.

Incidentally, there is a name for the correspondence we have described between points $(x, y, z)$ on the unit sphere and points $(u, v)$ in the plane: it is called **stereographic projection**. One can think of the sphere and the plane as being made of clear glass, and one puts one’s eye at the north pole of the sphere and looks downward and outward in all directions to see points on the sphere projected onto points in the plane, and vice versa. The north pole itself does not project onto any point in the plane, but points approaching the north pole project to points approaching infinity in the plane, so one can think of the north pole as corresponding to an imaginary infinitely distant “point” in the plane. This geometric viewpoint somehow makes infinity less of a mystery, as it just corresponds to a point on the sphere, and points on a sphere are not very mysterious. (Though in the early days of polar exploration the north pole may have seemed very mysterious and infinitely distant!)
4. Pythagorean Triples and Quadratic Forms

There are many questions one can ask about Pythagorean triples \((a,b,c)\). For example, we could begin by asking which numbers actually arise as the numbers \(a\), \(b\), or \(c\) in some Pythagorean triple. It is sufficient to answer the question just for primitive Pythagorean triples, since the remaining ones are obtained just by multiplying by arbitrary positive integers. We know all primitive Pythagorean triples arise from the formula

\[(a,b,c) = (2pq, p^2 - q^2, p^2 + q^2)\]

where \(p\) and \(q\) have no common factor and are not both odd. Determining whether a given number can be expressed in the form \(2pq\), \(p^2 - q^2\), or \(p^2 + q^2\) is a special case of the general question of deciding when an equation \(Ap^2 + Bpq + Cq^2 = n\) has an integer solution \(p, q\), for given integers \(A, B, C,\) and \(n\). Expressions of the form \(Ax^2 + Bxy + Cy^2\) are called quadratic forms. These will be the main topic studied in Chapters 4–8, where we will develop some general theory addressing the question of what values a quadratic form takes on when all the numbers involved are integers. For now, let us just look at the special cases at hand.

First let us consider which numbers occur as \(a\) or \(b\) in primitive Pythagorean triples \((a,b,c)\). A trivial case is the equation \(0^2 + 1^2 = 1^2\) which shows that 0 and 1 can be realized by the triple \((0,1,1)\) which is primitive, so let us focus on realizing numbers bigger than 1. If we look at the earlier table of Pythagorean triples we see that all the numbers up to 15 can be realized as \(a\) or \(b\) in primitive triples except for 2, 6, 10, and 14. This might lead us to guess that the numbers realizable as \(a\) or \(b\) in primitive Pythagorean triples are the numbers not of the form \(4k + 2\), or in other words, numbers not congruent to 2 modulo 4. This is indeed true, and can be proved as follows. First note that since \(2pq\) is even, \(p^2 - q^2\) must be odd, otherwise both \(a\) and \(b\) would be even, violating primitivity. Now, every odd number is expressible in the form \(p^2 - q^2\) since \(2k + 1 = (k + 1)^2 - k^2\), so in fact every odd number is the difference between two consecutive squares. Taking \(p = k + 1\) and \(q = k\) yields a primitive triple since \(k\) and \(k + 1\) always have opposite parity and no common factors. This takes care of realizing odd numbers. For even numbers, they would have to be expressible as \(2pq\) with \(p\) and \(q\) of opposite parity, which forces \(pq\) to be even so \(2pq\) is a multiple of 4 and hence cannot be of the form \(4k + 2\). On the other hand, if we take \(p = 2k\) and \(q = 1\) then \(2pq = 4k\) with \(p\) and \(q\) having opposite parity and no common factors.

To summarize, we have shown that all positive numbers \(2k + 1\) and \(4k\) occur as \(a\) or \(b\) in primitive Pythagorean triples but none of the numbers \(4k + 2\) occur. To finish the story, note that a number \(a = 4k + 2\) which can’t be realized in a primitive triple can be realized by a nonprimitive triple just by taking a triple \((a,b,c)\) with \(a = 2k + 1\) and doubling each of \(a, b,\) and \(c\). Thus all numbers can be realized as \(a\) or \(b\) in Pythagorean triples \((a,b,c)\).
Now let us ask which numbers \( c \) can occur in Pythagorean triples \((a, b, c)\), so we are trying to find a solution of \( p^2 + q^2 = c \) for a given number \( c \). Pythagorean triples \((p, q, r)\) give solutions when \( c \) is equal to a square \( r^2 \), but we are asking now about arbitrary numbers \( c \). It suffices to figure out which numbers \( c \) occur in primitive triples \((a, b, c)\), since by multiplying the numbers \( c \) in primitive triples by arbitrary numbers we get the numbers \( c \) in arbitrary triples. A look at the earlier table shows that the numbers \( c \) that can be realized by primitive triples \((a, b, c)\) seem to be fairly rare: only 5, 13, 17, 25, 29, 37, 41, 53, 61, 65, and 85 occur in the table. These are all odd, and in fact they are all congruent to 1 modulo 4. Thus we get the infinite sequence of solutions \( p \) with corresponding triples \((p, q)\). The left side doesn't factor using integer coefficients, so it's not so easy to find integer solutions this time. In the table there are only the two triples \((4, 3, 5)\) and \((20, 21, 29)\), with \((p, q) = (2, 1)\) and \((5, 2)\). After some trial and error one...
could find the next solution \((p, q) = (12, 5)\) which gives the triple \((120, 119, 169)\). Is there a pattern in the solutions \((2, 1), (5, 2), (12, 5)\)? One has the numbers \(1, 2, 5, 12\), and perhaps it isn’t too much of a stretch to notice that the third number is twice the second plus the first, while the fourth number is twice the third plus the second. If this pattern continued, the next number would be \(29 = 2 \cdot 12 + 5\), giving \((p, q) = (29, 12)\), and this does indeed satisfy \(p^2 - 2pq - q^2 = 1\), yielding the Pythagorean triple \((696, 697, 985)\). These numbers are increasing rather rapidly, and the next case \((p, q) = (70, 29)\) yields an even bigger Pythagorean triple \((4060, 4059, 5741)\). Could there be other solutions of \(p^2 - 2pq - q^2 = \pm 1\) with smaller numbers that we missed? We will develop tools in Chapters 4 and 5 to find all the integer solutions, and it will turn out that the sequence we have just discovered gives them all.

Although the quadratic form \(p^2 - 2pq - q^2\) does not factor using integer coefficients, it can be simplified slightly by rewriting it as \((p - q)^2 - 2q^2\). Then if we change variables by setting

\[
\begin{align*}
  x &= p - q \\
  y &= q
\end{align*}
\]

we obtain the quadratic form \(x^2 - 2y^2\). Finding integer solutions of \(x^2 - 2y^2 = n\) is equivalent to finding integer solutions of \(p^2 - 2pq - q^2 = n\) since integer values of \(p\) and \(q\) give integer values of \(x\) and \(y\), and conversely, integer values of \(x\) and \(y\) give integer values of \(p\) and \(q\) since when we solve for \(p\) and \(q\) in terms of \(x\) and \(y\) we again get equations with integer coefficients:

\[
\begin{align*}
  p &= x + y \\
  q &= y
\end{align*}
\]

Thus the quadratic forms \(p^2 - 2pq - q^2\) and \(x^2 - 2y^2\) are completely equivalent, and finding integer solutions of \(p^2 - 2pq - q^2 = \pm 1\) is equivalent to finding integer solutions of \(x^2 - 2y^2 = \pm 1\).

The equation \(x^2 - 2y^2 = \pm 1\) is an instance of the equation \(x^2 - Dy^2 = \pm 1\) which is known as Pell’s equation (although sometimes this term is used only when the right hand side of the equation is \(+1\) and the other case is called the negative Pell equation). This is a very famous equation in number theory which has arisen in many different contexts going back hundreds of years. We will develop techniques for finding all integer solutions of Pell’s equation for arbitrary values of \(D\) in Chapters 4 and 5. It is interesting that certain fairly small values of \(D\) can force the solutions to be quite large. For example for \(D = 61\) the smallest positive integer solution of \(x^2 - 61y^2 = 1\) is the rather large pair

\((x, y) = (1766319049, 226153980)\)

As far back as the eleventh and twelfth centuries mathematicians in India knew how to find this solution. It was rediscovered in the seventeenth century by Fermat in France,
who also gave the smallest solution of \( x^2 - 109y^2 = 1 \), the even larger pair

\[
(x, y) = (158070671986249, 15140424455100)
\]

The way that the size of the smallest solution of \( x^2 - Dy^2 = 1 \) depends upon \( D \) is very erratic and is still not well understood today.

5. Pythagorean Triples and Complex Numbers

There is another way of looking at Pythagorean triples that involves complex numbers, surprisingly enough. The starting point here is the observation that \( a^2 + b^2 \) can be factored as \((a + bi)(a - bi)\) where \( i = \sqrt{-1} \). If we rewrite the equation \( a^2 + b^2 = c^2 \) as \((a + bi)(a - bi) = c^2\) then since the right side of the equation is a square, we might wonder whether each term on the left side would have to be a square too. For example, in the case of the triple \((3, 4, 5)\) we have \((3 + 4i)(3 - 4i) = 5^2\) with \((3 + 4i) = (2 + i)^2\) and \((3 - 4i) = (2 - i)^2\). So let us ask optimistically whether the equation \((a + bi)(a - bi) = c^2\) can be rewritten as \((p + qi)^2(p - qi)^2 = c^2\) with \(a + bi = (p + qi)^2\) and \(a - bi = (p - qi)^2\). We might hope also that the equation \((p + qi)(p - qi)^2 = c^2\) was obtained by simply squaring the equation \((p + qi)(p - qi) = c\). Let us see what happens when we multiply these various products out:

\[
a + bi = (p + qi)^2 = (p^2 - q^2) + (2pq)i \\
\text{hence } a = p^2 - q^2 \text{ and } b = 2pq \\
a - bi = (p - qi)^2 = (p^2 - q^2) - (2pq)i \\
\text{hence again } a = p^2 - q^2 \text{ and } b = 2pq \\
c = (p + qi)(p - qi) = p^2 + q^2
\]

Thus we have miraculously recovered the formulas for Pythagorean triples that we obtained earlier by geometric means (with \( a \) and \( b \) switched, which doesn’t really matter):

\[
a = p^2 - q^2 \quad b = 2pq \quad c = p^2 + q^2
\]

Of course, our derivation of these formulas just now depended on several assumptions that we haven’t justified, but it does suggest that looking at complex numbers of the form \( a + bi \) where \( a \) and \( b \) are integers might be a good idea. There is a name for complex numbers of this form \( a + bi \) with \( a \) and \( b \) integers. They are called Gaussian integers, since the great mathematician and physicist C.F. Gauss made a thorough algebraic study of them some 200 years ago. We will develop the basic properties of Gaussian integers in Chapter 8, in particular explaining why the derivation of the formulas above is valid.

6. Diophantine Equations

Equations like \( x^2 + y^2 = z^2 \) or \( x^2 - Dy^2 = 1 \) that involve polynomials with integer coefficients, and where the solutions sought are required to be integers, are called
Diophantine equations after the Greek mathematician Diophantus (ca. 250 A.D.) who wrote a book about these equations that was very influential when European mathematicians started to consider this topic much later in the 1600s. Usually Diophantine equations are very hard to solve because of the restriction to integer solutions. The first really interesting case is quadratic Diophantine equations. By the year 1800 there was quite a lot known about the quadratic case, and we will be focusing on this case in this book.

Diophantine equations of higher degree than quadratic are much more challenging to understand. Probably the most famous one is $x^n + y^n = z^n$ where $n$ is a fixed integer greater than 2. When the French mathematician Fermat in the 1600s was reading about Pythagorean triples in his copy of Diophantus’ book he made a marginal note that, in contrast with the equation $x^2 + y^2 = z^2$, the equation $x^n + y^n = z^n$ has no solutions with positive integers $x, y, z$ when $n > 2$ and that he had a marvelous proof which unfortunately the margin was too narrow to contain. This is one of many statements that he claimed were true but never wrote proofs of for public distribution, nor have proofs been found among his manuscripts. Over the next century other mathematicians discovered proofs for all his other statements, but this one was far more difficult to verify. The issue is clouded by the fact that he only wrote this statement down the one time, whereas all his other important results were stated numerous times in his correspondence with other mathematicians of the time. So perhaps he only briefly believed he had a proof. In any case, the statement has become known as Fermat’s Last Theorem. It was finally proved in the 1990s by Andrew Wiles, using some very deep mathematics developed over the preceding couple decades.

We have seen that finding integer solutions of $x^2 + y^2 = z^2$ is equivalent to finding rational points on the circle $x^2 + y^2 = 1$, and in the same way finding integer solutions of $x^n + y^n = z^n$ is equivalent to finding rational points on the curve $x^n + y^n = 1$. For even values of $n > 2$ this curve looks like a flattened out circle while for odd $n$ it has a rather different shape, extending out to infinity in the second and fourth quadrants, asymptotic to the line $y = -x$:

Fermat’s Last Theorem is equivalent to the statement that these curves have no rational points except their intersections with the coordinate axes, where either $x$ or
\( y \) is 0. It is curious that these curves contain only a finite number of rational points (either two points or four points, depending on whether \( n \) is odd or even) whereas quadratic curves like \( x^2 + y^2 = n \) contain either no rational points or an infinite dense set of rational points.

**Exercises**

1. (a) Make a list of the 16 primitive Pythagorean triples \((a, b, c)\) with \( c \leq 100 \), regarding \((a, b, c)\) and \((b, a, c)\) as the same triple.
   (b) How many more would there be if we allowed nonprimitive triples?
   (c) How many triples (primitive or not) are there with \( c = 65 \)?

2. (a) Find all the positive integer solutions of \( x^2 - y^2 = 512 \) by factoring \( x^2 - y^2 \) as \((x + y)(x - y)\) and considering the possible factorizations of 512.
   (b) Show that the equation \( x^2 - y^2 = n \) has only a finite number of integer solutions for each value of \( n > 0 \).
   (c) Find a value of \( n > 0 \) for which the equation \( x^2 - y^2 = n \) has at least 100 different positive integer solutions.

3. (a) Show that there are only a finite number of Pythagorean triples \((a, b, c)\) with \( a \) equal to a given number \( n \).
   (b) Show that there are only a finite number of Pythagorean triples \((a, b, c)\) with \( c \) equal to a given number \( n \).

4. Find an infinite sequence of primitive Pythagorean triples where two of the numbers in each triple differ by 2.

5. Find a right triangle whose sides have integer lengths and whose acute angles are close to 30 and 60 degrees by first finding the irrational value of \( r \) that corresponds to a right triangle with acute angles exactly 30 and 60 degrees, then choosing a rational number close to this irrational value of \( r \).

6. Find a right triangle whose sides have integer lengths and where one of the nonhypotenuse sides is approximately twice as long as the other, using a method like the one in the preceding problem. (One possible answer might be the \((8, 15, 17)\) triangle, or a triangle similar to this, but you should do better than this.)

7. Find a rational point on the sphere \( x^2 + y^2 + z^2 = 1 \) whose \( x \), \( y \), and \( z \) coordinates are nearly equal.

8. (a) Derive formulas that give all the rational points on the circle \( x^2 + y^2 = 2 \) in terms of a rational parameter \( m \), the slope of the line through the point \((1,1)\) on the circle. (The value \( m = \infty \) should be allowed as well, yielding the point \((1,-1)\).) The calculations may be a little messy, but they work out fairly nicely in the end to give

\[
x = \frac{m^2 - 2m - 1}{m^2 + 1}, \quad y = \frac{-m^2 - 2m + 1}{m^2 + 1}
\]
(b) Using these formulas, find five different rational points on the circle in the first quadrant, and hence five solutions of $a^2 + b^2 = 2c^2$ with positive integers $a, b, c$.

(c) The equation $a^2 + b^2 = 2c^2$ can be rewritten as $c^2 = (a^2 + b^2)/2$, which says that $c^2$ is the average of $a^2$ and $b^2$, or in other words, the squares $a^2, c^2, b^2$ form an arithmetic progression. One can assume $a < b$ by switching $a$ and $b$ if necessary. Find four such arithmetic progressions of three increasing squares where in each case the three numbers have no common divisors.

9. (a) Find formulas that give all the rational points on the upper branch of the hyperbola $y^2 - x^2 = 1$.

(b) Can you find any relationship between these rational points and Pythagorean triples?

10. (a) For integers $x$, what are the possible values of $x^2$ modulo 8?

(b) Show that the equation $x^2 - 2y^2 = \pm 3$ has no integer solutions by considering this equation modulo 8.

(c) Show that there are no primitive Pythagorean triples $(a, b, c)$ with $a$ and $b$ differing by 3.

11. Show that for every Pythagorean triple $(a, b, c)$ the product $abc$ must be divisible by 60. (It suffices to show that $abc$ is divisible by 3, 4, and 5.)
Chapter 1. The Farey Diagram

Our goal is to use geometry to study numbers. Of the various kinds of numbers, the simplest are integers, along with their ratios, the rational numbers. Usually one thinks of rational numbers geometrically as points along a line interspersed with irrational numbers as well. In this chapter we introduce a two-dimensional pictorial representation of rational numbers that displays certain interesting relations between them that we will be exploring. This diagram, along with several variants of it that will be introduced later, is known as the Farey diagram. The origin of the name will be explained when we get to one of these variants.

1. The Diagram.

Here is what the Farey diagram looks like:

What is shown here is not the whole diagram but only a finite part of it. The actual diagram has infinitely many curvilinear triangles, getting smaller and smaller out near the boundary circle. The diagram can be constructed by first inscribing the two big triangles in the circle, then adding the four triangles that share an edge with the two big triangles, then the eight triangles sharing an edge with these four, then sixteen
more triangles, and so on forever. With a little practice one can draw the diagram without lifting one’s pencil from the paper: First draw the outer circle starting at the left or right side, then the diameter, then make the two large triangles, then the four next-largest triangles, etc.

The vertices of all the triangles are labeled with fractions $a/b$, including the fraction $1/0$ for $\infty$, according to the following scheme. In the upper half of the diagram first label the vertices of the big triangles $0/1$, $1/1$, and $1/0$ as shown. Then by induction, if the labels at the two ends of the long edge of a triangle are $a/b$ and $c/d$, the label on the third vertex of the triangle is $\frac{a+c}{b+d}$. This fraction is called the mediant of $a/b$ and $c/d$.

The labels in the lower half of the diagram follow the same scheme, starting with the labels $0/1$, $-1/1$, and $-1/0$ on the large triangle. Using $-1/0$ instead of $1/0$ as the label of the vertex at the far left means that we are regarding $+\infty$ and $-\infty$ as the same. The labels in the lower half of the diagram are the negatives of those in the upper half, and the labels in the left half are the reciprocals of those in the right half.

The labels occur in their proper order around the circle, increasing from $-\infty$ to $+\infty$ as one goes around the circle in the counterclockwise direction. To see why this is so, it suffices to look at the upper half of the diagram where all numbers are positive. What we want to show is that the mediant $\frac{a+c}{b+d}$ is always a number between $\frac{a}{b}$ and $\frac{c}{d}$ (hence the term “mediant”). Thus we want to see that if $\frac{a}{b} > \frac{c}{d}$ then $\frac{a}{b} > \frac{a+c}{b+d} > \frac{c}{d}$. Since we are dealing with positive numbers, the inequality $\frac{a}{b} > \frac{c}{d}$ is equivalent to $ad > bc$, and $\frac{a}{b} > \frac{a+c}{b+d}$ is equivalent to $ab + ad > ab + bc$ which follows from $ad > bc$. Similarly, $\frac{a+c}{b+d} > \frac{c}{d}$ is equivalent to $ad + cd > bc + cd$ which also follows from $ad > bc$.

We will show in the next chapter that the mediant rule for labeling vertices in the diagram automatically produces labels that are fractions in lowest terms. It is not immediately apparent why this should be so. For example, the mediant of $1/3$ and $2/3$ is $3/6$, which is not in lowest terms, and the mediant of $2/7$ and $3/8$ is $5/15$, again not in lowest terms. Somehow cases like this don’t occur in the diagram.

Another non-obvious fact about the diagram is that all rational numbers occur eventually as labels of vertices. This will be shown in the next chapter as well.

The construction we have described for the Farey diagram involves an inductive process, where more and more triangles are added in succession. With a construction like this it is not easy to tell by a simple calculation whether or not two given rational numbers $a/b$ and $c/d$ are joined by an edge in the diagram. Fortunately there is such a criterion:

Two rational numbers $a/b$ and $c/d$ are joined by an edge in the Farey diagram exactly when the determinant $ad - bc$ of the matrix \( \begin{pmatrix} a & c \\ b & d \end{pmatrix} \) is $\pm 1$. This applies also
when one of \( a/b \) or \( c/d \) is \( \pm 1/0 \).

We will prove this in the next chapter. What it means is that if one starts with the rational numbers together with \( 1/0 = -1/0 \) arranged in order around a circle and one inserts circular arcs inside this circle meeting it perpendicularly and joining each pair \( a/b \) and \( c/d \) such that \( ad - bc = \pm 1 \), with a diameter of the circle as the arc joining \( 1/0 \) and \( 0/1 \), then no two of these arcs would cross, and they would divide the interior of the circle into non-overlapping curvilinear triangles. This is really quite remarkable when you think about it, and it does not happen for other values of the determinant besides \( \pm 1 \). For example, for determinant \( \pm 2 \) the edges would be the dotted arcs in the figure at the right. Here there are three arcs crossing in each triangle of the original Farey diagram, and these arcs divide each triangle of the Farey diagram into six smaller triangles.

2. Farey Series

We can build the set of rational numbers by starting with the integers and then inserting in succession all the halves, thirds, fourths, fifths, sixths, and so on. Let us look at what happens if we restrict to rational numbers between 0 and 1. Starting with 0 and 1 we first insert 1/2, then 1/3 and 2/3, then 1/4 and 3/4, skipping 2/4 which we already have, then inserting 1/5, 2/5, 3/5, and 4/5, then 1/6 and 5/6, etc. This process can be pictured as in the following diagram:
The interesting thing to notice is:

*Each time a new number is inserted, it forms the third vertex of a triangle whose other two vertices are its two nearest neighbors among the numbers already listed, and if these two neighbors are $a/b$ and $c/d$ then the new vertex is exactly the mediant $\frac{a+c}{b+d}$.*

The discovery of this curious phenomenon in the early 1800s was initially attributed to a geologist and amateur mathematician named Farey, although it turned out that he was not the first person to have noticed it. In spite of this confusion, the sequence of fractions $a/b$ between 0 and 1 with denominator less than or equal to a given number $n$ is usually called the $n$th Farey series $F_n$. For example, here is $F_7$:

$$0 \quad 1 \quad 1 \quad 1 \quad 1 \quad 2 \quad 1 \quad 2 \quad 3 \quad 1 \quad 4 \quad 3 \quad 2 \quad 5 \quad 3 \quad 4 \quad 5 \quad 6 \quad 1$$

These numbers trace out the up-and-down path across the bottom of the figure above. For the next Farey series $F_8$ we would insert $1/8$ between $0/1$ and $1/7$, $3/8$ between $1/3$ and $2/5$, $5/8$ between $3/5$ and $2/3$, and finally $7/8$ between $6/7$ and $1/1$.

There is a cleaner way to draw the preceding diagram using straight lines in a square:

![Diagram](image_url)

One can construct this diagram in stages, as indicated in the sequence of figures below. Start with a square together with its diagonals and a vertical line from their intersection point down to the bottom edge of the square. Next, connect the resulting midpoint of the lower edge of the square to the two upper corners of the square and drop vertical lines down from the two new intersection points this produces. Now add a W-shaped zigzag and drop verticals again. It should then be clear how to continue.
A nice feature of this construction is that if we start with a square whose sides have length 1 and place this square so that its bottom edge lies along the x-axis with the lower left corner of the square at the origin, then the construction assigns labels to the vertices along the bottom edge of the square that are exactly the x-coordinates of these points. Thus the vertex labeled 1/2 really is at the midpoint of the bottom edge of the square, and the vertices labeled 1/3 and 2/3 really are 1/3 and 2/3 of the way along this edge, and so forth. In order to verify this fact the key observation is the following: For a vertical line segment in the diagram whose lower endpoint is at the point \((a/b, 0)\) on the x-axis, the upper endpoint is at the point \((a/b, 1/b)\). This is obviously true at the first stage of the construction, and it continues to hold at each successive stage since for a quadrilateral whose four vertices have coordinates as shown in the figure at the right, the two diagonals intersect at the point \((a+b/c, 1/d)\). For example, to verify that \((a+b/c, 1/d)\) is on the line from \((a/b, 0)\) to \((c/d, 1/d)\) it suffices to show that the line segments from \((a/b, 0)\) to \((a+b/c, 1/d)\) and from \((a+b/c, 1/d)\) to \((c/d, 1/d)\) have the same slope. These slopes are

\[
\frac{1/(b+d) - 0}{(a+c)/(b+d) - a/b} \cdot \frac{b(b+d)}{b(b+d)} = \frac{b}{b(a+c) - a(b+d)} = \frac{b}{bc - ad}
\]

and

\[
\frac{1/d - 1/(b+d)}{c/d - (a+c)/(b+d)} \cdot \frac{d(b+d)}{d(b+d)} = \frac{b+d - d}{c(b+d) - d(a+c)} = \frac{b}{bc - ad}
\]

so they are equal. The same argument works for the other diagonal, just by interchanging \(a/b\) and \(c/d\).

Going back to the square diagram, this fact that we have just shown implies that the successive Farey series can be obtained by taking the vertices that lie above the line \(y = 1/2\), then the vertices above \(y = 1/3\), then above \(y = 1/4\), and so on. Here we are assuming the two properties of the Farey diagram that will be shown in the next chapter, that all rational numbers occur eventually as labels on vertices, and that these labels are always fractions in lowest terms.
3. The Upper Half-Plane Farey Diagram

In the square diagram depicting the Farey series, the most important thing for our purposes is the triangles, not the vertical lines. We can get rid of all the vertical lines by shrinking each one to its lower endpoint, converting each triangle into a curvilinear triangle with semicircles as edges, as shown in the diagram below.

This looks more like a portion of the Farey diagram we started with at the beginning of the chapter, but with the outer boundary circle straightened into a line. The advantage of the new version is that the labels on the vertices are exactly in their correct places along the $x$-axis, so the vertex labeled $\frac{a}{b}$ is exactly at the point $\frac{a}{b}$ on the $x$-axis.

This diagram can be enlarged so as to include similar diagrams for fractions between all pairs of adjacent integers, not just 0 and 1, all along the $x$-axis:

We can also put in vertical lines at the integer points, extending upward to infinity. These correspond to the edges having one endpoint at the vertex $1/0$ in the original Farey diagram.

We could also form a linear version of the full Farey diagram from copies of the square:
4. Relation with Pythagorean Triples

Next we describe a variant of the circular Farey diagram that is closely related to Pythagorean triples. Recall from Chapter 0 that rational points \((x, y)\) on the unit circle correspond to rational points \(p/q\) on the \(x\)-axis by means of lines through the point \((0, 1)\) on the circle. In formulas, \((x, y) = \left(\frac{2pq}{p^2+q^2}, \frac{p^2-q^2}{p^2+q^2}\right)\). Using this correspondence, we can label the rational points on the circle by the corresponding rational points on the \(x\)-axis and then construct a new Farey diagram in the circle by filling in triangles by the mediant rule just as before.

The result is a version of the circular Farey diagram that is rotated by 90 degrees to put \(1/0\) at the top of the circle, and there are also some perturbations of the positions of the other vertices and the shapes of the triangles. The next figure shows an enlargement of the new part of the diagram, with the vertices labeled by both the fraction \(p/q\) and the coordinates \((x, y)\) of the vertex:
Exercises

1. This problem involves another version of the Farey diagram, or at least the positive part of the diagram, the part consisting of the triangles whose vertices are labeled by fractions $p/q$ with $p \geq 0$ and $q \geq 0$. In this variant of the diagram the vertex labeled $p/q$ is placed at the point $(q, p)$ in the plane. Thus $p/q$ is the slope of the line through the origin and $(q, p)$. The edges of this new Farey diagram are straight line segments connecting the pairs of vertices that are connected in the original Farey diagram. For example there is a triangle with vertices $(1, 0)$, $(0, 1)$, and $(1, 1)$ corresponding to the big triangle in the upper half of the circular Farey diagram. What you are asked to do in this problem is just to draw the portion of the new Farey diagram consisting of all the triangles whose vertices $(q, p)$ satisfy $0 \leq q \leq 5$ and $0 \leq p \leq 5$. Note that since fractions $p/q$ labeling vertices are always in lowest terms, the points $(q, p)$ such that $q$ and $p$ have a common divisor greater than 1 are not vertices of the diagram.
A parenthetical comment: With this model of the Farey diagram the operation of forming the mediant of two fractions just corresponds to standard vector addition 
\[(a, b) + (c, d) = (a + c, b + d),\] which may make the mediant operation seem more natural.

2. Compute the Farey series \(F_{10}\). 
Here are two typical examples of continued fractions:

\[
\frac{7}{16} = 2 + \frac{1}{3 + \frac{1}{2}} \quad \quad \quad \quad \quad \quad \frac{67}{24} = 2 + \frac{1}{3 + \frac{1}{1 + \frac{1}{4}}}
\]

To compute the value of a continued fraction one starts in the lower right corner and works one’s way upward. For example in the continued fraction for \( \frac{7}{16} \), one starts with \( 3 + \frac{1}{2} = \frac{7}{2} \), then taking 1 over this gives \( \frac{2}{7} \), and adding the 2 to this gives \( \frac{16}{7} \), and finally 1 over this gives \( \frac{7}{16} \).

Here is the general form of a continued fraction:

\[
\frac{p}{q} = a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}}
\]

To write this in more compact form on a single line one can write it as

\[
\frac{p}{q} = a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}}
\]

For example:

\[
\frac{7}{16} = 1/2 + 1/3 + 1/2 \quad \quad \quad \quad \quad \frac{67}{24} = 2 + 1/1 + 1/3 + 1/1 + 1/4
\]

This way of writing continued fractions with upward-pointing diagonal arrows is intended to be a more legible version of the classical notation

\[
a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}}
\]

often found in older books. A more abbreviated notation common in other books is simply \([a_0; a_1, a_2, \ldots, a_n]\), although we will not use this notation here.

To compute the continued fraction for a given rational number one starts in the upper left corner and works one’s way downward, as the following example shows:

\[
\frac{67}{24} = 2 + \frac{19}{24} = 2 + \frac{1}{24/19} = 2 + \frac{1}{1 + 5/19} = 2 + \frac{1}{1 + \frac{1}{19/5}}
\]

\[
= 2 + \frac{1}{1 + \frac{1}{3 + 4/5}} = 2 + \frac{1}{1 + \frac{1}{3 + \frac{1}{5/4}}} = 2 + \frac{1}{1 + \frac{1}{3 + \frac{1}{1 + \frac{1}{4}}}}
\]

If one is good at mental arithmetic and the numbers aren’t too large, only the final form of the answer needs to be written down: \( \frac{67}{24} = 2 + \frac{1}{1} + \frac{1}{3} + \frac{1}{1} + \frac{1}{4} \).
1. The Euclidean Algorithm

The process for computing the continued fraction for a given rational number is known as the *Euclidean Algorithm*. It consists of repeated division, at each stage dividing the previous remainder into the previous divisor. The procedure for 67/24 is shown at the right. Note that the numbers in the shaded box are the numbers \(a_i\) in the continued fraction. These are the quotients of the successive divisions. They are sometimes called the *partial quotients* of the original fraction.

One of the classical uses for the Euclidean algorithm is to find the greatest common divisor of two given numbers. If one applies the algorithm to two numbers \(p\) and \(q\), dividing the smaller into the larger, then the remainder into the first divisor, and so on, then the greatest common divisor of \(p\) and \(q\) turns out to be the last nonzero remainder. For example, starting with \(p = 72\) and \(q = 201\) the calculation is shown at the right, and the last nonzero remainder is 3, which is the greatest common divisor of 72 and 201. (In fact the fraction 201/72 equals 67/24, which explains why the successive quotients for this example are the same as in the preceding example.) It is easy to see from the displayed equations why 3 has to be the greatest common divisor of 72 and 201, since from the first equation it follows that any divisor of 72 and 201 must also divide 57, then the second equation shows it must divide 15, the third equation then shows it must divide 12, and the fourth equation shows it must divide 3, the last nonzero remainder. Conversely, if a number divides the last nonzero remainder 3, then the last equation shows it must also divide the 12, and the next-to-last equation then shows it must divide 15, and so on until we conclude that it divides all the numbers not in the shaded rectangle, including the original two numbers 72 and 201. The same reasoning applies in general.

A more obvious way to try to compute the greatest common divisor of two numbers would be to factor each of them into a product of primes, then look to see which primes occurred as factors of both, and to what power. But to factor a large number into its prime factors is a very laborious and time-consuming process. For example, even a large computer would have a hard time factoring a number of a hundred digits into primes, so it would not be feasible to find the greatest common divisor of a pair of hundred-digit numbers this way. However, the computer would have no trouble at all applying the Euclidean algorithm to find their greatest common divisor.

Having seen what continued fractions are, let us now see what they have to do with the Farey diagram. Some examples will illustrate this best, so let us first look at the continued fraction for 7/16 again. This has 2, 3, 2 as its sequence of partial quotients.
We use these three numbers to build a strip of three large triangles subdivided into 2, 3, and 2 smaller triangles, from left to right:

\[
\frac{7}{16} = \frac{1}{2 + \frac{1}{3 + \frac{1}{2}}}
\]

We can think of the diagram as being formed from three “fans”, where the first fan is made from the first 2 small triangles, the second fan from the next 3 small triangles, and the third fan from the last 2 small triangles. Now we begin labeling the vertices of this strip. On the left edge we start with the labels 1/0 and 0/1. Then we use the mediant rule for computing the third label of each triangle in succession as we move from left to right in the strip. Thus we insert, in order, the labels 1/1, 1/2, 1/3, 2/5, 3/7, 4/9, and finally 7/16.

Was it just an accident that the final label was the fraction 7/16 that we started with, or does this always happen? Doing more examples should help us decide. Here is a second example:

\[
\frac{9}{31} = \frac{1}{3 + \frac{1}{2 + \frac{1}{4}}}
\]

Again the final vertex on the right has the same label as the fraction we started with. The reader is encouraged to try more examples to make sure we are not rigging things to get a favorable outcome by only choosing examples that work.

In fact this always works for fractions \( p/q \) between 0 and 1. For fractions larger than 1 the procedure works if we modify it by replacing the label 0/1 with the initial integer \( a_0/1 \) in the continued fraction \( a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}} \). This is illustrated by the 67/24 example:

\[
\frac{67}{24} = 2 + \frac{1}{1 + \frac{1}{3 + \frac{1}{1 + \frac{1}{4}}}}
\]

For comparison, here is the corresponding strip for the reciprocal, 24/67:
Now let us see how all this relates to the Farey diagram. Since the rule for labeling vertices in the triangles along the horizontal strip for a fraction \( p/q \) is the mediant rule, each of the triangles in the strip is a triangle in the Farey diagram, somewhat distorted in shape, and the strip of triangles can be regarded as a sequence of adjacent triangles in the diagram. Here is what this looks like for the fraction \( 7/16 \) in the circular Farey diagram, slightly distorted for the sake of visual clarity:

In the strip of triangles for a fraction \( p/q \) there is a zigzag path from \( 1/0 \) to \( p/q \) that we have indicated by the heavily shaded edges. The vertices that this zigzag path passes through have a special significance. They are the fractions that occur as the values of successively larger initial portions of the continued fraction, as illustrated in the following example:

These fractions are called the *convergents* for the given fraction. Thus the convergents for \( 67/24 \) are \( 2, 3, 11/4, 14/5, \) and \( 67/24 \) itself.

From the preceding examples one can see that each successive vertex label \( p_t/q_t \) along the zigzag path for a continued fraction \( \frac{p}{q} = a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) is
computed in terms of the two preceding vertex labels according to the rule

\[ \frac{p_i}{q_i} = \frac{a_i p_{i-1} + p_{i-2}}{a_i q_{i-1} + q_{i-2}} \]

This is because the mediant rule is being applied \( a_i \) times, ‘adding’ \( p_{i-1}/q_{i-1} \) to the previously obtained fraction each time until the next label \( p_i/q_i \) is obtained.

It is interesting to see what the zigzag paths corresponding to continued fractions look like in the upper half-plane Farey diagram. The next figure shows the simple example of the continued fraction for \( 3/8 \). We can see here that the five triangles of the strip correspond to the four curvilinear triangles lying directly above \( 3/8 \) in the Farey diagram, plus the fifth ‘triangle’ extending upward to infinity, bounded on the left and right by the vertical lines above \( 0/1 \) and \( 1/1 \), and bounded below by the semicircle from \( 0/1 \) to \( 1/1 \).

This example is typical of the general case, where the zigzag path for a continued fraction \( \frac{p}{q} = a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) becomes a ‘pinball path’ in the Farey diagram, starting down the vertical line from \( 1/0 \) to \( a_0/1 \), then turning left across \( a_1 \) triangles, then right across \( a_2 \) triangles, then left across \( a_3 \) triangles, continuing to alternate left and right turns until reaching the final vertex \( p/q \). Two consequences of this are:

1. The convergents are alternately smaller than and greater than \( p/q \).
2. The triangles that form the strip of triangles for \( p/q \) are exactly the triangles in the Farey diagram that lie directly above the point \( p/q \) on the \( x \)-axis.
Here is a general statement describing the relationship between continued fractions and the Farey diagram that we have observed in all our examples so far:

**Theorem 2.1.** The convergents for a continued fraction \( \frac{p}{q} = a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) are the vertices along a zigzag path consisting of a finite sequence of edges in the Farey diagram, starting at \( \frac{1}{0} \) and ending at \( \frac{p}{q} \). The path starts along the edge from \( \frac{1}{0} \) to \( \frac{a_0}{1} \), then turns left across a fan of \( a_1 \) triangles, then right across a fan of \( a_2 \) triangles, etc., finally ending at \( \frac{p}{q} \).

In particular, since every positive rational number has a continued fraction expansion, we see that every positive rational number occurs eventually as the label of some vertex in the positive half of the diagram. All negative rational numbers then occur as labels in the negative half.

**Proof of the Theorem:** The continued fraction \( \frac{p}{q} = a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) determines a strip of triangles:

![Diagram of Farey diagram with continued fraction](image)

We will show that the label \( \frac{p_n}{q_n} \) on the final vertex in this strip is equal to \( \frac{p}{q} \), the value of the continued fraction. Replacing \( n \) by \( i \), we conclude that this holds also for each initial segment \( a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_i} \) of the continued fraction. This is just saying that the vertices \( \frac{p_i}{q_i} \) along the strip are the convergents to \( \frac{p}{q} \), which is what the theorem claims.

To prove that \( \frac{p_n}{q_n} = \frac{p}{q} \) we will use \( 2 \times 2 \) matrices. Consider the product

\[
p = \begin{pmatrix} 1 & a_0 \\ 0 & 1 \end{pmatrix} \begin{pmatrix} 0 & 1 \\ 1 & a_1 \end{pmatrix} \begin{pmatrix} 0 & 1 \\ 1 & a_2 \end{pmatrix} \cdots \begin{pmatrix} 0 & 1 \\ 1 & a_n \end{pmatrix}
\]

We can multiply this product out starting either from the left or from the right. Suppose first that we multiply starting at the left. The initial matrix is \( \begin{pmatrix} 1 & a_0 \\ 0 & 1 \end{pmatrix} \) and we can view the two columns of this matrix as the two fractions \( \frac{1}{0} \) and \( \frac{a_0}{1} \) labeling the left edge of the strip of triangles. When we multiply this matrix by the next matrix we get

\[
\begin{pmatrix} 1 & a_0 \\ 0 & 1 \end{pmatrix} \begin{pmatrix} 0 & 1 \\ 1 & a_1 \end{pmatrix} = \begin{pmatrix} a_0 & 1 + a_0 a_1 \\ 1 & a_1 \end{pmatrix} = \begin{pmatrix} p_0 & p_1 \\ a_0 & a_1 \end{pmatrix}
\]

The two columns here give the fractions at the ends of the second edge of the zigzag path. The same thing happens for subsequent matrix multiplications, as multiplying by the next matrix in the product takes the matrix corresponding to one edge of the
zigzag path to the matrix corresponding to the next edge:

\[
\begin{pmatrix}
  p_{i-2} & p_{i-1} \\
  q_{i-2} & q_{i-1}
\end{pmatrix}
\begin{pmatrix}
  0 & 1 \\
  1 & a_i
\end{pmatrix}
= \begin{pmatrix}
  p_{i-1} & p_{i-2} + a_ip_{i-1} \\
  q_{i-1} & q_{i-2} + a_iq_{i-1}
\end{pmatrix}
= \begin{pmatrix}
  p_{i-1} & p_i \\
  q_{i-1} & q_i
\end{pmatrix}
\]

In the end, when all the matrices have been multiplied, we obtain the matrix corresponding to the last edge in the strip from \( p_{n-1}/q_{n-1} \) to \( p_n/q_n \). Thus the second column of the product \( P \) is \( p_n/q_n \), and what remains is to show that this equals the value \( p/q \) of the continued fraction \( a_0 + 1/a_1 + 1/a_2 + \cdots + 1/a_n \).

The value of the continued fraction \( a_0 + 1/a_1 + 1/a_2 + \cdots + 1/a_n \) is computed by working from right to left. If we let \( r_i/s_i \) be the value of the tail \( 1/a_i + 1/a_{i+1} + \cdots + 1/a_n \) of the continued fraction, then \( r_n/s_n = 1/a_n \) and we have

\[
\frac{r_i}{s_i} = \frac{1}{a_i + \frac{r_{i+1}}{s_{i+1}}} = \frac{s_{i+1}}{a_is_{i+1} + r_{i+1}} \quad \text{and finally} \quad \frac{p}{q} = a_0 + \frac{r_1}{s_1} = \frac{a_0s_1 + r_1}{s_1}
\]

In terms of matrices this implies that we have

\[
\begin{pmatrix}
  r_n \\
  s_n
\end{pmatrix}
= \begin{pmatrix}
  1 & 0 \\
  a_n & 1
\end{pmatrix}
\begin{pmatrix}
  r_{i+1} \\
  s_{i+1}
\end{pmatrix}
= \begin{pmatrix}
  s_{i+1} \\
  r_{i+1} + a_is_{i+1}
\end{pmatrix} = \begin{pmatrix}
  r_i \\
  s_i
\end{pmatrix}
\]

and

\[
\begin{pmatrix}
  1 & 0 \\
  a_0 & 1
\end{pmatrix}
\begin{pmatrix}
  r_1 \\
  s_1
\end{pmatrix}
= \begin{pmatrix}
  r_1 + a_0s_1 \\
  s_1
\end{pmatrix} = \begin{pmatrix}
  p \\
  q
\end{pmatrix}
\]

This means that when we multiply out the product \( P \) starting from the right, then the second columns will be successively \( \begin{pmatrix} r_n \\ s_n \end{pmatrix} \), \( \begin{pmatrix} r_{n-1} \\ s_{n-1} \end{pmatrix} \), \ldots \, \begin{pmatrix} r_1 \\ s_1 \end{pmatrix} \) and finally \( \begin{pmatrix} p \\ q \end{pmatrix} \).

We already showed this second column is \( \begin{pmatrix} p_n \\ q_n \end{pmatrix} \), so \( p/q = p_n/q_n \) and the proof is complete.

An interesting fact that can be deduced from the preceding proof is that for a continued fraction \( \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) with no initial integer \( a_0 \), if we reverse the order of the numbers \( a_i \), this leaves the denominator unchanged. For example

\[
\frac{1}{2} + \frac{1}{3} + \frac{1}{4} = \frac{13}{30} \quad \text{and} \quad \frac{1}{4} + \frac{1}{3} + \frac{1}{2} = \frac{7}{30}
\]

To see why this must always be true we use the operation of transposing a matrix to interchange its rows and columns. For a \( 2 \times 2 \) matrix this just amounts to interchanging the upper-right and lower-left entries:

\[
\begin{pmatrix}
  a & c \\
  b & d
\end{pmatrix}^T = \begin{pmatrix}
  a & b \\
  c & d
\end{pmatrix}
\]

Transposing a product of matrices reverses the order of the factors: \( (AB)^T = B^TA^T \), as can be checked by direct calculation. In the product

\[
\begin{pmatrix}
  0 & 1 \\
  1 & a_1
\end{pmatrix}
\begin{pmatrix}
  0 & 1 \\
  1 & a_2
\end{pmatrix}\cdots\begin{pmatrix}
  0 & 1 \\
  1 & a_n
\end{pmatrix}
= \begin{pmatrix}
  p_{n-1} & p_n \\
  q_{n-1} & q_n
\end{pmatrix}
\]
the individual matrices on the left side of the equation are symmetric with respect to transposition, so the transpose of the product is obtained by just reversing the order of the factors:

\[
\begin{pmatrix} 0 & 1 \\ 1 & a_n \end{pmatrix} \begin{pmatrix} 0 & 1 \\ 1 & a_{n-1} \end{pmatrix} \ldots \begin{pmatrix} 0 & 1 \\ 1 & a_1 \end{pmatrix} = \begin{pmatrix} p_{n-1} & q_{n-1} \\ p_n & q_n \end{pmatrix}
\]

Thus the denominator \( q_n \) is unchanged, as claimed.

There is also a fairly simple relationship between the numerators. In the example of \( 13/30 \) and \( 7/30 \) we see that the product of the numerators, 91, is congruent to 1 modulo the denominator. In the general case the product of the numerators is \( p_n q_{n-1} \) and this is congruent to \((-1)^{n+1}\) modulo the denominator \( q_n \). To verify this, we note that the determinant of each factor \( \begin{pmatrix} 0 & 1 \\ 1 & a_i \end{pmatrix} \) is \(-1\) so since the determinant of a product is the product of the determinants, we have \( p_{n-1} q_n - p_n q_{n-1} = (-1)^n \), which says that \( p_n q_{n-1} \) is congruent to \((-1)^{n+1}\) modulo \( q_n \).

2. Determinants and the Farey Diagram

We constructed the Farey diagram by an inductive procedure, inserting successive edges according to the mediant rule, but there is another rule that can be used to characterize the edges in the diagram:

**Theorem 2.2.** In the Farey diagram, two vertices labeled \( a/b \) and \( c/d \) are joined by an edge if and only if the determinant \( ad - bc \) of the matrix \( \begin{pmatrix} a & c \\ b & d \end{pmatrix} \) is equal to \( \pm 1 \).

**Proof:** First we show that for an arbitrary edge in the diagram joining \( a/b \) to \( c/d \), the associated matrix \( \begin{pmatrix} a & c \\ b & d \end{pmatrix} \) has determinant \( \pm 1 \). This is obviously true for the edges in the two largest triangles in the circular version of the diagram. For the smaller triangles we proceed by induction. The figure at the right shows the three matrices corresponding to the edges of one of these smaller triangles. By induction we assume we know that \( ad - bc = \pm 1 \) for the long edge of the triangle. Then the determinant condition holds also for the two shorter edges of the triangle since \( a(b + d) - b(a + c) = ad - bc \) and \( (a + c)d - (b + d)c = ad - bc \).

Before proving the converse let us pause to apply what we have shown so far to deduce a basic fact about the Farey diagram that was mentioned but not proved when we first constructed the diagram:

**Corollary 2.3.** The mediant rule for labeling the vertices in the Farey diagram always produces labels \( a/b \) that are fractions in lowest terms.

**Proof:** Consider an edge joining a vertex labeled \( a/b \) to some other vertex labeled \( c/d \). We know from the argument given above that \( ad - bc = \pm 1 \). This equation
implies that $a$ and $b$ can have no common divisor greater than 1 since any common divisor of $a$ and $b$ must divide the products $ad$ and $bc$, hence also the difference $ad - bc = \pm 1$, but the only divisors of $\pm 1$ are $\pm 1$.

Now we return to proving the converse half of the theorem, which says that there is an edge joining $a/b$ to $c/d$ whenever $ad - bc = \pm 1$. To do this we will examine how all the edges emanating from a fixed vertex $a/b$ are related. To begin, if $a/b = 0/1$ then the matrices $\begin{pmatrix} 0 & c \\ 1 & d \end{pmatrix}$ with determinant $\pm 1$ are the matrices $\begin{pmatrix} 0 & \pm 1 \\ 1 & d \end{pmatrix}$, and these correspond exactly to the edges in the diagram from $0/1$ to $\pm 1/d$. There is a similar exact correspondence for the edges from $1/0$. For the other vertices $a/b$, the example $a/b = 5/8$ is shown in the left half of the figure below. The first edges drawn to this vertex come from $2/3$ and $3/5$, and after this all the other edges from $5/8$ are drawn in turn. As one can see, they are all obtained by adding $(5,8)$ to $(2,3)$ or $(3,5)$ repeatedly. If we choose any one of these edges from $5/8$, say the edge to $2/3$ for example, then the edges from $5/8$ have their other endpoints at the fractions $(2 + 5k)/(3 + 8k)$ as $k$ ranges over all integers, with positive values of $k$ giving the edges on the upper side of the edge to $2/3$ and negative values of $k$ giving the edges on the lower side of the edge to $2/3$.

The same thing happens for an arbitrary value of $a/b$ as shown in the right half of the figure, where $a/b$ initially arises as the mediant of $c/d$ and $e/f$. In this case if we choose the edge to $c/d$ as the starting edge, then the other edges go from $a/b$ to $(c + ka)/(d + kb)$. In particular, when $k = -1$ we get the edge to $(c - a)/(d - b) = (a - c)/(b - d) = e/f$.

To finish the argument we need to know how the various matrices $\begin{pmatrix} a & x \\ b & y \end{pmatrix}$ of determinant $ay - bx = \pm 1$ having the same first column are related. This can be deduced from the following result about integer solutions of linear equations with integer coefficients:

**Lemma 2.4.** Suppose $a$ and $b$ are integers with no common divisor greater than 1. If one solution of $ay - bx = n$ is $(x, y) = (c, d)$, then the general solution is $(x, y) = (c + ka, d + kb)$ for $k$ an arbitrary integer.
The proof will use the same basic argument as is used in linear algebra to show that the general solution of a system of nonhomogeneous linear equations is obtained from any particular solution by adding the general solution of the associated system of homogeneous equations.

Before giving the proof let us introduce a convenient bit of standard terminology. If two integers \( a \) and \( b \) have no common divisor greater than 1 then \( a \) and \( b \) are said to be coprime. One can also say that \( a \) is coprime to \( b \), or symmetrically, \( b \) is coprime to \( a \). A commonly used synonym for coprime is relatively prime.

**Proof**: One solution \((x, y) = (c, d)\) of \(ay - bx = n\) is given. For an arbitrary solution \((x, y)\) we look at the difference \((x_0, y_0) = (x - c, y - d)\). This satisfies \(ay_0 - bx_0 = 0\), or in other words, \(ay_0 = bx_0\). Since \(a\) and \(b\) are coprime, the equation \(ay_0 = bx_0\) implies that \(x_0\) must be a multiple of \(a\) and \(y_0\) must be a multiple of \(b\), in fact the same multiple in both cases so that the equation becomes \(a(kb) = b(ka)\). Thus we have \((x_0, y_0) = (ka, kb)\) for some integer \(k\). Thus every solution of \(ay - bx = n\) has the form \((x, y) = (c + x_0, d + y_0) = (c + ka, d + kb)\), and it is clear that these formulas for \(x\) and \(y\) give solutions for all values of \(k\). \(\square\)

Now we can easily finish the proof of the theorem. The lemma in the cases \(n = \pm 1\) implies that the edges in the Farey diagram with \(a/b\) at one endpoint account for all matrices \(\begin{pmatrix} a & x \\ b & y \end{pmatrix}\) of determinant \(ay - bx = \pm 1\). \(\square\)

There is some ambiguity in the correspondence between edges of the Farey diagram and matrices \(\begin{pmatrix} a & c \\ b & d \end{pmatrix}\) of determinant \(\pm 1\). For one thing, either column of the matrix can be multiplied by \(-1\), changing the sign of the determinant without changing the value of the fractions \(a/b\) and \(c/d\). This ambiguity can be eliminated by choosing all of \(a, b, c,\) and \(d\) to be positive for edges in the upper half of the circular Farey diagram, and choosing just the numerators \(a\) and \(c\) to be negative for edges in the lower half of the diagram. The only other ambiguity is that both \(\begin{pmatrix} a & c \\ b & d \end{pmatrix}\) and \(\begin{pmatrix} c & a \\ d & b \end{pmatrix}\) correspond to the same edge. This ambiguity can be eliminated by orienting the edges by placing an arrowhead on each edge pointing from the vertex corresponding to the first column of the matrix to the vertex corresponding to the second column. Changing the orientation of an edge switches the two columns of the matrix, which changes the sign of the determinant.

The identity matrix \(\begin{pmatrix} 1 & 0 \\ 0 & 1 \end{pmatrix}\) has determinant \(+1\) and corresponds to the edge from \(1/0\) to \(0/1\) oriented from left to right in the circular diagram. We can use this orientation to give orientations to all other edges when we build the diagram using the mediant rule. In the upper half of the diagram this makes all edges be oriented toward the right, or in
other words from \( a/b \) to \( c/d \) with \( a/b > c/d \). With this orientation, all the corresponding matrices have determinant \(+1\) since \( \begin{pmatrix} 1 & 0 \\ 0 & 1 \end{pmatrix} \) has determinant \(+1\) and we have seen that the determinant doesn’t change when we add new edges by the mediant rule. When we use the mediant rule to construct the lower half of the diagram we have to start with \(-1/0\) instead of \(1/0\). This means that we are starting with the matrix \( \begin{pmatrix} -1 & 0 \\ 0 & 1 \end{pmatrix} \) instead of \( \begin{pmatrix} 1 & 0 \\ 0 & 1 \end{pmatrix} \). Since the determinant of \( \begin{pmatrix} -1 & 0 \\ 0 & 1 \end{pmatrix} \) is \(-1\), this means that the edges in the lower half of the diagram, when oriented toward the right as in the upper half, correspond to matrices of determinant \(-1\).

### 3. The Diophantine Equation \( ax + by = n \)

The Euclidean algorithm and continued fractions can be used to compute all the integer solutions of a linear equation \( ax + by = n \) when \( a, b, \) and \( n \) are integers. We can assume neither \( a \) nor \( b \) is zero, otherwise the equation is rather trivial. Changing the signs of \( x \) or \( y \) if necessary, we can rewrite the equation in the form \( ax - by = n \) where \( a \) and \( b \) are both positive.

If \( a \) and \( b \) have greatest common divisor \( d > 1 \), then since \( d \) divides \( a \) and \( b \) it must divide \( ax - by \), so \( d \) must divide \( n \) if the equation is to have any solutions at all. If \( d \) does divide \( n \) we can divide both sides of the equation by \( d \) to get a new equation of the same type as the original one and having the same solutions, but with the new coefficients \( a \) and \( b \) being coprime. For example, the equation \( 6x - 15y = 21 \) reduces in this way to the equation \( 2x - 5y = 7 \). Thus we can assume from now on that \( a \) and \( b \) are coprime.

The Lemma from a page or two back shows how to find the general solution of \( ax - by = n \) once we have found one particular solution. To find a particular solution it suffices to do the case \( n = 1 \) since if we have a solution of \( ax - by = 1 \), we can multiply \( x \) and \( y \) by \( n \) to get a solution of \( ax - by = n \). For small values of \( a \) and \( b \) a solution of \( ax - by = 1 \) can be found more or less by inspection since the equation \( ax - by = 1 \) says that we have a multiple of \( a \) that is 1 greater than a multiple of \( b \). For example, for the equation \( 2x - 5y = 1 \) the smallest multiples of 2 that is one greater than a multiple of 5 is \( 2 \cdot 3 > 5 \cdot 1 \), so a solution of \( 2x - 5y = 1 \) is \( (x, y) = (3, 1) \). A solution of \( 2x - 5y = 7 \) is then \( (x, y) = (21, 7) \). By the earlier Lemma, the general solution of \( 2x - 5y = 7 \) is \( (x, y) = (21 + 5k, 7 + 2k) \) for arbitrary integers \( k \). The smallest positive solution is \( (6, 1) \), obtained by setting \( k = -3 \). This means we could also write the general solution as \( (6 + 5k, 1 + 2k) \).

Solutions of \( ax - by = 1 \) always exist when \( a \) and \( b \) are coprime, and a way to find one is to find an edge in the Farey diagram with \( a/b \) at one end of the edge. This can be done by using the Euclidean algorithm to compute the strip of triangles from \( 1/0 \) to \( a/b \). As an example, let us solve \( 67x - 24y = 1 \). We already computed the strip of triangles for \( 67/24 \) earlier in the chapter. The vertex preceding \( 67/24 \) in the zigzag path is \( 14/5 \) and this vertex lies above \( 67/24 \) so we have \( 14/5 > 67/24 \) and hence
the matrix \( \begin{pmatrix} 14 & 67 \\ 5 & 24 \end{pmatrix} \) has determinant +1. (One can easily distinguish determinant +1 from determinant -1 by computing just the last digit of the determinant.) Thus we have 14 \cdot 24 - 5 \cdot 67 = 1 so one solution of 67x - 24y = 1 is \((x, y) = (-5, -14)\) and the general solution is \((x, y) = (-5 + 24k, -14 + 67k)\). We could also use the edge from 53/19 to 67/24, so \( \begin{pmatrix} 67 & 53 \\ 24 & 19 \end{pmatrix} \) has determinant +1, yielding another formula for the general solution \((19 + 24k, 53 + 67k)\).

From a geometric point of view, finding the integer solutions of \(ax + by = n\) is finding the points on the line \(ax + by = n\) in the \(xy\)-plane having both coordinates integers. The points in the plane having both coordinates integers form a square grid called the integer lattice. Thus we wish to see which points in the integer lattice lie on the line \(ax + by = n\). This equation can be written in the form \(y = mx + b\) where the slope \(m\) and the \(y\)-intercept \(b\) are both rational. Conversely, an equation \(y = mx + b\) with \(m\) and \(b\) rational can be written as an equation \(ax + by = n\) with \(a, b,\) and \(n\) integers by multiplying through by a common denominator of \(m\) and \(b\). Sometimes the equation \(ax + by = n\) has no integer solutions, as we have seen, namely when \(n\) is not a multiple of the greatest common divisor of \(a\) and \(b\), for example the equation \(2x + 2y = 1\). In these cases the line \(ax + by = n\) passes through no integer lattice points. In the opposite case that there does exist an integer solution, there are infinitely many, and they correspond to integer lattice points spaced at equal intervals along the line.

4. Infinite Continued Fractions

We have seen that all rational numbers can be represented as continued fractions \(a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}}\), but what about irrational numbers? It turns out that these can be represented as infinite continued fractions \(a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \frac{1}{a_3 + \cdots}}}\). A simple example is \(\frac{1}{1 + \frac{1}{1 + \frac{1}{1 + \cdots}}}\), or in its expanded form:

\[
\frac{1}{1 + \frac{1}{1 + \frac{1}{1 + \cdots}}}
\]

The corresponding strip of triangles is infinite:
Notice that these fractions after \( 1/0 \) are the successive ratios of the famous Fibonacci sequence \( 0, 1, 1, 2, 3, 5, 8, 13, 21, \cdots \) where each number is the sum of its two predecessors. The sequence of convergents is thus \( 0/1, 1/1, 1/2, 2/3, 3/5, 5/8, 8/13, \cdots \), the vertices along the zigzag path. The way this zigzag path looks in the standard Farey diagram is shown in the figure at the right. What happens when we follow this path farther and farther? The path consists of an infinite sequence of semicircles, each one shorter than the preceding one and sharing a common endpoint. The left endpoints of the semicircles form an increasing sequence of numbers which have to be approaching a certain limiting value \( x \). We know \( x \) has to be finite since it is certainly less than each of the right-hand endpoints of the semicircles, the convergents \( 1/1, 2/3, 5/8, \cdots \). Similarly the right endpoints of the semicircles form a decreasing sequence of numbers approaching a limiting value \( y \) greater than each of the left-hand endpoints \( 0/1, 1/2, 3/5, \cdots \). Obviously \( x \leq y \). Is it possible that \( x \) is not equal to \( y \)? If this happened, the infinite sequence of semicircles would be approaching the semicircle from \( x \) to \( y \). Above this semicircle there would then be an infinite number of semicircles, all the semicircles in the infinite sequence. Between \( x \) and \( y \) there would have to be a rational numbers \( p/q \) (between any two real numbers there is always a rational number), so above this rational number there would be an infinite number of semicircles, hence an infinite number of triangles in the Farey diagram. But we know that there are only finitely many triangles above any rational number \( p/q \), namely the triangles that appear in the strip for the continued fraction for \( p/q \). This contradiction shows that \( x \) has to be equal to \( y \). Thus the sequence of convergents along the edges of the infinite strip of triangles converges to a unique real number \( x \). (This is why the convergents are called convergents.)

This argument works for arbitrary infinite continued fractions, so we have shown the following general result:

**Proposition 2.5.** For every infinite continued fraction \( a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \frac{1}{a_3} + \cdots \) the convergents converge to a unique limit.

This limit is by definition the value of the infinite continued fraction. There is a simple method for computing the value in the example involving Fibonacci numbers. We begin by setting

\[
x = \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \cdots
\]

Then if we take the reciprocals of both sides of this equation we get

\[
\frac{1}{x} = 1 + \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \cdots
\]
The right side of this equation is just $1 + x$, so we can easily solve for $x$:

\[
\frac{1}{x} = 1 + x \\
1 = x + x^2 \\
x^2 + x - 1 = 0 \\
x = \frac{-1 \pm \sqrt{5}}{2}
\]

We know $x$ is positive, so this rules out the negative root and we are left with the final value $x = (-1 + \sqrt{5})/2$. (This number, approximately .618, goes by the name of the golden ratio because of its many interesting and beautiful properties.)

**Proposition 2.6.** Every irrational number has an expression as an infinite continued fraction, and this continued fraction is unique.

**Proof:** In the Farey diagram consider the vertical line $L$ going upward from a given irrational number $x$ on the $x$-axis. The lower endpoint of $L$ is not a vertex of the Farey diagram since $x$ is irrational. Thus as we move downward along $L$ we cross a sequence of triangles, entering each triangle by crossing its upper edge and exiting the triangle by crossing one of its two lower edges. When we exit one triangle we are entering another, the one just below it, so the sequence of triangles and edges we cross must be infinite. The left and right endpoints of the edges in the sequence must be approaching the single point $x$ by the argument we gave in the preceding proposition, so the edges themselves are approaching $x$. Thus the triangles in the sequence form a single infinite strip consisting of an infinite sequence of fans with their pivot vertices on alternate sides of the strip. The zigzag path along this strip gives a continued fraction for $x$.

For the uniqueness, we have seen that an infinite continued fraction for $x$ corresponds to a zigzag path in the infinite strip of triangles lying above $x$. This set of triangles is unique so the strip is unique, and there is only one path in this strip that starts at $1/0$ and then does left and right turns alternately, starting with a left turn. The initial turn must be to the left because the first two convergents are $a_0$ and $a_0 + \frac{1}{a_1}$, with $a_0 + \frac{1}{a_1} > a_0$ since $a_1 > 0$. After the path traverses the initial edge from $1/0$ to $a_0/1$ no subsequent edge of the path can be in the border of the strip since this would entail two successive left turns or two successive right turns. □

The arguments we have just given can be used to prove a fact about the standard Farey diagram that we have been taking more or less for granted. This is the fact that the triangles in the diagram completely cover the upper halfplane. In other words, every point $(x, y)$ with $y > 0$ lies either in the interior of some triangle or on the common edge between two triangles. To see why, consider the vertical line $L$ in the upper halfplane through the given point $(x, y)$. If $x$ is an integer then $(x, y)$ is on one of the vertical edges of the diagram. Thus we can assume $x$ is not an integer
and hence \( L \) is not one of the vertical edges of the diagram. The line \( L \) will then be contained in the strip of triangles corresponding to the continued fraction for \( x \). This is a finite strip if \( x \) is rational and an infinite strip if \( x \) is irrational. In either case the point \((x, y)\), being in \( L \), will be in one of the triangles of the strip or on an edge separating two triangles in the strip. This proves what we wanted to prove.

To compute the infinite continued fraction \( a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \frac{1}{a_3 + \cdots}}} \) for a given irrational number \( x \) we can follow the same procedure as for rational numbers, but it doesn’t terminate after a finite number of steps. Recall the original example that we did:

\[
\begin{align*}
\frac{67}{24} &= 2 + \frac{19}{24} = 2 + \frac{1}{24/19} = 2 + \frac{1}{1 + \frac{5}{19}} \\
&= 2 + \frac{1}{1 + \frac{1}{3 + 4/5}} = 2 + \frac{1}{1 + \frac{1}{3 + \frac{1}{5/4}}} = 2 + \frac{1}{1 + \frac{1}{3 + \frac{1}{4}}}
\end{align*}
\]

The sequence of steps is the following:

1. Write \( x = a_0 + r_1 \) where \( a_0 \) is an integer and \( 0 \leq r_1 < 1 \)
2. Write \( 1/r_1 = a_1 + r_2 \) where \( a_1 \) is an integer and \( 0 \leq r_2 < 1 \)
3. Write \( 1/r_2 = a_2 + r_3 \) where \( a_2 \) is an integer and \( 0 \leq r_3 < 1 \)

and so on, repeatedly. Thus one first finds the largest integer \( a_0 \leq x \), with \( r_1 \) the ‘remainder’, then one inverts \( r_1 \) and finds the greatest integer \( a_1 \leq 1/r_1 \), with \( r_2 \) the remainder, etc.

Here is how this works for \( x = \sqrt{2} \):

1. \( \sqrt{2} = 1 + (\sqrt{2} - 1) \) where \( a_0 = 1 \) since \( \sqrt{2} \) is between 1 and 2. Before going on to step (2) we have to compute \( \frac{1}{r_1} = \frac{1}{\sqrt{2} - 1} \). Multiplying numerator and denominator by \( \sqrt{2} + 1 \) gives \( \frac{1}{\sqrt{2} - 1} = \frac{\sqrt{2} + 1}{\sqrt{2} - 1} = \sqrt{2} + 1 \). This is the number we use in the next step.

2. \( \sqrt{2} + 1 = 2 + (\sqrt{2} - 1) \) since \( \sqrt{2} + 1 \) is between 2 and 3.

Notice that something unexpected has happened: The remainder \( r_2 = \sqrt{2} - 1 \) is exactly the same as the previous remainder \( r_1 \). There is then no need to do the calculation of \( \frac{1}{r_2} = \frac{1}{\sqrt{2} - 1} \) since we know it will have to be \( \sqrt{2} + 1 \). This means that the next step (3) will be exactly the same as step (2), and the same will be true for all subsequent steps. Hence we get the continued fraction

\[
\sqrt{2} = 1 + \frac{1}{2} + \frac{1}{2} + \frac{1}{2} + \cdots
\]

We can check this calculation by finding the value of the continued fraction in the same way that we did earlier for \( \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \cdots \). First we set \( x = \frac{1}{2} + \frac{1}{2} + \frac{1}{2} + \cdots \).
Taking reciprocals gives \(1/x = 2 + 1/2 + 1/2 + \cdots = 2 + x\). This leads to the quadratic equation \(x^2 + 2x - 1 = 0\), which has roots \(x = -1 \pm \sqrt{2}\). Since \(x\) is positive we can discard the negative root. Thus we have \(-1 + \sqrt{2} = 1/2 + 1/2 + 1/2 + \cdots\). Adding 1 to both sides of this equation gives the formula for \(\sqrt{2}\) as a continued fraction.

We can get good rational approximations to \(\sqrt{2}\) by computing the convergents in its continued fraction \(1 + 1/2 + 1/2 + 1/2 + \cdots\). It’s a little easier to compute the convergents in \(2 + 1/2 + 1/2 + 1/2 + \cdots = 1 + \sqrt{2}\) and then subtract 1 from each of these. For \(2 + 1/2 + 1/2 + 1/2 + \cdots\) there is a nice pattern to the convergents:

\[
\begin{array}{cccccccc}
2 & 5 & 12 & 29 & 70 & 169 & 408 & 985 \\
1' & 2' & 5' & 12' & 29' & 70' & 169' & 408' \\
\end{array}
\]

Notice that the sequence of numbers 1, 2, 5, 12, 29, 70, 169, \(\cdots\) is constructed in a way somewhat analogous to the Fibonacci sequence, except that each number is twice the preceding number plus the number before that. (It’s easy to see why this has to be true, because each convergent is constructed from the previous one by inverting the fraction and adding 2.) After subtracting 1 from each of these fractions we get the convergents to \(\sqrt{2}\):

\[
\sqrt{2} = 1.41421356 \cdots \\
1/1 = 1.00000000 \cdots \\
3/2 = 1.50000000 \cdots \\
7/5 = 1.40000000 \cdots \\
17/12 = 1.41666666 \cdots \\
41/29 = 1.41379310 \cdots \\
99/70 = 1.41428571 \cdots \\
239/169 = 1.41420118 \cdots \\
577/408 = 1.41421356 \cdots 
\]

We can compute the continued fraction for \(\sqrt{3}\) by the same method as for \(\sqrt{2}\), but something slightly different happens:

1. \(\sqrt{3} = 1 + (\sqrt{3} - 1)\) since \(\sqrt{3}\) is between 1 and 2. Computing \(\frac{1}{\sqrt{3} - 1}\), we have \(\frac{1}{\sqrt{3} - 1} = \frac{1}{\sqrt{3} - 1} \cdot \frac{\sqrt{3} + 1}{\sqrt{3} + 1} = \frac{\sqrt{3} + 1}{2}\).

2. \(\frac{\sqrt{3} + 1}{2} = 1 + (\frac{\sqrt{3} - 1}{2})\) since the numerator \(\sqrt{3} + 1\) of \(\frac{\sqrt{3} + 1}{2}\) is between 2 and 3. Now we have a remainder \(r_2 = \frac{\sqrt{3} - 1}{2}\) which is different from the previous remainder \(r_1 = \sqrt{3} - 1\), so we have to compute \(\frac{1}{r_2} = \frac{2}{\sqrt{3} - 1}\), namely \(\frac{2}{\sqrt{3} - 1} = \frac{2}{\sqrt{3} - 1} \cdot \frac{\sqrt{3} + 1}{\sqrt{3} + 1} = \sqrt{3} + 1\).

3. \(\sqrt{3} + 1 = 2 + (\sqrt{3} - 1)\) since \(\sqrt{3} + 1\) is between 2 and 3.

Now this remainder \(r_3 = \sqrt{3} - 1\) is the same as \(r_1\), so instead of the same step being repeated infinitely often, as happened for \(\sqrt{2}\), the same two steps will repeat infinitely often. This means we get the continued fraction

\[
\sqrt{3} = 1 + \frac{1}{1} + \frac{1}{2} + \frac{1}{1} + \frac{1}{2} + \frac{1}{1} + \frac{1}{2} + \cdots
\]
The answer is given by a theorem of Lagrange from around 1766:

\[ \sqrt{\text{irrational number}} \]

This example illustrates two other curious facts about the continued fraction for an irrational number \( \sqrt{\text{irrational number}} \):

(i) The last term of the period (12 in the example) is always twice the integer \( a_0 \) (the initial 6).

(ii) If the last term of the period is omitted, the preceding terms in the period form a palindrome, reading the same backwards as forwards.

We will see in Section 4.3 of Chapter 4 why these two properties have to be true.

It is natural to ask exactly which irrational numbers have continued fractions that are periodic, or at least eventually periodic, like for example

\[ \frac{1}{2} + \frac{1}{4} + \frac{1}{3} + \frac{1}{5} + \frac{1}{7} = \frac{1}{2} + \frac{1}{4} + \frac{1}{3} + \frac{1}{5} + \frac{1}{7} + \frac{1}{3} + \frac{1}{5} + \frac{1}{7} + \frac{1}{3} + \frac{1}{5} + \frac{1}{7} + \cdots \]

The answer is given by a theorem of Lagrange from around 1766:
Theorem 2.7 (Lagrange's Theorem). The irrational numbers whose continued fractions are eventually periodic are exactly the numbers of the form \( a + b\sqrt{n} \) where \( a \) and \( b \) are rational numbers, \( b \neq 0 \), and \( n \) is a positive integer that is not a square.

These numbers \( a + b\sqrt{n} \) are called quadratic irrationals because they are roots of quadratic equations with integer coefficients. The easier half of the theorem is the statement that the value of an eventually periodic infinite continued fraction is always a quadratic irrational. This can be proved by showing that the method we used for finding a quadratic equation satisfied by an eventually periodic continued fraction works in general. Rather than following this purely algebraic approach, however, we will develop a more geometric version of the procedure in the next chapter, so we will wait until then to give the argument that proves this half of Lagrange's Theorem.

The more difficult half of the theorem is the assertion that the continued fraction expansion of every quadratic irrational is eventually periodic. It is not at all apparent from the examples of \( \sqrt{2} \) and \( \sqrt{3} \) why this should be true in general, but in Chapter 5 we will develop some theory that will make it clear.

What can be said about the continued fraction expansions of irrational numbers that are not quadratic, such as \( 3\sqrt{2} \), \( \pi \), or \( e \), the base for natural logarithms? It happens that \( e \) has a continued fraction whose terms have a very nice pattern, even though they are not periodic or eventually periodic:

\[
e = 2 + \frac{1}{1 + \frac{1}{6 + \frac{1}{1 + \frac{1}{1 + \frac{1}{1 + \frac{1}{1 + \cdots}}}}}}
\]

where the terms are grouped by threes with successive even numbers as middle denominators. Even simpler are the continued fractions for certain numbers built from \( e \) that have arithmetic progressions for their denominators:

\[
\frac{e - 1}{e + 1} = \frac{1}{2} + \frac{1}{6} + \frac{1}{10} + \frac{1}{14} + \cdots
\]

\[
\frac{e^2 - 1}{e^2 + 1} = \frac{1}{1} + \frac{1}{3} + \frac{1}{5} + \frac{1}{7} + \cdots
\]

The continued fractions for \( e \) and \( (e - 1)/(e + 1) \) were discovered by Euler in 1737 while the formula for \( (e^2 - 1)/(e^2 + 1) \) was found by Lambert in 1766 as a special case of a slightly more complicated formula for \( (e^x - 1)/(e^x + 1) \).

For \( 3\sqrt{2} \) and \( \pi \), however, the continued fractions have no known pattern. For \( \pi \) the continued fraction begins

\[
\pi = 3 + \frac{1}{7 + \frac{1}{15 + \frac{1}{1 + \frac{1}{292 + \cdots}}}}
\]

Here the first four convergents are 3, 22/7, 333/106, and 355/113. We recognize 22/7 as the familiar approximation \( 3\frac{1}{7} \) to \( \pi \). The convergent 355/113 is a particularly good approximation to \( \pi \) since its decimal expansion begins 3.14159282 whereas \( \pi = 3.14159265 \cdots \). It is no accident that the convergent 355/113 obtained by truncating the continued fraction just before the 292 term gives a good approximation
to $\pi$ since it is a general fact that a convergent immediately preceding a large term in the continued fraction always gives an especially good approximation. This is because the next jump in the zigzag path in the Farey diagram will be rather small since it crosses a fan with a large number of triangles, and all succeeding jumps will of course be smaller still.

There are nice continued fractions for $\pi$ if one allows numerators larger than 1, as in the following formula discovered by Euler:

$$\pi = 3 + \frac{1}{6} + \frac{1}{6 + \frac{1}{6}} + \frac{1}{6 + \frac{1}{6}} + \frac{1}{6 + \frac{1}{6}} + \cdots$$

However, it is the continued fractions with numerator 1 that have the nicest properties, so we will not consider the more general sort in this book.

**Exercises**

1. (a) Compute the values of the continued fractions $\frac{1}{1} + \frac{1}{3} + \frac{1}{5} + \frac{1}{7}$ and $\frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{2}$.

   (b) Compute the continued fraction expansions of $\frac{19}{44}$ and $\frac{101}{1020}$.

2. (a) Compute the continued fraction for $\frac{38}{83}$ and display the steps of the Euclidean algorithm as a sequence of equations involving just integers.

   (b) For the same number $\frac{38}{83}$ compute the associated strip of triangles (with large triangles subdivided into fans of smaller triangles), including the labeling of the vertices of all the triangles.

   (c) Take the continued fraction $\frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n}$ you got in part (a) and reverse the order of the numbers $a_i$ to get a new continued fraction $\frac{1}{a_n} + \frac{1}{a_{n-1}} + \cdots + \frac{1}{a_1}$. Compute the value $\frac{p}{q}$ of this continued fraction, and also compute the strip of triangles for this fraction $\frac{p}{q}$.

3. Let $\frac{p_n}{q_n}$ be the value of the continued fraction $\frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n}$ where each of the $n$ terms $a_i$ is equal to 2. For example, $\frac{p_1}{q_1} = \frac{1}{2}$ and $\frac{p_2}{q_2} = \frac{1}{2} + \frac{1}{2} = \frac{2}{5}$.

   (a) Find equations expressing $p_n$ and $q_n$ in terms of $p_{n-1}$ and $q_{n-1}$, and use these to write down the values of $\frac{p_n}{q_n}$ for $n = 1, 2, 3, 4, 5, 6, 7$.

   (b) Compute the strip of triangles for $\frac{p_7}{q_7}$.

4. (a) A rectangle whose sides have lengths 13 and 48 can be partitioned into squares in the following way:

Determine the lengths of the sides of all the squares, and relate the numbers of squares of each size to the continued fraction for $\frac{13}{48}$.  

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**Chapter 2 Continued Fractions**
(b) Draw the analogous figure decomposing a rectangle of sides 19 and 42 into squares, and relate this to the continued fraction for 19/42.

5. This exercise is intended to illustrate the proof of Theorem 2.1 in the concrete case of the continued fraction $\frac{2}{3} + \frac{1}{4} + \frac{1}{5}$.

(a) Write down the product $A_1 A_2 A_3 A_4 = \left( \begin{array}{cc} 0 & 1 \\ 1 & a_1 \end{array} \right) \left( \begin{array}{cc} 0 & 1 \\ 1 & a_2 \end{array} \right) \left( \begin{array}{cc} 0 & 1 \\ 1 & a_3 \end{array} \right) \left( \begin{array}{cc} 0 & 1 \\ 1 & a_4 \end{array} \right)$ associated to $\frac{1}{2} + \frac{1}{3} + \frac{1}{4} + \frac{1}{5}$.

(b) Compute the four matrices $A_1, A_2, A_3, A_4$ and relate these to the edges of the zigzag path in the strip of triangles for $\frac{1}{2} + \frac{1}{3} + \frac{1}{4} + \frac{1}{5}$.

(c) Compute the four matrices $A_5, A_6, A_7, A_8$ and relate these to the successive fractions that one gets when one computes the value of $\frac{1}{2} + \frac{1}{3} + \frac{1}{4} + \frac{1}{5}$, namely $\frac{1}{5}, \frac{1}{4} + \frac{1}{5}, \frac{1}{3} + \frac{1}{4} + \frac{1}{5}$, and $\frac{1}{2} + \frac{1}{3} + \frac{1}{4} + \frac{1}{5}$.

6. (a) Find all integer solutions of the equations $40x + 89y = 1$ and $40x + 89y = 5$.

(b) Find another equation $ax + by = 1$ with integer coefficients $a$ and $b$ that has an integer solution in common with $40x + 89y = 1$. [Hint: use the Farey diagram.]

7. There is a close connection between the Diophantine equation $ax + by = n$ and the congruence $ax \equiv n \mod b$, where the symbol $\equiv$ means “is congruent to”. Namely, if one has a solution $(x, y)$ to $ax + by = n$ then $ax \equiv n \mod b$, and conversely, if one has a number $x$ such that $ax \equiv n \mod b$ then this means that $ax - n$ is a multiple of $b$, say $k$ times $b$, so $ax - n = kb$ or equivalently $ax - kb = n$ so one has a solution of $ax + by = n$ with $y = -k$.

Using this viewpoint, find all integers $x$ satisfying the congruence $31x \equiv 1 \mod 71$, and then do the same for the congruence $31x \equiv 10 \mod 71$. Are the solutions unique mod 71, i.e., unique up to adding multiples of 71?

8. Compute the values of the following infinite continued fractions:

(a) $\frac{1}{\sqrt{4}}$

(b) $\frac{1}{\sqrt{k}}$ for an arbitrary positive integer $k$.

(c) $\frac{1}{\sqrt{2}} + \frac{1}{\sqrt{3}}$ and $\frac{1}{\sqrt{1}} + \frac{1}{\sqrt{2}} + \frac{1}{\sqrt{3}}$

(d) $\frac{1}{\sqrt{1}} + \frac{1}{\sqrt{2}} + \frac{1}{\sqrt{3}} + \frac{1}{\sqrt{6}}$ and $\frac{1}{\sqrt{1}} + \frac{1}{\sqrt{4}} + \frac{1}{\sqrt{1}} + \frac{1}{\sqrt{2}} + \frac{1}{\sqrt{4}} + \frac{1}{\sqrt{1}} + \frac{1}{\sqrt{6}}$

(e) $\frac{1}{\sqrt{2}} + \frac{1}{\sqrt{3}} + \frac{1}{\sqrt{5}}$

9. Compute the continued fractions for $\sqrt{5}$ and $\sqrt{23}$.

10. Compute the continued fractions for $\sqrt{n^2 + 1}$ and $\sqrt{n^2 + n}$ where $n$ is an arbitrary positive integer.
Chapter 3. Linear Fractional Transformations

One thing one notices about the various versions of the Farey diagram is their symmetry. For the circular Farey diagram the symmetries are the reflections across the horizontal and vertical axes and the 180 degree rotation about the center. For the standard Farey diagram in the upper halfplane there are symmetries that translate the diagram by any integer distance to the left or the right, as well as reflections across certain vertical lines, the vertical lines through an integer or half-integer point on the $x$-axis. The Farey diagram could also be drawn to have 120 degree rotational symmetry and three reflectional symmetries.

Our purpose in this chapter is to study all possible symmetries of the Farey diagram, where we interpret the word “symmetry” in a broader sense than the familiar meaning from Euclidean geometry. For our purposes, symmetries will be invertible transformations that take vertices to vertices, edges to edges, and triangles to triangles. There are simple algebraic formulas for these more general symmetries, and these formulas lead to effective means of calculation. One of the applications will be to computing the values of periodic or eventually periodic continued fractions.

1. Symmetries of the Farey Diagram

From linear algebra one is familiar with the way in which $2 \times 2$ matrices \(
\begin{pmatrix} a & b \\ c & d \end{pmatrix}
\) correspond to linear transformations of the plane $\mathbb{R}^2$, transformations of the form

\[
T\begin{pmatrix} x \\ y \end{pmatrix} = \begin{pmatrix} a & b \\ c & d \end{pmatrix} \begin{pmatrix} x \\ y \end{pmatrix} = \begin{pmatrix} ax + by \\ cx + dy \end{pmatrix}
\]

In our situation we are going to restrict $a, b, c, d, x, y$ to be integers. Then by associating to a pair $(x, y)$ the fraction $x/y$ one obtains a closely related transformation

\[
T\left(\frac{x}{y}\right) = \frac{ax + by}{cx + dy} = \frac{a\left(\frac{x}{y}\right) + b}{c\left(\frac{x}{y}\right) + d}
\]

If we set $z = x/y$ then $T$ can also be written in the form

\[
T(z) = \frac{az + b}{cz + d}
\]

Such a transformation is called a linear fractional transformation since it is defined by a fraction whose numerator and denominator are linear functions.

In the formula $T(x/y) = (ax + by)/(cx + dy)$ there is no problem with allowing $x/y = \pm 1/0$ just by setting $(x, y) = (\pm 1, 0)$, and the result is that $T(\pm 1/0) = a/c$. 

The value \( T(x/y) = (ax + by)/(cx + dy) \) can also be \( \pm 1/0 \), when \((x, y) = (d, -c)\) and the matrix has determinant \( ad - bc = \pm 1 \). This means that \( T \) defines a function from vertices of the Farey diagram to vertices of the Farey diagram. We would like \( T \) to take edges of the diagram to edges of the diagram, and the following result gives a condition for this to happen.

**Proposition 3.1.** If the matrix \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) has determinant \( \pm 1 \) then the associated linear fractional transformation \( T \) takes each pair of vertices in the Farey diagram that lie at the ends of an edge of the diagram to another such pair of vertices.

**Proof:** We showed in Theorem 2.2 that two vertices labeled \( p/q \) and \( r/s \) are joined by an edge in the diagram exactly when \( ps - qr = \pm 1 \), or in other words when the matrix \( \begin{pmatrix} p & r \\ q & s \end{pmatrix} \) has determinant \( \pm 1 \). The two columns of the product matrix \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \begin{pmatrix} p & r \\ q & s \end{pmatrix} \) correspond to the two vertices \( T(p/q) \) and \( T(r/s) \), by the definition of matrix multiplication:

\[
\begin{pmatrix} a & b \\ c & d \end{pmatrix} \begin{pmatrix} p & r \\ q & s \end{pmatrix} = \begin{pmatrix} ap + bq & ar + bs \\ cp + dq & cr + ds \end{pmatrix}
\]

The proposition can then be restated as saying that if \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) and \( \begin{pmatrix} p & r \\ q & s \end{pmatrix} \) each have determinant \( \pm 1 \) then so does their product \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \begin{pmatrix} p & r \\ q & s \end{pmatrix} \). But it is a general fact about determinants that the determinant of a product is the product of the determinants. (This is easy to prove by a direct calculation in the case of \( 2 \times 2 \) matrices.) So the product of two matrices of determinant \( \pm 1 \) has determinant \( \pm 1 \).

As notation, we will use \( LF(\mathbb{Z}) \) to denote the set of all linear fractional transformations \( T(x/y) = (ax + by)/(cx + dy) \) with coefficients \( a, b, c, d \) in \( \mathbb{Z} \) such that the matrix \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) has determinant \( \pm 1 \). (Here \( \mathbb{Z} \) is the set of all integers.)

Changing the matrix \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) to its negative \( \begin{pmatrix} -a & -b \\ -c & -d \end{pmatrix} \) produces the same linear fractional transformation since \((-ax - by)/(cx + dy) = (ax + by)/(cx + dy)\). This is in fact the only way that different matrices can give the same linear fractional transformation \( T \), as we will see later in this chapter. Note that changing \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) to its negative \( \begin{pmatrix} -a & -b \\ -c & -d \end{pmatrix} \) does not change the determinant. Thus each linear fractional transformation in \( LF(\mathbb{Z}) \) has a well-defined determinant, either \( +1 \) or \( -1 \). Later in this chapter we will also see how the distinction between determinant \( +1 \) and determinant \( -1 \) has a geometric interpretation in terms of orientations.

A useful fact about \( LF(\mathbb{Z}) \) is that each transformation \( T \) in \( LF(\mathbb{Z}) \) has an inverse \( T^{-1} \) in \( LF(\mathbb{Z}) \) because the inverse of a \( 2 \times 2 \) matrix is given by the formula

\[
\begin{pmatrix} a & b \\ c & d \end{pmatrix}^{-1} = \frac{1}{ad - bc} \begin{pmatrix} d & -b \\ -c & a \end{pmatrix}
\]

Thus if \( a, b, c, d \) are integers with \( ad - bc = \pm 1 \) then the inverse matrix also has integer entries and determinant \( \pm 1 \). The factor \( \frac{1}{ad - bc} \) is \( \pm 1 \) so it can be ignored since
the matrices \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) and \( -\begin{pmatrix} a & b \\ c & d \end{pmatrix} \) determine the same linear fractional transformation, as we observed in the preceding paragraph.

The preceding proposition says that each linear fractional transformation \( T \) in \( \text{LF}(\mathbb{Z}) \) not only sends vertices of the Farey diagram to vertices, but also edges to edges. It follows that \( T \) must take triangles in the diagram to triangles in the diagram, since triangles correspond to sets of three vertices, each pair of which forms the endpoints of an edge. Since each transformation \( T \) in \( \text{LF}(\mathbb{Z}) \) has an inverse in \( \text{LF}(\mathbb{Z}) \), this implies that \( T \) gives a one-to-one (injective) and onto (surjective) transformation of vertices, and also of edges and triangles. For example, if two edges \( e_1 \) and \( e_2 \) have the same image \( T(e_1) = T(e_2) \) then we must have \( T^{-1}(T(e_1)) = T^{-1}(T(e_2)) \) or in other words \( e_1 = e_2 \), so \( T \) cannot send two different edges to the same edge, which means it is one-to-one on edges. Also, every edge \( e_1 \) is the image \( T(e_2) \) of some edge \( e_2 \) since we can write \( e_1 = T(T^{-1}(e_1)) \) and let \( e_2 = T^{-1}(e_1) \). The same reasoning works with vertices and triangles as well as edges.

A useful property of linear fractional transformations that we will use repeatedly is that the way an element of \( \text{LF}(\mathbb{Z}) \) acts on the Farey diagram is uniquely determined by where a single triangle is sent. This is because once one knows where one triangle goes, this uniquely determines where the three adjacent triangles go, and this in turn determines where the six new triangles adjacent to these three go, and so on.

We will now give examples illustrating seven different ways that elements of \( \text{LF}(\mathbb{Z}) \) can act on the Farey diagram.

1. The transformation \( T(x/y) = y/x \) with matrix \( \begin{pmatrix} 0 & 1 \\ 1 & 0 \end{pmatrix} \) gives a reflection of the circular Farey diagram across its vertical axis of symmetry. This is a reflection across a line perpendicular to an edge of the diagram.

2. The reflection across the horizontal axis of symmetry is the element \( T(x/y) = -x/y \) with matrix \( \begin{pmatrix} -1 & 0 \\ 0 & 1 \end{pmatrix} \). This is a reflection across an edge of the diagram.

3. If we compose the two preceding reflections we get the transformation \( T(x/y) = -y/x \) with matrix \( \begin{pmatrix} 0 & 1 \\ -1 & 0 \end{pmatrix} \). This rotates the Farey diagram 180 degrees about its center, interchanging \( 1/0 \) and \( 0/1 \) and also interchanging \( 1/1 \) and \( -1/1 \). Thus it rotates the diagram 180 degrees about the centerpoint of an edge.

4. Consider \( T(x/y) = y/(y - x) \) corresponding to the matrix \( \begin{pmatrix} 0 & 1 \\ -1 & 1 \end{pmatrix} \). This has the effect of “rotating” the triangle \( \langle 1/0, 0/1, 1/1 \rangle \) about its centerpoint, taking \( 1/0 \) to
0/1, 0/1 to 1/1 and 1/1 back to 1/0. The whole Farey diagram is then “rotated” about the same point.

(5) Next let $T(x/y) = x/(x + y)$, corresponding to the matrix $\begin{pmatrix} 1 & 0 \\ 1 & 1 \end{pmatrix}$. In particular $T(0/1) = 0/1$, so $0/1$ is a fixed point of $T$, a point satisfying $T(z) = z$. Also we have $T(1/0) = 1/1$ and more generally $T(1/n) = 1/(n + 1)$. Thus the triangle $\langle 0/1, 1/0, 1/1 \rangle$ is taken to the triangle $\langle 0/1, 1/1, 1/2 \rangle$. This implies that $T$ is a “rotation” of the Farey diagram about the vertex $0/1$, taking each triangle with $0/1$ as a vertex to the next triangle in the clockwise direction about this vertex.

(6) A different sort of behavior is exhibited by $T(x/y) = (2x + y)/(x + y)$ corresponding to $\begin{pmatrix} 2 & 1 \\ 1 & 1 \end{pmatrix}$. To visualize $T$ as a transformation of the Farey diagram let us look at the infinite strip

![Infinite Farey Strip with Transformation](image)

We claim that $T$ translates the whole strip one unit to the right. To see this, notice first that since $T$ takes $1/0$ to $2/1$, $0/1$ to $1/1$, and $1/1$ to $3/2$, it takes the triangle $\langle 1/0, 0/1, 1/1 \rangle$ to the triangle $\langle 2/1, 1/1, 3/2 \rangle$. This implies that $T$ takes the triangle just to the right of $\langle 1/0, 0/1, 1/1 \rangle$ to the triangle just to the right of $\langle 2/1, 1/1, 3/2 \rangle$, and similarly each successive triangle is translated one unit to the right. The same argument shows that each successive triangle to the left of the original one is also translated one unit to the right. Thus the whole strip is translated one unit to the right.

(7) Using the same figure as in the preceding example, consider the transformation $T(x/y) = (x + y)/x$ corresponding to the matrix $\begin{pmatrix} 1 & 1 \\ 1 & 0 \end{pmatrix}$. This sends the triangle $\langle 1/0, 0/1, 1/1 \rangle$ to $\langle 1/1, 1/0, 2/1 \rangle$ which is the next triangle to the right in the infinite strip. Geometrically, $T$ translates the first triangle half a unit to the right and reflects it across the horizontal axis of the strip. It follows that the whole strip is translated half a unit to the right and reflected across the horizontal axis. Such a motion is referred to as a glide-reflection. Notice that performing this motion twice in succession yields a translation of the strip one unit to the right, the transformation in the preceding example.

Thus we have seven types of symmetries of the Farey diagram: reflections across an edge or a line perpendicular to an edge; rotations about the centerpoint of an edge or a triangle, or about a vertex; and translations and glide-reflections of periodic infinite strips. (Not all periodic strips have glide-reflection symmetries.) It is a true fact, though we won’t prove it here, that every element of $LF(\mathbb{Z})$ acts on the Farey
diagram in one of these seven ways, except for the identity transformation \( T(x/y) = x/y \) of course.

2. Specifying Where a Triangle Goes

As we observed earlier, the action of an element of \( LF(Z) \) on the Farey diagram is completely determined by where it sends a single triangle. Now we will see that there always exists an element of \( LF(Z) \) sending any triangle to any other triangle, and in fact, one can do this specifying where each individual vertex of the triangle goes.

As an example, suppose we wish to find an element \( T \) of \( LF(Z) \) that takes the triangle \( (2/5, 1/3, 3/8) \) to the triangle \( (5/8, 7/11, 2/3) \), preserving the indicated ordering of the vertices, so \( T(2/5) = 5/8 \), \( T(1/3) = 7/11 \), and \( T(3/8) = 2/3 \). For this problem to even make sense we might want to check first that these really are triangles in the Farey diagram. In the first case, \( (2/5, 1/3) \) is an edge since the matrix \( \begin{pmatrix} 2 & 1 \\ 5 & 3 \end{pmatrix} \) has determinant 1, and there is a triangle joining this edge to \( 3/8 \) since \( 3/8 \) is the mediant of \( 2/5 \) and \( 1/3 \). For the other triangle, the determinant of \( \begin{pmatrix} 5 & 2 \\ 8 & 3 \end{pmatrix} \) is \(-1\) and the mediant of \( 5/8 \) and \( 2/3 \) is \( 7/11 \).

As a first step toward constructing the desired transformation \( T \) we will do something slightly weaker: We construct a transformation \( T \) taking the edge \( (2/5, 1/3) \) to the edge \( (5/8, 7/11) \). This is rather easy if we first notice the general fact that the transformation \( T(x/y) = (ax + by)/(cx + dy) \) with matrix \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) takes \( 1/0 \) to \( a/c \) and \( 0/1 \) to \( b/d \). Thus the transformation \( T_1 \) with matrix \( \begin{pmatrix} 2 & 1 \\ 5 & 3 \end{pmatrix} \) takes \( 1/0, 0/1 \) to \( (2/5, 1/3) \), and the transformation \( T_2 \) with matrix \( \begin{pmatrix} 5 & 7 \\ 8 & 11 \end{pmatrix} \) takes \( 1/0, 0/1 \) to \( (5/8, 7/11) \). Then the product

\[
T_2 T_1^{-1} = \begin{pmatrix} 5 & 7 \\ 8 & 11 \end{pmatrix} \begin{pmatrix} 2 & 1 \\ 5 & 3 \end{pmatrix}^{-1}
\]

takes \( (2/5, 1/3) \) first to \( (1/0, 0/1) \) and then to \( (5/8, 7/11) \). Doing the calculation, we get

\[
\begin{pmatrix} 5 & 7 \\ 8 & 11 \end{pmatrix} \begin{pmatrix} 2 & 1 \\ 5 & 3 \end{pmatrix}^{-1} = \begin{pmatrix} 5 & 7 \\ 8 & 11 \end{pmatrix} \begin{pmatrix} 3 & -1 \\ -5 & 2 \end{pmatrix} = \begin{pmatrix} -20 & 9 \\ -31 & 14 \end{pmatrix}
\]

This takes the edge \( (2/5, 1/3) \) to the edge \( (5/8, 7/11) \), but does it do the right thing on the third vertex of the triangle \( (2/5, 1/3, 3/8) \), taking it to the third vertex of \( (5/8, 7/11, 2/3) \)? This is not automatic since there are always two triangles containing a given edge, and in this case the other triangle having \( (5/8, 7/11) \) as an edge is \( (5/8, 7/11, 12/19) \) since \( 12/19 \) is the mediant of \( 5/8 \) and \( 7/11 \). In fact, if we compute what our \( T \) does to \( 3/8 \) we get

\[
\begin{pmatrix} -20 & 9 \\ -31 & 14 \end{pmatrix} \begin{pmatrix} 3 \\ 8 \end{pmatrix} = \begin{pmatrix} 12 \\ 19 \end{pmatrix}
\]

so we don’t have the right \( T \) yet. To fix the problem, notice that we have a little flexibility in the choice of a matrix \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \) taking \( 1/0 \) to \( a/c \) and \( 0/1 \) to \( b/d \) since
we can multiply either column by \(-1\) without affecting the fractions \(a/b\) and \(c/d\). It doesn’t matter which column we multiply by \(-1\) since multiplying both columns by \(-1\) multiplies the whole matrix by \(-1\) which doesn’t change the associated element of \(LF(\mathbb{Z})\), as noted earlier. In the case at hand, suppose we change the sign of the first column of \(\begin{pmatrix} 5 & 7 \\ 8 & 11 \end{pmatrix}\). Then we get

\[
\begin{pmatrix} -5 & 7 \\ -8 & 11 \end{pmatrix} \begin{pmatrix} 2 & 1 \\ 5 & 3 \end{pmatrix}^{-1} = \begin{pmatrix} -5 & 7 \\ -8 & 11 \end{pmatrix} \begin{pmatrix} 3 & -1 \\ -5 & 2 \end{pmatrix} = \begin{pmatrix} -50 & 19 \\ -79 & 30 \end{pmatrix}
\]

This fixes the problem since

\[
\begin{pmatrix} -50 & 19 \\ -79 & 30 \end{pmatrix} \begin{pmatrix} 3 \\ 8 \end{pmatrix} = \begin{pmatrix} 2 \\ 3 \end{pmatrix}
\]

Here is a general statement summarizing what we saw in this one example:

**Proposition 3.2.** (a) For any two triangles \(\langle p/q, r/s, t/u \rangle\) and \(\langle p'/q', r'/s', t'/u' \rangle\) in the Farey diagram there is a unique element \(T\) in \(LF(\mathbb{Z})\) taking the first triangle to the second triangle preserving the ordering of the vertices, so \(T(p/q) = p'/q'\), \(T(r/s) = r'/s'\), and \(T(t/u) = t'/u'\).

(b) The matrix \(\begin{pmatrix} a & b \\ c & d \end{pmatrix}\) representing a given transformation \(T\) in \(LF(\mathbb{Z})\) is unique except for replacing it by \(\begin{pmatrix} -a & -b \\ -c & -d \end{pmatrix}\).

**Proof:** As we saw in the example above, there is a composition \(T_2T_1^{-1}\) taking the edge \(\langle p/q, r/s \rangle\) to \(\langle p'/q', r'/s' \rangle\), where \(T_1\) has matrix \(\begin{pmatrix} p & r \\ q & s \end{pmatrix}\) and \(T_2\) has matrix \(\begin{pmatrix} p' & r' \\ q' & s' \end{pmatrix}\).

If this composition \(T_2T_1^{-1}\) does not take \(t/u\) to \(t'/u'\) we modify \(T_2\) by changing the sign of one of its columns, say the first column. Thus we change \(\begin{pmatrix} p' & r' \\ q' & s' \end{pmatrix}\) to \(\begin{pmatrix} -p' & r' \\ -q' & s' \end{pmatrix}\), which equals the product \(\begin{pmatrix} p' & r' \\ q' & s' \end{pmatrix}\left(\begin{pmatrix} -1 & 0 \\ 0 & 1 \end{pmatrix}\right)\). The matrix \(\begin{pmatrix} -1 & 0 \\ 0 & 1 \end{pmatrix}\) corresponds to the transformation \(R(x/y) = -x/y\) reflecting the Farey diagram across the edge \(\langle 1/0, 0/1 \rangle\). Thus we are replacing \(T_2T_1^{-1}\) by \(T_2RT_1^{-1}\), inserting a reflection that interchanges the two triangles containing the edge \(\langle 1/0, 0/1 \rangle\). By inserting \(R\) we change where the composition \(T_2T_1^{-1}\) sends the third vertex \(t/u\) of the triangle \(\langle p/q, r/s, t/u \rangle\), so we can guarantee that \(t/u\) is taken to \(t'/u'\). This proves part (a).

For part (b), note first that the transformation \(T\) determines the values \(T(1/0) = a/c\) and \(T(0/1) = b/d\). The fractions \(a/c\) and \(b/d\) are in lowest terms (because \(ad - bc = \pm 1\)) so this means that we know the two columns of the matrix \(\begin{pmatrix} a & b \\ c & d \end{pmatrix}\) up to multiplying either or both columns by \(-1\). We need to check that changing the sign of one column without changing the sign of the other column gives a different transformation. It doesn’t matter which column we change since \(\begin{pmatrix} a & b \\ -c & -d \end{pmatrix}\) = \(-\begin{pmatrix} a & b \\ c & d \end{pmatrix}\).

As we saw in part (a), changing the sign in the first column amounts to replacing \(T\) by the composition \(TR\), but this is a different transformation from \(T\) since it has a different effect on the triangles containing the edge \(\langle 1/0, 0/1 \rangle\). \(\square\)
3. Continued Fractions Again

Linear fractional transformations can be used to compute the values of periodic or eventually periodic continued fractions, and to see that these values are always quadratic irrational numbers. To illustrate this, consider the periodic continued fraction

$$\frac{1}{2} + \frac{1}{3} + \frac{1}{1} + \frac{1}{4}$$

The associated periodic strip in the Farey diagram is the following:

We would like to compute the element \( T \) of \( LF(\mathbb{Z}) \) that gives the rightward translation of this strip that exhibits the periodicity. A first guess is the \( T \) with matrix \( \begin{pmatrix} 4 & 19 \\ 9 & 43 \end{pmatrix} \) since this sends \( \langle 1/0, 0/1 \rangle \) to \( \langle 4/9, 19/43 \rangle \). This is actually the correct \( T \) since it sends the vertex 1/1 just to the right of 1/0, which is the mediant of 1/0 and 0/1, to the vertex \((4 + 19)/(9 + 43)\) just to the right of 4/9, which is the mediant of 4/9 and 19/43. This is a general fact since \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \begin{pmatrix} 1 \\ 1 \end{pmatrix} = \begin{pmatrix} a+b \\ c+d \end{pmatrix} \).

The fractions labeling the vertices along the zigzag path in the strip moving toward the right are the convergents to \(\frac{1}{2} + \frac{1}{3} + \frac{1}{1} + \frac{1}{4} \). Call these convergents \( z_1, z_2, \cdots \) and their limit \( z \). When we apply the translation \( T \) we are taking each convergent to a later convergent in the sequence, so both the sequence \( \{z_n\} \) and the sequence \( \{T(z_n)\} \) converge to \( z \). On the other hand the sequence \( \{T(z_n)\} \) converges to \( T(z) \) since this is just saying that \( \frac{4z_n+19}{9z_n+43} \) converges to \( \frac{4z+19}{9z+43} \) as \( z_n \) converges to \( z \). Thus we have \( T(z) = z \).

In summary, what we have just argued is that the value \( z \) of the periodic continued fraction \( \frac{1}{2} + \frac{1}{3} + \frac{1}{1} + \frac{1}{4} \) satisfies the equation \( T(z) = z \), or in other words, \( \frac{4z+19}{9z+43} = z \). This can be rewritten as \( 4z + 19 = 9z^2 + 43z \), which simplifies to \( 9z^2 + 39z - 19 = 0 \). Computing the roots of this quadratic equation, we get

\[
z = \frac{-39 \pm \sqrt{39^2 + 4 \cdot 9 \cdot 19}}{18} = \frac{-39 \pm 3\sqrt{13^2 + 4 \cdot 19}}{18} = \frac{-13 \pm \sqrt{245}}{6} = \frac{-13 \pm 7\sqrt{5}}{6}
\]

The positive root is the one that the right half of the infinite strip converges to, so we have

\[
-\frac{13 + 7\sqrt{5}}{6} = \frac{1}{2} + \frac{1}{3} + \frac{1}{1} + \frac{1}{4}
\]

Incidentally, the other root \((-13 - 7\sqrt{5})/6\) has an interpretation in terms of the diagram as well: It is the limit of the numbers labeling the vertices of the zigzag path moving off to the left rather than to the right. This follows by the same sort of argument as above.
If a periodic continued fraction has period of odd length, the transformation giving the periodicity is a glide-reflection of the periodic strip rather than a translation. As an example, consider

\[
\frac{1}{1} + \frac{1}{2} + \frac{1}{3}
\]

Here the periodic strip is

\[
\begin{array}{ccc}
\frac{1}{0} & \frac{1}{1} & \frac{7}{10} \\
0 & \frac{2}{3} & \\
\end{array}
\]

The transformation \( T \) with matrix \( \begin{pmatrix} 2 & 7 \\ 3 & 10 \end{pmatrix} \) takes \( (1/0, 0/1) \) to \( (2/3, 7/10) \) and the mediant \( 1/1 \) of \( 1/0 \) and \( 0/1 \) to the mediant \( 9/13 \) of \( 2/3 \) and \( 7/10 \) so this transformation is a glide-reflection of the strip. The equation \( T(z) = z \) becomes \( 2z^2 + 7z - 10 = z \) which simplifies to \( 3z^2 + 8z - 7 = 0 \) with roots \( (-4 \pm \sqrt{37})/3 \). The positive root gives

\[
\frac{-4 + \sqrt{37}}{3} = \frac{1}{1} + \frac{1}{2} + \frac{1}{3}
\]

Continued fractions that are only eventually periodic can be treated in a similar fashion. For example, consider

\[
\frac{1}{2} + \frac{1}{2} + \frac{1}{1} + \frac{1}{2} + \frac{1}{3}
\]

The corresponding infinite strip is

\[
\begin{array}{cccc}
\frac{1}{0} & \frac{1}{2} & \frac{3}{7} & \frac{27}{64} \\
0 & \frac{2}{5} & \frac{8}{19} & \\
\end{array}
\]

In this case if we discard the triangles corresponding to the initial nonperiodic part of the continued fraction, \( 1/2 + 1/2 \), and then extend the remaining periodic part in both directions, we obtain a periodic strip that is carried to itself by the glide-reflection \( T \) taking \( (1/2, 2/5) \) to \( (8/19, 27/64) \):
We can compute $T$ as the composition $\langle 1/2, 2/5 \rangle \rightarrow \langle 1/0, 0/1 \rangle \rightarrow \langle 8/19, 27/64 \rangle$ corresponding to the product

$$
\begin{pmatrix}
8 & 27 \\
19 & 64
\end{pmatrix}
\begin{pmatrix}
1 & 2 \\
2 & 5
\end{pmatrix}
^{-1}
= \begin{pmatrix}
8 & 27 \\
19 & 64
\end{pmatrix}
\begin{pmatrix}
5 & -2 \\
-2 & 1
\end{pmatrix}
= \begin{pmatrix}
-14 & 11 \\
-33 & 26
\end{pmatrix}
$$

Since this transformation takes $3/7$ to the mediant $(8+27)/(19+64)$, it is the glide-reflection we want. Now we solve $T(z) = z$. This means $\frac{-14z+11}{-33z+26} = z$, which reduces to the equation $33z^2 - 40z + 11 = 0$ with roots $z = (20 \pm \sqrt{37})/33$. Both roots are positive, and we want the smaller one, $(20 - \sqrt{37})/33$, because along the top edge of the strip the numbers decrease as we move to the right, approaching the smaller root, and they increase as we move to the left, approaching the larger root. Thus we have

$$
(20 - \sqrt{37})/33 = \frac{1}{2} + \frac{1}{2} + \frac{1}{1} + \frac{1}{2} + \frac{1}{3}
$$

Notice that $\sqrt{37}$ occurs in both this example and the preceding one where we computed the value of $\frac{1}{1} + \frac{1}{2} + \frac{1}{3}$. This is not just an accident. It had to happen because to get from $\frac{1}{1} + \frac{1}{2} + \frac{1}{3}$ to $\frac{1}{2} + \frac{1}{2} + \frac{1}{1} + \frac{1}{2} + \frac{1}{3}$ one adds 2 and inverts, then adds 2 and inverts again, and each of these operations of adding an integer or taking the reciprocal takes place within the set $\mathbb{Q}(\sqrt{37})$ of all numbers of the form $a + b\sqrt{37}$ with $a$ and $b$ rational. More generally, this argument shows that any eventually periodic continued fraction whose periodic part is $\frac{1}{1} + \frac{1}{2} + \frac{1}{3}$ has as its value some number in $\mathbb{Q}(\sqrt{37})$. However, not all irrational numbers in $\mathbb{Q}(\sqrt{37})$ have eventually periodic continued fractions with periodic part $\frac{1}{1} + \frac{1}{2} + \frac{1}{3}$. For example, the continued fraction for $\sqrt{37}$ itself is $6 + \frac{1}{12}$, with a different periodic part. (Check this by computing the value of this continued fraction.)

The procedure we have used in these examples works in general for any irrational number $z$ whose continued fraction is eventually periodic. From the periodic part of the continued fraction one constructs a periodic infinite strip in the Farey diagram, where the periodicity is given by a linear fractional transformation $T(z) = \frac{az+b}{cz+d}$ with integer coefficients, with $T$ either a translation or a glide-reflection of the strip. As we argued in the first example, the number $z$ satisfies the equation $T(z) = z$. This becomes the quadratic equation $az + b = cz^2 + dz$ with integer coefficients, or in more standard form, $cz^2 + (d-a)z - b = 0$. By the quadratic formula, the roots of this equation have the form $A + B\sqrt{n}$ for some rational numbers $A$ and $B$ and some integer $n$. We know that the real number $z$ is a root of the equation so $n$ can’t be negative, and it can’t be a square since $z$ is irrational.

Thus we have an argument that proves one half of Lagrange’s Theorem:

**Proposition 3.3.** A number whose continued fraction is periodic or eventually periodic is a quadratic irrational.

**Proof:** The main part of the argument was given above, but there is one technical point that needs be addressed. Could the leading coefficient $c$ in the quadratic equation
$cz^2 + (d - a)z - b = 0$ be zero? If this were the case then we couldn’t apply the quadratic formula to solve for $z$, so we need to show that $c$ cannot be zero. We do this in the following way. If $c$ were zero the equation would become the linear equation $(d - a)z = b$. If the coefficient of $z$ in this equation is nonzero, we have only one root, $z = b/(d - a)$, a rational number contrary to the fact that $z$ is irrational since its continued fraction is infinite. Thus we are left with the possibility that $c = 0$ and $a = d$, so the equation for $z$ reduces to the equation $b = 0$. Then the transformation $T$ would have the form $T(z) = az/a = z$ so it would be the identity transformation. However we know it is a genuine translation or a glide-reflection, so it is not the identity. We conclude from all this that $c$ cannot be zero, and the technical point is taken care of.

4. Orientations

Elements of $LF(\mathbb{Z})$ are represented by integer matrices $\begin{pmatrix} a & b \\ c & d \end{pmatrix}$ of determinant $\pm 1$. The distinction between determinant $+1$ and $-1$ has a very nice geometric interpretation in terms of orientations, which can be described in terms of triangles. A triangle in the Farey diagram can be oriented by choosing either the clockwise or counterclockwise ordering of its three vertices. An element $T$ of $LF(\mathbb{Z})$ takes each triangle to another triangle in a way that either preserves the two possible orientations or reverses them.

For example, among the seven types of transformations we looked at earlier, only reflections and glide-reflections reverse the orientations of triangles. Note that if a transformation $T$ preserves the orientation of one triangle, it has to preserve the orientation of the three adjacent triangles, and then of the triangles adjacent to these, and so on for all the triangles. Similarly, if the orientation of one triangle is reversed by $T$, then the orientations of all triangles are reversed.

**Proposition 3.4.** A transformation $T(x/y) = (ax + by)/(cx + dy)$ in $LF(\mathbb{Z})$ preserves orientations of triangles in the Farey diagram when the determinant $ad - bc$ is $+1$ and reverses the orientations when the determinant is $-1$.

**Proof:** We will first prove a special case and then deduce the general case from the special case. The special case is that $a, b, c, d$ are all positive or zero. The transformation $T$ with matrix $\begin{pmatrix} a & b \\ c & d \end{pmatrix}$ takes the edge $\langle 1/0, 0/1 \rangle$ in the circular Farey diagram to the edge $\langle a/c, b/d \rangle$, and if $a, b, c, d$ are all positive or zero, this edge lies in the upper half
of the diagram. Since \( T(1/1) = (a+b)/(c+d) \), the triangle \( \langle 1/0, 0/1, 1/1 \rangle \) is taken to the triangle \( \langle a/c, b/d, (a+b)/(c+d) \rangle \) whose third vertex \( (a+b)/(c+d) \) lies above the edge \( \langle a/c, b/d \rangle \), by the way the Farey diagram was constructed using mediants, since we assume \( a, b, c, d \) are positive or zero. We know that the edge \( \langle a/c, b/d \rangle \) is oriented to the right if \( ad - bc = +1 \) and to the left if \( ad - bc = -1 \). This means that \( T \) preserves the orientation of the triangle \( \langle 1/0, 0/1, 1/1 \rangle \) if the determinant is \(+1\) and reverses the orientation if the determinant is \(-1\).

This proves the special case.

The general case can be broken into two subcases, according to whether the edge \( \langle a/c, b/d \rangle \) lies in the upper or the lower half of the diagram. If \( \langle a/c, b/d \rangle \) lies in the upper half of the diagram, then after multiplying one or both columns of \( \binom{a}{b} \) by \(-1\) if necessary, we will be in the special case already considered. Multiplying both columns by \(-1\) doesn’t affect \( T \). Multiplying one column by \(-1\) corresponds to first reflecting across the edge \( \langle 1/0, 0/1 \rangle \), as we have seen earlier. Modifying \( T \) in this way changes the sign of the determinant and it also changes whether \( T \) preserves or reverses orientation, so the special case already proved implies the case that \( T \) takes \( \langle 1/0, 0/1 \rangle \) to an edge in the upper half of the diagram.

The remaining possibility is that \( T \) takes the edge \( \langle 1/0, 0/1 \rangle \) to an edge in the lower half of the diagram. In this case if we follow \( T \) by reflection across the edge \( \langle 1/0, 0/1 \rangle \) we get a new transformation taking \( \langle 1/0, 0/1 \rangle \) to an edge in the upper half of the diagram. As before, composing with this reflection changes \( T \) from orientation-preserving to orientation-reversing and vice versa, and it also changes the sign of the determinant since the matrix \( \binom{a}{b} \) is changed to \( \binom{-a}{-b} \), so this case follows from the previous case.\[\Box\]

As we noted in Chapter 2, to determine whether a matrix representing an element of \( LF(\mathbb{Z}) \) has determinant \(+1\) or \(-1\) it suffices to compute just the last digit of the determinant, and this can be done using just the last digit of the entries in the matrix. This is easy to do in one’s head even if the entries in the matrix have many digits.

We will let \( LF^+(\mathbb{Z}) \) denote the elements of \( LF(\mathbb{Z}) \) corresponding to matrices of determinant \(+1\).
Proposition 3.5. For any two edges \( \langle p/q, r/s \rangle \) and \( \langle p'/q', r'/s' \rangle \) of the Farey diagram there exists a unique element \( T \in LF^+(\mathbb{Z}) \) taking the first edge to the second edge preserving the ordering of the vertices, so \( T(p/q) = p'/q' \) and \( T(r/s) = r'/s' \).

Proof: We already know that there exists an element \( T \) in \( LF(\mathbb{Z}) \) with \( T(p/q) = p'/q' \) and \( T(r/s) = r'/s' \), and in fact there are exactly two choices for \( T \) which are distinguished by which of the two triangles containing \( \langle p'/q', r'/s' \rangle \) a triangle containing \( \langle p/q, r/s \rangle \) is sent to. One of these choices will make \( T \) preserve orientation and the other will make \( T \) reverse orientation. So there is only one choice where the determinant is \(+1\). \( \square \)

Exercises

1. Find a formula for the linear fractional transformation that rotates the triangle \( \langle 0/1, 1/2, 1/1 \rangle \) to \( \langle 1/1, 0/1, 1/2 \rangle \).

2. Find the linear fractional transformation that reflects the Farey diagram across the edge \( \langle 1/2, 1/3 \rangle \) (so in particular, the transformation takes \( 1/2 \) to \( 1/2 \) and \( 1/3 \) to \( 1/3 \)).

3. Find a formula for the linear fractional transformation that reflects the upper half-plane version of the Farey diagram across the vertical line \( x = 3/2 \).

4. Find an infinite periodic strip of triangles in the Farey diagram such that the transformation \( \begin{pmatrix} 0 & 1 \\ 1 & 2 \end{pmatrix} \) is a glide-reflection along this strip and the transformation \( \begin{pmatrix} 0 & 1 \\ 1 & 2 \end{pmatrix} \begin{pmatrix} 0 & 1 \\ 1 & 2 \end{pmatrix} = \begin{pmatrix} 1 & 2 \\ 2 & 5 \end{pmatrix} \) is a translation along this strip.

5. Let \( T \) be an element of \( LF(\mathbb{Z}) \) with matrix \( \begin{pmatrix} a & b \\ c & d \end{pmatrix} \). Show that the composition \( T \begin{pmatrix} -1 & 0 \\ 0 & 1 \end{pmatrix} T^{-1} \) is the reflection across the edge \( \langle a/c, b/d \rangle = T(\langle 1/0, 0/1 \rangle) \).

For each of the remaining six problems, compute the value of the given periodic or eventually periodic continued fraction by first drawing the associated infinite strip of triangles, then finding a linear fractional transformation \( T \) in \( LF(\mathbb{Z}) \) that gives the periodicity in the strip, then solving \( T(z) = z \).

6. \( \frac{1}{2} + \frac{1}{5} \)

7. \( \frac{1}{2} + \frac{1}{4} + \frac{1}{1} \)

8. \( \frac{1}{1} + \frac{1}{4} + \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{2} \)

9. \( 2 + \frac{1}{1} + \frac{1}{4} + \frac{1}{1} \)

10. \( 2 + \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{4} \)

11. \( \frac{1}{1} + \frac{1}{1} + \frac{1}{2} + \frac{1}{3} \)
Chapter 4. Quadratic Forms

Finding Pythagorean triples is answering the question, *When is the sum of two squares equal to a square?* More generally one can ask, *Exactly which numbers are sums of two squares?* In other words, when does an equation \( x^2 + y^2 = n \) have integer solutions, and how can one find these solutions? The brute force approach of simply plugging in values for \( x \) and \( y \) leads to the following list of all solutions for \( n \leq 50 \) (apart from interchanging \( x \) and \( y \)):

\[
1 = 1^2 + 0^2, \quad 2 = 1^2 + 1^2, \quad 4 = 2^2 + 0^2, \quad 5 = 2^2 + 1^2, \quad 8 = 2^2 + 2^2, \quad 9 = 3^2 + 0^2, \\
10 = 3^2 + 1^2, \quad 13 = 3^2 + 2^2, \quad 16 = 4^2 + 0^2, \quad 17 = 4^2 + 1^2, \quad 18 = 3^2 + 3^2, \\
20 = 4^2 + 2^2, \quad 25 = 5^2 + 0^2 = 4^2 + 3^2, \quad 26 = 5^2 + 1^2, \quad 29 = 5^2 + 2^2, \quad 32 = 4^2 + 4^2, \\
34 = 5^2 + 3^2, \quad 36 = 6^2 + 0^2, \quad 37 = 6^2 + 1^2, \quad 40 = 6^2 + 2^2, \quad 41 = 5^2 + 4^2, \\
45 = 6^2 + 3^2, \quad 49 = 7^2 + 0^2, \quad 50 = 5^2 + 5^2 = 7^2 + 1^2
\]

Notice that in some cases there is more than one solution for a given value of \( n \). Our first goal will be to describe a more efficient way to find the integer solutions of \( x^2 + y^2 = n \) and to display them graphically in a way that sheds much light on their structure. The technique for doing this will work not just for the function \( x^2 + y^2 \) but also for any function \( Q(x, y) \), where \( a, b, \) and \( c \) are integer constants. Such a function \( Q(x, y) \) with at least one of the coefficients \( a, b, c \) nonzero is called a *quadratic form*, or sometimes just a *form* for short.

Solving \( x^2 + y^2 = n \) amounts to representing \( n \) in the form of the sum of two squares. More generally, solving \( Q(x, y) = n \) is called representing \( n \) by the form \( Q(x, y) \). So the overall goal is to solve the *representation problem*: Which numbers \( n \) are represented by a given form \( Q(x, y) \), and how does one find such representations?

Before starting to describe the method for displaying the values of a quadratic form graphically let us make a preliminary observation: If the greatest common divisor of two integers \( x \) and \( y \) is \( d \), then we can write \( x = dx' \), \( y = dy' \), and \( Q(x, y) = d^2Q(x', y') \) where the greatest common divisor of \( x' \) and \( y' \) is 1. Hence it suffices to find the values of \( Q \) on *primitive* pairs \((x, y)\), the pairs whose greatest common divisor is 1, and then multiply these values by arbitrary squares \( d^2 \).

In a similar way, if the coefficients \( a, b, c \) of a form \( Q(x, y) = ax^2 + bxy + cy^2 \) have greatest common divisor \( d \), so \( a = da' \), \( b = db' \), and \( c = dc' \) for integers \( a', b', c' \) whose greatest common divisor is 1, then \( Q(x, y) = d(a'x'^2 + b'xy + c'y'^2) = dQ'(x, y) \) for the form \( Q'(x, y) = a'x'^2 + b'xy + c'y'^2 \). Multiplying all the values of a form by a constant \( d \) is a fairly trivial operation, so for most purposes it suffices to restrict attention to forms for which the greatest common divisor of the coefficients is 1. Such forms are called *primitive forms*.

Primitive pairs \((x, y)\) correspond almost exactly to fractions \( x/y \) that are reduced to lowest terms, the only ambiguity being that both \((x, y)\) and \((-x, -y)\) correspond to the same fraction \( x/y \). However, this ambiguity does not affect the value
of a quadratic form \( Q(x, y) = ax^2 + bxy + cy^2 \) since \( Q(x, y) = Q(-x, -y) \). This means that we can regard \( Q(x, y) \) as being essentially a function \( f(x/y) \). Notice that we are not excluding the possibility \( (x, y) = (1, 0) \) which corresponds to the “fraction” \( 1/0 \).

1. The Topograph

We already have a nice graphical representation of the rational numbers \( x/y \) and \( 1/0 \) as the vertices in the Farey diagram. Here is a picture of the diagram with the so-called dual tree superimposed:

![Dual Tree](image)

The dual tree has a vertex in the center of each triangle of the Farey diagram, and it has an edge crossing each edge of the Farey diagram. As with the Farey diagram, we can only draw a finite part of the dual tree. The actual dual tree has branching that repeats infinitely often, an unending bifurcation process with smaller and smaller twigs.

The tree divides the interior of the large circle into regions, each of which is adjacent to one vertex of the original diagram. We can write the value \( Q(x, y) \) in the region adjacent to the vertex \( x/y \). This is shown in the figure below for the quadratic form \( Q(x, y) = x^2 + y^2 \), where to unclutter the picture we no longer draw the triangles of the original Farey diagram.
For example the 13 in the region adjacent to the fraction $2/3$ represents the value $2^2 + 3^2$, and the 29 in the region adjacent to $5/2$ represents the value $5^2 + 2^2$.

For a quadratic form $Q$ this picture showing the values $Q(x,y)$ is called the topograph of $Q$. It turns out that there is a very simple method for computing the topograph from just a very small amount of initial data. This method is based on the following:

**Arithmetic Progression Rule.** If the values of $Q(x,y)$ in the four regions surrounding an edge in the tree are $p$, $q$, $r$, and $s$ as indicated in the figure, then the three numbers $p$, $q + r$, $s$ form an arithmetic progression.

We can check this in the topograph of $x^2 + y^2$ shown above. Consider for example one of the edges separating the values 1 and 2. The values in the four regions surrounding this edge are 1, 1, 2, 5 and the arithmetic progression is 1, 1 + 2, 5. For an edge separating the values 1 and 5 the arithmetic progression is 2, 1 + 5, 10. For an edge separating the values 5 and 13 the arithmetic progression is 2, 5 + 13, 34. And similarly for all the other edges.

The arithmetic progression rule implies that the values of $Q$ in the three regions surrounding a single vertex of the tree determine the values in all other regions, by starting at the vertex where the three adjacent values are known and working one's way outward in the dual tree. The easiest place to start for a quadratic form $Q(x,y) = ax^2 + bxy + cy^2$ is with the three values $Q(1,0) = a$, $Q(0,1) = c$, and $Q(1,1) = a + b + c$ for the three fractions $1/0$, $0/1$, and $1/1$. Here are two examples:
In the first case we start with the values 1 and 2 together with the 3 just above them. These determine the value 9 above the 2 via the arithmetic progression 1, 2 + 3, 9. Similarly the 6 above the 1 is determined by the arithmetic progression 2, 1 + 3, 6. Next one can fill in the 19 next to the 9 we just computed, using the arithmetic progression 3, 2 + 9, 19, and so on for as long as one likes.

The procedure for the other form $x^2 - 2y^2$ is just the same, but here there are negative as well as positive values. The edges that separate positive values from negative values will be important later, so we have indicated these edges by special shading.

Perhaps the most noticeable thing in both the examples $x^2 + 2y^2$ and $x^2 - 2y^2$ is the fact that the values in the lower half of the topograph are the same as those in the upper half. We could have predicted in advance that this would happen because $Q(x, y) = Q(-x, y)$ whenever $Q(x, y)$ has the form $ax^2 + cy^2$, with no $xy$ term. The topograph for $x^2 + y^2$ has even more symmetry since the values of $x^2 + y^2$ are unchanged when $x$ and $y$ are switched, so the topograph has left-right symmetry as well.

Here is a general observation: The three values around one vertex of the topograph can be specified arbitrarily. For if we are given three numbers $a$, $b$, $c$ then the quadratic form $ax^2 + (c - a - b)xy + by^2$ takes these three values for $(x, y)$ equal to $(1, 0)$, $(0, 1)$, $(1, 1)$.

**Proof of the Arithmetic Progression Rule:** Let the two vertices of the Farey diagram corresponding to the values $q$ and $r$ have labels $x_1/y_1$ and $x_2/y_2$ as in the figure below. Then by the mediant rule for labeling vertices, the labels on the $p$ and $s$ regions are the fractions shown. Note that these labels are correct even when $x_1/y_1 = 1/0$ and $x_2/y_2 = 0/1$. 

Chapter 4 Quadratic Forms

For a quadratic form $Q(x, y) = ax^2 + bxy + cy^2$ we then have

$$s = Q(x_1 + x_2, y_1 + y_2) = a(x_1 + x_2)^2 + b(x_1 + x_2)(y_1 + y_2) + c(y_1 + y_2)^2$$

$$= ax_1^2 + bx_1 y_1 + cy_1^2 + ax_2^2 + bx_2 y_2 + cy_2^2 + (\cdots)$$

$$Q(x_1, y_1) = q$$ 

$$Q(x_2, y_2) = r$$

Similarly we have

$$p = Q(x_1 - x_2, y_1 - y_2) = ax_1^2 + bx_1 y_1 + cy_1^2 + ax_2^2 + bx_2 y_2 + cy_2^2 - (\cdots)$$

$$Q(x_1, y_1) = q$$ 

$$Q(x_2, y_2) = r$$

The terms in $(\cdots)$ are the same in both cases, namely the terms involving both subscripts 1 and 2. If we compute $p + s$ by adding the two formulas together, the terms $(\cdots)$ will therefore cancel, leaving just $p + s = 2(q + r)$. This equation can be rewritten as $(q + r) - p = s - (q + r)$, which just says that $p, q + r, s$ forms an arithmetic progression.

2. Periodic Separator Lines

For most quadratic forms that take on both positive and negative values, such as $x^2 - 2y^2$, there is another way of drawing the topograph that reveals some hidden and unexpected properties. For the form $x^2 - 2y^2$ there is a zigzag path of edges in the topograph separating the positive and negative values, and if we straighten this path out to be a line, called the separator line, what we see is the following infinitely repeated pattern:
To construct this, one can first build the separator line starting with the three values $Q(1, 0) = 1$, $Q(0, 1) = -2$, and $Q(1, 1) = -1$. Place these as shown in part (a) of the figure below, with a horizontal line segment separating the positive from the negative values.

To extend the separator line one step farther to the right, apply the arithmetic progression rule to compute the next value 2 using the arithmetic progression $-2, 1 - 1, 2$. Since this value 2 is positive, we place it above the horizontal line and insert a vertical edge to separate this 2 from the 1 to the left of it, as in (b) of the figure. Now we repeat the process with the next arithmetic progression $1, 2 - 1, 1$ and put the new 1 above the horizontal line with a vertical edge separating it from the previous 2, as shown in (c). At the next step we compute the next value $-2$ and place it below the horizontal line since it is negative, giving (d). One more step produces (e) where we see that further repetitions will produce a pattern that repeats periodically as we move to the right. The arithmetic progression rule also implies that it repeats periodically to the left, so it is periodic in both directions:

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(a)   (b)   (c)   (d)   (e)

Thus we have the periodic separator line. To get the rest of the topograph we can then work our way upward and downward from the separator line, as shown in the original figure. As one moves upward from the separator line, the values of $Q$ become larger and larger, approaching $+\infty$ monotonically, and as one moves downward the values approach $-\infty$ monotonically. The reason for this will become clear in the next chapter when we discuss something called the Monotonicity Property.
An interesting property of this form $x^2 - 2y^2$ that is evident from its topograph is that it takes on the same negative values as positive values. This would have been hard to predict from the formula $x^2 - 2y^2$. Indeed, for the similar-looking quadratic form $x^2 - 3y^2$ the negative values are quite different from the positive values, as one can see in its straightened-out topograph:

$$Q_1(x, y) = x^2 - 3y^2$$

3. Continued Fractions Once More

There is a close connection between the topograph for a quadratic form $x^2 - dy^2$ and the infinite continued fraction for $\sqrt{d}$ when $d$ is a positive integer that is not a square. In fact, we will see that the topograph can be used to compute the continued fraction for $\sqrt{d}$. As an example let us look at the case $d = 2$. The relevant portion of the topograph for $x^2 - 2y^2$ is the strip along the line separating the positive and negative values:

$$\begin{array}{c|c|c|c|c|c|c|c|c|c}
2 & 1 & 2 & 1 & 2 & 1 & 2 & 1 & 1 \\
-1 & -2 & -1 & -2 & -1 & -2 & -1 & -2 & -1
\end{array}$$

This is a part of the dual tree of the Farey diagram. If we superimpose the triangles of the Farey diagram corresponding to this part of the dual tree we obtain an infinite strip of triangles:

Ignoring the dotted triangles to the left, the infinite strip of triangles corresponds to the infinite continued fraction $1 + \frac{1}{\sqrt{2}}$. We could compute the value of this continued fraction by the method in Chapter 2, but there is an easier way using the quadratic
form $x^2 - 2y^2$. For fractions $\frac{x}{y}$ labeling the vertices along the infinite strip, the corresponding values $n = x^2 - 2y^2$ are either $\pm 1$ or $\pm 2$. We can rewrite the equation $x^2 - 2y^2 = n$ as $(\frac{x}{y})^2 = 2 + \frac{n}{y^2}$. As we go farther and farther to the right in the infinite strip, both $x$ and $y$ are getting larger and larger while $n$ only varies through finitely many values, namely $\pm 1$ and $\pm 2$, so the quantity $\frac{n}{y^2}$ is approaching 0. The equation $(\frac{x}{y})^2 = 2 + \frac{n}{y^2}$ then implies that $(\frac{x}{y})^2$ is approaching 2, so we see that $\frac{x}{y}$ is approaching $\sqrt{2}$. Since the fractions $\frac{x}{y}$ are also approaching the value of the infinite continued fraction $1 + \frac{1}{2}$ that corresponds to the infinite strip, this implies that the value of the continued fraction $1 + \frac{1}{2}$ is $\sqrt{2}$.

Here is another example, for the quadratic form $x^2 - 3y^2$, showing how $\sqrt{3} = 1 + \frac{1}{1} + \frac{1}{2}$.

After looking at these two examples one can see that it is not really necessary to draw the strip of triangles, and one can just read off the continued fraction directly from the periodic separator line. Let us illustrate this by considering the form $x^2 - 10y^2$:

If one moves toward the right along the horizontal line starting at a point in the edge separating the $\frac{1}{1}$ region from the $\frac{0}{1}$ region, one first encounters 3 edges leading off to the right (downward), then 6 edges leading off to the left (upward), then 6 edges leading off to the right, and thereafter 6 edges leading off to the left and right alternately. This means that the continued fraction for $\sqrt{10}$ is $3 + \frac{1}{6}$.

Here is a more complicated example showing how to compute the continued fraction for $\sqrt{19}$ from the form $x^2 - 19y^2$:
From this we read off that \( \sqrt{19} = 4 + \frac{1}{2} + \frac{1}{1 + \frac{1}{3 + \frac{1}{1 + \frac{1}{2 + \frac{1}{8}}}}} \).

In the next chapter we will prove that the topograph of the form \( x^2 - dy^2 \) always has a periodic separator line whenever \( d \) is a positive integer that is not a square. As in the examples above, this separator line always includes the edge of the dual tree separating the vertices 1/0 and 0/1 since the form takes the positive value +1 on 1/0 and the negative value \(-d\) on 0/1. The periodicity then implies that the continued fraction for \( \sqrt{d} \) has the form

\[
\sqrt{d} = a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}}
\]

with the periodic part starting immediately after the initial term \( a_0 \). In addition to being periodic, the separator line also has mirror symmetry with respect to reflection across the vertical line corresponding to the edge connecting 1/0 to 0/1 in the Farey diagram. This is because the form \( x^2 - dy^2 \) has no xy term, so replacing \( x/y \) by \(-x/y\) does not change the value of the form. Replacing \( x/y \) by \(-x/y\) reflects the circular Farey diagram across the horizontal edge from 1/0 to 0/1, and this reflects the periodic separator line across the vertical line through the 1/0 and 0/1 regions. Once the separator line has symmetry with respect to this vertical line, the periodicity forces it to have mirror symmetry with respect to an infinite sequence of vertical lines, as illustrated in the following figure for the form \( x^2 - 19y^2 \):

\[
\begin{array}{cccccccc}
1 & 6 & 5 & 9 & 9 & 5 & 6 & 1 \\
\end{array}
\]

\( a_0 \) \hspace{1cm} \text{palindrome} \hspace{1cm} a_n = 2a_0

In particular, these mirror symmetries imply that the continued fraction

\[
\sqrt{d} = a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}}
\]

always has two special properties:

(a) \( a_n = 2a_0 \).

(b) The intermediate terms \( a_1, a_2, \ldots, a_{n-1} \) form a palindrome, reading the same forward as backward.

Thus in \( \sqrt{19} = 4 + \frac{1}{2} + \frac{1}{1 + \frac{1}{3 + \frac{1}{1 + \frac{1}{2 + \frac{1}{8}}}}} \) the final 8 is twice the initial 4, and the intermediate terms 2, 1, 3, 1, 2 form a palindrome. These special properties held
also in the earlier examples, but were less apparent because there were fewer terms in the repeated part of the continued fraction.

In some cases there is an additional kind of symmetry along the separator line, as illustrated for the form \( x^2 - 13y^2 \):

\[
\begin{array}{cccccccc}
1 & 3 & 4 & 9 & 12 & 13 & 12 & 9 & 4 & 3 & 1 \\
-13 & -12 & -9 & -4 & -3 & -1 & -3 & -4 & -9 & -12 & -13
\end{array}
\]

As before there is a horizontal translation giving the periodicity and there are mirror symmetries across vertical lines, but now there is an extra glide-reflection along the strip that interchanges the positive and negative values of the form. Performing this glide-reflection twice in succession gives the translational periodicity. Notice that there are also 180 degree rotational symmetries about the points marked with dots on the separator line, and these rotations account for the palindromic middle part of the continued fraction

\[
\sqrt{13} = 3 + \frac{1}{1 + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{6}}}}}}
\]

The fact that the periodic part has odd length corresponds to the separator strip having the glide-reflection symmetry. We could rewrite the continued fraction to have a periodic part of even length by doubling the period,

\[
\sqrt{13} = 3 + \frac{1}{\frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{1} + \frac{1}{\sqrt{6}}}}}}}}}}
\]

and this corresponds to ignoring the glide-reflection and just considering the translational periodicity.

We have been using quadratic forms \( x^2 - dy^2 \) to compute the continued fractions for irrational numbers \( \sqrt{d} \), but everything works just the same for irrational numbers \( \sqrt{p/q} \) using the quadratic form \( qx^2 - py^2 \) in place of \( x^2 - dy^2 \). Following the same reasoning as before, if the equation \( qx^2 - py^2 = n \) is rewritten as \( q(\frac{x}{y})^2 = p + \frac{n}{y^2} \) then we see that as we move out along the periodic separator line the numbers \( x \) and \( y \) approach infinity while \( n \) cycles through finitely many values, so the term \( \frac{n}{y^2} \) approaches 0 and the fractions \( \frac{x}{y} \) approach a number \( z \) satisfying \( qz^2 = p \), so \( z = \sqrt{p/q} \). This argument depends of course on the existence of a periodic separator line, and we will prove in the next chapter that forms \( qx^2 - py^2 \) always have a periodic separator line if \( p \) and \( q \) are positive and the roots \( \pm \sqrt{p/q} \) of \( qz^2 - p = 0 \) are irrational.

Here are two examples. For the first one we use the form \( 3x^2 - 7y^2 \) to compute the continued fraction for \( \sqrt{7/3} \).
This gives \( \sqrt{7/3} = 1 + \frac{1}{1} + \frac{1}{8} + \frac{1}{1} + \frac{1}{1} + \frac{1}{2} \). For the second example we use \( 10x^2 - 29y^2 \) to compute the continued fraction for \( \sqrt{29/10} \),

\[
\begin{array}{cccccccccccc}
3 & 5 & 12 & 17 & 20 & 21 & 20 & 17 & 12 & 5 & 3 & 5 & 12 & 17 & 20 & 21 \\
\vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots \\
-7 & -4 & -1 & -4 & -7 & -4 & 1 \\
\end{array}
\]

with the result that \( \sqrt{29/10} = 1 + \frac{1}{1} + \frac{1}{2} + \frac{1}{2} + \frac{1}{1} + \frac{1}{2} \). The period of odd length here corresponds to the existence of the glide-reflection and 180 degree rotation symmetries.

The palindrome property and the relation \( a_n = 2a_0 \) that we observed in the continued fraction for \( \sqrt{d} \) still hold for irrational numbers \( \sqrt{p/q} \) assuming that \( a_0 > 0 \), which is equivalent to the condition \( p/q > 1 \) since \( a_0 \) is the integer part of \( \sqrt{p/q} \). The key point is that the form \( qx^2 - py^2 \) is unchanged when the sign of \( x \) is changed, so its topograph has mirror symmetry with respect to reflection across a line through the 1/0 and 0/1 regions, and this symmetry implies the special properties of the continued fraction for \( \sqrt{p/q} \).

Fractions \( p/q \) between 0 and 1 can be dealt with just by inverting them, interchanging \( p \) and \( q \). Inverting a continued fraction \( a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}} \) changes it to \( \frac{1}{a_0 + \frac{1}{a_1 + \frac{1}{a_2 + \cdots + \frac{1}{a_n}}}} \). For example, from the earlier computation of \( \sqrt{7/3} \) we obtain \( \sqrt{3/7} = \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{8} + \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{2} \). Interchanging \( p \) and \( q \) changes the form \( qx^2 - py^2 \) to \( px^2 - qy^2 \), and we could achieve the same effect by first interchanging \( x \) and \( y \) to get \( qy^2 - px^2 \) and then changing the sign of this form to get \( px^2 - qy^2 \). Interchanging \( x \) and \( y \) inverts the fractions \( x/y \) labeling the regions along the separator line, which corresponds to reflecting the periodic separator line across a horizontal line, and then changing the sign of the form to put the positive values above the horizontal line again. In the example of \( \sqrt{7/3} \) the reflected separator line for \( 7x^2 - 3y^2 \) is thus

\[
\begin{array}{cccccccccccc}
7 & 4 & 1 & 4 & 7 & 4 & 1 \\
-3 & -5 & -12 & -17 & -20 & -21 & -17 & -12 & -5 & -3 & -5 & -12 & -17 & -20 & -21 \\
\end{array}
\]
so we see that this yields the continued fraction \( \frac{1}{1} + \frac{1}{1} + \frac{1}{1} + \frac{1}{8} + \frac{1}{1} + \frac{1}{2} \) for \( \sqrt{3/7} \).

One might ask whether the irrational numbers \( \sqrt{p/q} \) are the only numbers having a continued fraction \( a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) or \( \frac{1}{a_0} + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) satisfying the palindrome property and the relation \( a_n = 2a_0 \). Here we should restrict attention only to positive irrational numbers since the numbers \( a_0, a_1, \ldots, a_n \) must all be positive. The answer is Yes, as we will see later in this section.

Quadratic forms can be used to compute the continued fractions for all quadratic irrationals. To illustrate the general method let us find the continued fraction for \( \frac{10 + \sqrt{2}}{14} \) which is a root of the equation \( 14z^2 - 20z + 7 = 0 \). The associated quadratic form is \( 14x^2 - 20xy + 7y^2 \), obtained by setting \( z = x/y \) and then multiplying by \( y^2 \). We would like to find a periodic separator line in the topograph of this form. To do this we start with the three values at \( 1/0, 0/1 \), and \( 1/1 \), which are the positive numbers 14, 7 and 1, and then use the arithmetic progression rule to move in a direction that leads to negative values since the separator line separates positive and negative values of the form. In this way we are led to a separator line which is indeed periodic:

\[
\begin{array}{cccccccc}
-2 & -1 & -2 & -1 & -2 & -1 & -2 & -1 \\
1 & 2 & 1 & 2 & 1 & 2 & 1 & 2 \\
\end{array}
\]

This figure lies in the upper half of the circular Farey diagram where the fractions \( x/y \) labeling the regions in the topograph are positive. If we follow the separator line out to either end the labels \( x/y \) have both \( x \) and \( y \) increasing monotonically and approaching infinity, as a consequence of the mediant rule for labeling vertices of the Farey diagram. Hence the values

\[
14z^2 - 20z + 7 = 14\left(\frac{x}{y}\right)^2 - 20\left(\frac{x}{y}\right) + 7 = \frac{14x^2 - 20xy + 7y^2}{y^2}
\]

are approaching zero since the values of the numerator \( 14x^2 - 20xy + 7y^2 \) on the right just cycle through a finite set of numbers repeatedly, the values of the form along the separator line, while the denominators \( y^2 \) approach infinity. Thus the labels \( x/y \) are approaching the roots of the equation \( 14z^2 - 20z + 7 = 0 \). Since we are in the upper half of the Farey diagram, the smaller of the two roots, which is \( \frac{10 - \sqrt{2}}{14} \), is the limit toward the right along the separator line and the larger root \( \frac{10 + \sqrt{2}}{14} \) is the limit toward the left.

To get the continued fraction for the smaller root we follow the path in the topograph that starts with the edge between \( 1/0 \) and \( 0/1 \), then zigzags up to the separator
line, then goes out this line to the right. If we straighten this path out it looks like the following:

\[
\begin{array}{c}
\frac{1}{0} \\
\frac{0}{1}
\end{array}
\]

The continued fraction is therefore

\[
\frac{10 - \sqrt{2}}{14} = \frac{1}{1 + \frac{1}{1 + \frac{1}{1 + \frac{1}{1 + \frac{1}{2}}}}}
\]

It is not actually necessary to redraw the straightened-out path since in the original form of the topograph we can read off the sequence of left and right “side roads” as we go along the path, the sequence \(LRLRLRR\) where \(L\) denotes a side road to the left and \(R\) a side road to the right. This sequence determines the continued fraction.

For the other root \(\frac{10 + \sqrt{2}}{14}\) the straightened-out path has the following shape:

\[
\begin{array}{c}
\frac{1}{0} \\
\frac{0}{1}
\end{array}
\]

The sequence of side roads is \(LRRR\) so the continued fraction is

\[
\frac{10 + \sqrt{2}}{14} = \frac{1}{1 + \frac{1}{4 + \frac{1}{2}}}
\]

We will show that this procedure works for all quadratic irrational numbers, and this will prove the harder half of Lagrange’s Theorem:

**Proposition 4.1.** *The continued fraction for every quadratic irrational is eventually periodic.*

The proof will involve associating a quadratic form to each quadratic irrational, and we will need to use the fact that the quadratic forms arising in this way all have periodic separator lines. This will be proved in the next chapter, so the proof will not be officially complete until then.

Before beginning the proof let us say a few things about the structure of all infinite strips in the Farey diagram, whether periodic or not. By an infinite strip in the Farey diagram we mean a collection of fans each consisting of a finite number of triangles, each fan intersecting the next along an edge of a zigzag path extending infinitely far at both ends.

To see how the strip lies in the upper half-plane model of the Farey diagram let \(L\) be a line running down the middle of the strip from end to end. The line \(L\) cannot
cross only vertical edges of the diagram, the edges with one end at 1/0, otherwise the strip would consist of a single infinite fan, which is not allowed as an infinite strip. Thus \( L \) must cross some semicircular edges. As it crosses such an edge we can assume it crosses from above the edge to below it, by re-orienting the direction of \( L \) if necessary. After \( L \) crosses the edge downward into a triangle, it will next cross one of the other two shorter edges of this triangle moving downward again. All subsequent crossings will then be downward as well. The semicircles crossed are becoming smaller and smaller with diameters approaching zero, as we saw in our initial discussion of infinite continued fractions, and there is a unique limiting point \( \alpha \) on the \( x \)-axis for this end of the strip of triangles. This is the unique point that lies between the two endpoints of each semicircle crossed by \( L \) on its downward path.

The vertical line \( V_\alpha \) going upward from \( \alpha \) will pass through triangles of the strip, either staying in the strip forever or leaving the strip by crossing the upper semicircular edge of a triangle \( T \) of the strip. In the latter case the line \( L \), which passes through the same upward sequence of triangles as \( V_\alpha \) until reaching \( T \), must exit \( T \) by turning and crossing the other smaller semicircular edge of \( T \) in the downward direction. As we saw for the other end of \( L \) it will then continue downward forever, passing through all the triangles of the other end of the strip and limiting on an irrational number \( \beta \). The vertical line \( V_\beta \) going upward from \( \beta \) will pass through the same set of triangles until reaching the triangle \( T \) where it will also exit the strip by crossing the upper edge of \( T \). We can then deform \( L \) so that it consists of the parts of \( V_\alpha \) and \( V_\beta \) below \( T \) joined by a bending arc within \( T \). Notice that the vertex 1/0 is not a vertex of the strip in this case.

There remains the possibility that \( V_\alpha \) remains in the strip forever as we move upward, so eventually it lies in a triangle \( T_\alpha \) of the strip having 1/0 as a vertex. One end of the line \( L \) runs parallel to \( V_\alpha \) until it reaches \( T_\alpha \), then it turns right or left to cross a finite number of other triangles having 1/0 as a vertex before turning downward to cross the lower edge of one of these triangles \( T_\beta \). After this it will travel monotonically downward by the earlier argument, limiting on an irrational number \( \beta \) in the \( x \)-axis. We can deform \( L \) to consist of parts of \( V_\alpha \) and the vertical line \( V_\beta \) through \( \beta \), joined by an arc crossing from \( T_\alpha \) to \( T_\beta \).

A consequence of these considerations is that there can be only one infinite strip.
in the Farey diagram whose ends converge to a given pair of irrationals \( \alpha \) and \( \beta \). This is because the vertical lines \( V_\alpha \) and \( V_\beta \) are uniquely determined by \( \alpha \) and \( \beta \), and the triangles \( T \) or \( T_\alpha \) and \( T_\beta \) are also determined uniquely since in the case of a triangle \( T \) both \( \alpha \) and \( \beta \) lie in the same interval in the \( x \) axis between consecutive integers and \( T \) is the smallest triangle of the Farey diagram whose projection to the \( x \) axis contains both \( \alpha \) and \( \beta \), while in the case of two triangles \( T_\alpha \) and \( T_\beta \) the numbers \( \alpha \) and \( \beta \) lie in different intervals between consecutive integers and \( T_\alpha \) and \( T_\beta \) are the triangles with vertex \( 1/0 \) that project to these two intervals.

A nice way to construct an infinite strip joining any two irrationals \( \alpha \) and \( \beta \) is to take all the triangles in the Farey diagram that meet the semicircle in the upper half-plane with endpoints \( \alpha \) and \( \beta \). This semicircle can cross an edge of the Farey diagram only once since if two semicircles with endpoints on the \( x \)-axis intersect in more than two points then they must coincide. Nor can two semicircles with endpoints on the \( x \)-axis be tangent unless the point of tangency is one of the endpoints, but this does not happen here since \( \alpha \) and \( \beta \) are irrational while the endpoints of edges of the Farey diagram are rational. From these observations we see that if the semicircle from \( \alpha \) to \( \beta \) intersects a triangle of the Farey diagram, then it crosses this triangle from one edge to another edge. The semicircle also cannot cross an infinite number of triangles having a common vertex, otherwise the semicircle would have to have this common vertex as an endpoint, making \( \alpha \) or \( \beta \) rational. Thus the union of all the triangles crossed by the semicircle is an infinite strip.

We have shown that an infinite strip is uniquely determined by its endpoints, so this implies that the semicircle from \( \alpha \) to \( \beta \) crosses exactly the same triangles as the line we constructed earlier consisting of two vertical segments joined at the top by a 180 degree bend. This may seem odd at first glance, but what it means is that the height of the vertical segments cannot be very large compared to the distance between them.

The construction of a strip connecting two irrational numbers \( \alpha \) and \( \beta \) via the semicircle with endpoints \( \alpha \) and \( \beta \) works equally well when \( \alpha \) or \( \beta \) is rational, but in this case the strip has only a finite number of triangles at a rational end. A very special case is when \( \alpha \) and \( \beta \) are the endpoints of an edge of the Farey diagram, when the strip degenerates to just this edge.

Proof of Proposition 4.1. Quadratic irrationals are the numbers \( \alpha = A + B \sqrt{n} \) where \( A \) and \( B \) are rational, \( B \) is nonzero, and \( n \) is a positive integer that is not a square. The first step is to find a quadratic equation with integer coefficients having \( \alpha \) as a root. From the quadratic formula we know the other root will have to be the conjugate \( \overline{\alpha} = A - B \sqrt{n} \), so a quadratic equation having \( \alpha \) and \( \overline{\alpha} \) as roots is \( (z - \alpha)(z - \overline{\alpha}) = 0 \). Multiplied out, this becomes \( z^2 - (\alpha + \overline{\alpha})z + \alpha \overline{\alpha} = z^2 - 2Az + (A^2 - B^2n) = 0 \) which has rational coefficients since \( A \) and \( B \) are rational. After multiplying by a common denominator for the coefficients this becomes an equation \( az^2 + bz + c = 0 \) with
integer coefficients having $\alpha$ and $\overline{\alpha}$ as roots. Here $a > 0$ since it is the common denominator we multiplied by.

The associated quadratic form is $ax^2 + bxy + cy^2$. This form has two special properties:

(1) Its topograph contains both positive and negative values. This is because the quadratic polynomial $az^2 + bz + c = a(z - \alpha)(z - \overline{\alpha})$ takes negative values when $z$ is between the two roots $\alpha$ and $\overline{\alpha}$, where the two factors in parentheses have opposite sign, and positive values when $z$ is greater than both roots or less than both roots, so the two parenthetical factors have the same sign. Thus there are rational numbers $z = x/y$ where the left side of the equation

$$a\left(\frac{x}{y}\right)^2 + b\left(\frac{x}{y}\right) + c = \frac{ax^2 + bxy + cy^2}{y^2}$$

has both signs, hence the same is true for the numerator on the right.

(2) The topograph of $ax^2 + bxy + cy^2$ does not contain the value 0. For suppose there was a pair $(x, y) \neq (0, 0)$ with $ax^2 + bxy + cy^2 = 0$. Since $a \neq 0$ we cannot have $y = 0$. Then for $y \neq 0$ the previous displayed equation would say that $x/y$ was a rational root of $az^2 + bz + c = 0$, contradicting the fact that its roots $\alpha$ and $\overline{\alpha}$ are irrational.

In Theorem 5.2 in the next chapter we will show that every form $ax^2 + bxy + cy^2$ satisfying these two conditions has a periodic separator line in its topograph. This corresponds to an infinite periodic strip in the Farey diagram. We claim that the ends of this strip must be at the roots $\alpha$ and $\overline{\alpha}$ of the equation $az^2 + bz + c = 0$. To see why this is true, consider first the case of an end that approaches a number on the positive $x$-axis in the upper half-plane model of the Farey diagram. The labels $x/y$ of the vertices along this end of the strip have both $x$ and $y$ approaching infinity since these labels are obtained by applying the mediant rule to positive numbers repeatedly, so each application of the rule increases both $x$ and $y$. (The vertex $0/1$ might create a problem here, but this vertex can belong to only finitely many triangles in the strip so eventually all the labels $x/y$ have both $x$ and $y$ strictly positive.)

The other case would be an end of the strip approaching a negative number, and in this case the vertex labels $x/y$ have $x$ approaching minus infinity and $y$ approaching plus infinity since we can just reflect the strip across the vertical axis of the plane to reduce to the previous case.

Since the denominators $y$ approach infinity as we go out to an end of the periodic infinite strip while the values of the form $ax^2 + bxy + cy^2$ cycle through finitely many values, it follows that the values of the right side of the equation

$$a\left(\frac{x}{y}\right)^2 + b\left(\frac{x}{y}\right) + c = \frac{ax^2 + bxy + cy^2}{y^2}$$

are approaching zero. This means that the vertex labels $x/y$ are approaching one of
the roots of the equation \( az^2 + bz + c = 0 \). By the same reasoning, if we go out toward the other end of the strip the labels \( x/y \) approach the other root.

Finally, to get the continued fraction for the given root \( \alpha \) we just take the strip of triangles meeting the vertical line through \( \alpha \). This strip will start at the vertex \( 1/0 \) at the top and then move downward through an infinite sequence of triangles that eventually coincide with the triangles in one end of the infinite periodic strip. This means that the continued fraction for \( \alpha \) is eventually periodic.

**Proposition 4.2.** The numbers \( \sqrt{p/q} \) are the only quadratic irrationals having continued fractions \( a_0 + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) or \( \frac{1}{a_0} + \frac{1}{a_1} + \frac{1}{a_2} + \cdots + \frac{1}{a_n} \) satisfying the palindrome property and the relation \( a_n = 2a_0 \).

**Proof:** Suppose the continued fraction for a quadratic irrational \( \alpha \) satisfies these conditions. In particular \( \alpha \) must be positive since \( a_0 \) is positive, being half the positive number \( a_n \). The strip in the Farey diagram corresponding to this continued fraction starts at the \( (1/0, 0/1) \) edge and goes out to \( \alpha \) at its end. Combining this strip with its reflection across the \( (1/0, 0/1) \) edge gives an infinite strip with mirror symmetry across the \( (1/0, 0/1) \) edge, and this strip is periodic everywhere, even at the junction along the \( (1/0, 0/1) \) edge since \( a_n = 2a_0 \). The other end of this strip is \( \overline{\alpha} \) since we have seen that the two endpoints of a periodic strip satisfy a single quadratic equation \( T(z) = z \) where \( T \) is the periodicity transformation. The two roots of this equation are conjugates \( \alpha \) and \( \overline{\alpha} \) but they are also negatives of each other by the mirror symmetry across the edge \( (1/0, 0/1) \). Thus \( \overline{\alpha} = -\alpha \). If \( \alpha = A + B\sqrt{n} \) with \( A \) and \( B \) rational this implies that \( A = 0 \). Since \( \alpha \) is positive, it is then the square root of the rational number \( B^2n \).

**Proposition 4.3.** Every periodic line in the dual tree of the Farey diagram occurs as the separator line for some form.

**Proof:** Given a periodic line, the periodicity of this line and of the corresponding infinite strip is realized by some linear fractional transformation \( T \). As we saw in Chapter 3, the endpoints of the strip are the fixed points of \( T \), the solutions of \( T(z) = z \). This can be rewritten as a quadratic equation \( az^2 + bz + c = 0 \) with integer coefficients. The coefficient \( a \) must be nonzero, otherwise we would have an equation \( bz + c = 0 \) with only one root if \( b \neq 0 \), while if \( b = 0 \) the equation would have no roots if \( c \neq 0 \). If \( c = 0 \) as well as \( a = 0 \) and \( b = 0 \) the equation would degenerate to \( 0 = 0 \), meaning that every \( z \) satisfied \( T(z) = z \) so \( T \) would be the identity transformation rather than the periodicity transformation, a contradiction. Thus \( a \) must be nonzero, and by multiplying the equation by \(-1\) if necessary we may assume that \( a > 0 \).

We claim that the the periodic line we started with is a separator line in the topograph of the form \( ax^2 + bxy + cy^2 \). This just means that the values of the form at vertices on one edge of the associated periodic strip are all positive and the values
on the other edge are all negative. To see why this is so let us factor $az^2 + bz + c$ as $a(z - \alpha)(z - \overline{\alpha})$ where $\alpha$ and $\overline{\alpha}$ are the roots of $az^2 + bz + c = 0$ at the ends of the strip. From this factorization and the fact that $a$ is positive we see that the product $a(z - \alpha)(z - \overline{\alpha})$ is negative if $z$ is between $\alpha$ and $\overline{\alpha}$ and positive if $z$ is greater than both $\alpha$ and $\overline{\alpha}$ or less than both $\alpha$ and $\overline{\alpha}$. (We saw this previously in the proof of Proposition 4.1.) Taking $z$ to be a rational number $x/y$, the equation

$$a\left(\frac{x}{y}\right)^2 + b\left(\frac{x}{y}\right) + c = \frac{ax^2 + bx y + c y^2}{y^2}$$

implies that the form $ax^2 + bx y + c y^2$ takes negative values for $x/y$ in the interval between $\alpha$ and $\overline{\alpha}$ and positive values for $x/y$ outside this interval, assuming $x/y \neq 1/0$ so we are not dividing by 0 in the equation above.

In terms of the circular Farey diagram the roots $\alpha$ and $\overline{\alpha}$ divide the boundary circle into two arcs, with the form taking positive values at vertices in one arc and negative values at vertices in the other arc, with the possible exception of the vertex $1/0$. However, this vertex is not actually exceptional since it lies in the arc with positive values and the form takes the value $a > 0$ when $x/y = 1/0$. This proves what we wanted to show since vertices along one edge of the strip lie in one arc and vertices along the other edge lie in the other arc.

To illustrate the procedure in the preceding proof let us find a quadratic form whose periodic separator line is the following:

The fractional labels correspond to vertices of the underlying Farey diagram, and from these we see that the translation giving the periodicity sends $1/0$ to $25/36$ and $0/1$ to $84/121$. The matrix of this transformation is \(\begin{pmatrix} 25 & 84 \\ 36 & 121 \end{pmatrix}\) so it is the transformation $T(z) = (25z + 84)/(36z + 121)$. The fixed points of $T$ are determined by setting this equal to $z$. The resulting equation simplifies to $25z + 84 = 36z^2 + 121z$ and then $36z^2 + 96z - 84 = 0$ or just $3z^2 + 8z - 7 = 0$. The roots $\alpha$ and $\overline{\alpha}$ of this equation are the fixed points, but we do not actually have to compute them since we know the quadratic form we want is $ax^2 + bx y + c y^2$ which in this example is just $3x^2 + 8xy - 7y^2$. As a check, we can compute the separator line of this form:
This provides a realization of the given periodic line as the separator line of a hyperbolic form. Any constant multiple of this form would also have the same separator line since we would just be multiplying all the labels along the line by the same constant.

We could have simplified the calculation slightly by noting that the periodic line we started with is taken to itself by a glide reflection that moves the line only half as far along itself as the translation $T$ that we used. This glide reflection is $T'(z) = (2z + 7)/(3z + 10)$ and it has the same fixed points as $T$ so we could use the equation $T'(z) = z$ instead of $T(z) = z$. This gives $2z + 7 = 3z^2 + 10z$ which simplifies immediately to $3z^2 + 8z - 7 = 0$.

4. Pell’s Equation

We encountered the equation $x^2 - dy^2 = 1$ briefly in Chapter 0. It is traditionally called Pell’s equation, and the similar equation $x^2 - dy^2 = -1$ is sometimes called Pell’s equation as well, or else the negative Pell’s equation. If $d$ is a square then the equations are not very interesting since in this case $d$ can be incorporated into the $y^2$ term, so one is looking at the equations $x^2 - y^2 = 1$ and $x^2 - y^2 = -1$, which have only the trivial solutions $(x, y) = (\pm 1, 0)$ for the first equation and $(x, y) = (0, \pm 1)$ for the second equation since these are the only cases when the difference between two squares is $\pm 1$. We will therefore assume that $d$ is not a square in what follows.

As an example let us look at the equation $x^2 - 19y^2 = 1$. We drew a portion of the periodic separator line for the form $x^2 - 19y^2$ earlier, and here it is again with some of the fractional labels $x/y$ shown as well.

Ignoring the label $741/170$ for the moment, the other fractional labels are the first few convergents for the continued fraction for $\sqrt{19}$ that we computed before, $4 + \frac{1}{2 + \frac{1}{3 + \frac{1}{1 + \frac{1}{2 + \frac{1}{8}}}}}$. These fractional labels are the labels on the vertices of the zigzag path in the infinite strip of triangles in the Farey diagram, which we can imagine being superimposed on the separator line in the figure. The fractional label we are most interested in is the $170/39$ because this is the label on a region where
the value of the form \( x^2 - 19y^2 \) is 1. This means exactly that \((x, y) = (170, 39)\) is a solution of \( x^2 - 19y^2 = 1 \). In terms of continued fractions, the fraction 170/39 is the value of the initial portion \( 4 + \frac{1}{2} + \frac{1}{3} + \frac{1}{1} + \frac{1}{2} \) of the continued fraction for \( \sqrt{19} \), with the final term of the period omitted.

Since the topograph of \( x^2 - 19y^2 \) is periodic along the separator line, there are infinitely many different solutions of \( x^2 - 19y^2 = 1 \) along the separator line. Going toward the left just gives the negatives \(-x/y\) of the fractions \( x/y \) to the right, changing the signs of \( x \) or \( y \), so it suffices to see what happens toward the right. One way to do this is to use the linear fractional transformation that gives the periodicity translation toward the right. This transformation sends the edge \( (1/0, 0/1) \) of the Farey diagram to the edge \( (170/39, 741/170) \). Here 741/170 is the value of the continued fraction \( 4 + \frac{1}{2} + \frac{1}{3} + \frac{1}{1} + \frac{1}{2} + \frac{1}{4} \) obtained from the continued fraction for \( \sqrt{19} \) by replacing the final number 8 in the period by one-half of its value, 4. The figure above shows why this is the right thing to do. We get an infinite sequence of larger and larger positive solutions of \( x^2 - 19y^2 = 1 \) by applying the periodicity transformation with matrix \( \begin{pmatrix} 170 & 741 \\ 39 & 170 \end{pmatrix} \) to the vector \((1, 0)\) repeatedly. For example,

\[
\begin{pmatrix} 170 & 741 \\ 39 & 170 \end{pmatrix} \begin{pmatrix} 170 \\ 39 \end{pmatrix} = \begin{pmatrix} 57799 \\ 13260 \end{pmatrix}
\]

so the next solution of \( x^2 - 19y^2 = 1 \) after \((170, 39)\) is \((57799, 13260)\), and we could compute more solutions if we wanted. Obviously they are getting large rather quickly.

The two 170’s in the matrix \( \begin{pmatrix} 170 & 741 \\ 39 & 170 \end{pmatrix} \) can hardly be just a coincidence. Notice also that the entry 741 factors as 19 \cdot 39 which hardly seems like it should be just a coincidence either. Let’s check that these numbers had to occur. In general, for the form \( x^2 - dy^2 \) let us suppose that we have found the first solution \((x, y) = (p, q)\) after \((1, 0)\) for Pell’s equation \( x^2 - dy^2 = 1 \), so \( p^2 - dq^2 = 1 \). Then based on the previous example we suspect that the periodicity transformation is the transformation

\[
\begin{pmatrix} p & dq \\ q & p \end{pmatrix} \begin{pmatrix} x \\ y \end{pmatrix} = \begin{pmatrix} px + dqy \\ qx + py \end{pmatrix}
\]

To check that this is correct the main thing to verify is that this transformation preserves the values of the quadratic form. When we plug in \((px + dqy, qx + dy)\) for \((x, y)\) in \( x^2 - dy^2 \) we get

\[
(px + dqy)^2 - d(qx + py)^2 = p^2x^2 + 2pdqx + d^2q^2y^2 - dx^2 - 2pdqx + dp^2y^2 = (p^2 - dq^2)x^2 - d(p^2 - dq^2)y^2 = x^2 - dy^2 \text{ since } p^2 - dq^2 = 1
\]

so the transformation \( \begin{pmatrix} p & dq \\ q & p \end{pmatrix} \) does preserve the values of the form. Also it takes \(1/0\) to \(p/q\), and its determinant is \( p^2 - dq^2 = 1 \), so it has to be the translation giving
the periodicity along the separator line. (We haven’t actually proved yet that periodic separator lines always exist for forms \(x^2 - dy^2\), but we will do this in the next chapter.)

Are there other solutions of \(x^2 - 19y^2 = 1\) besides the ones we have just described that occur along the separator line? The answer is No because we will see in the next chapter that as one moves away from the separator line in the topograph, the values of the quadratic form change in a monotonic fashion, steadily increasing toward \(+\infty\) as one moves upward above the separator line, and decreasing steadily toward \(-\infty\) as one moves downward below the separator line. Thus the value 1 occurs only along the separator line itself. Also we see that the value \(-1\) never occurs, which means that the equation \(x^2 - 19y^2 = -1\) has no integer solutions.

For an example where \(x^2 - dy^2 = -1\) does have solutions, let us look again at the earlier example of \(x^2 - 13y^2\).

The first positive solution \((x, y) = (p, q)\) of \(x^2 - 13y^2 = -1\) corresponds to the value \(-1\) in the middle of the figure. This is determined by the continued fraction \(p/q = 3 + \frac{1}{1 + \frac{1}{4}} = \frac{18}{5}\), so we have \((p, q) = (18, 5)\). The matrix \(\begin{pmatrix} p & dq \\ q & p \end{pmatrix}\) in this case is \(\begin{pmatrix} 18 & 65 \\ 5 & 18 \end{pmatrix}\) with determinant \(18^2 - 13 \cdot 5^2 = -1\) so this gives the glide-reflection along the periodic separator line taking \(1/0\) to \(18/5\) and \(0/1\) to \(65/18\). The smallest positive solution of \(x^2 - 13y^2 = +1\) is obtained by applying this glide-reflection to \((18, 5)\), which gives

\[
\begin{pmatrix} 18 & 65 \\ 5 & 18 \end{pmatrix} \begin{pmatrix} 18 \\ 5 \end{pmatrix} = \begin{pmatrix} 324 + 325 \\ 90 + 90 \end{pmatrix} = \begin{pmatrix} 649 \\ 180 \end{pmatrix}
\]

Repeated applications of the glide-reflection will give solutions of \(x^2 - 13y^2 = +1\) and \(x^2 - 13y^2 = -1\) alternately.

**Exercises**

1. Draw the topograph for the form \(Q(x, y) = 2x^2 + 5y^2\), showing all the values of \(Q(x, y) \leq 60\) in the topograph, with the associated fractional labels \(x/y\). If there is symmetry in the topograph, you only need to draw one half of the topograph and state that the other half is symmetric.

2. Do the same for the form \(Q(x, y) = 2x^2 + xy + 2y^2\), in this case displaying all values \(Q(x, y) \leq 40\) in the topograph.

3. Do the same for the form \(Q(x, y) = x^2 - y^2\), showing all the values between +30 and −30 in the topograph, but omitting the labels \(x/y\) this time.
4. For the form \( Q(x, y) = 2x^2 - xy + 3y^2 \) do the following:
(a) Draw the topograph, showing all the values \( Q(x, y) \leq 30 \) in the topograph, and including the labels \( x/y \).
(b) List all the values \( Q(x, y) \leq 30 \) in order, including the values when the pair \( (x, y) \) is not primitive.
(c) Find all the integer solutions of \( Q(x, y) = 24 \), both primitive and nonprimitive. (And don’t forget that quadratic forms always satisfy \( Q(x, y) = Q(-x, -y) \).)

5. Determine the periodic separator line in the topograph for each of the following quadratic forms (you do not need to include the fractional labels \( x/y \)):
(a) \( x^2 - 7y^2 \)  
(b) \( 3x^2 - 4y^2 \)  
(c) \( x^2 + xy - y^2 \)

6. Using your answers in the preceding problem, write down the continued fraction expansions for \( \sqrt{7} \), \( 2\sqrt{3}/3 \), and \( (1 + \sqrt{5})/2 \).

7. For the following quadratic forms, draw enough of the topograph, starting with the edge separating the 1/0 and 0/1 regions, to locate the periodic separator line, and include the separator line itself in your topograph:
(a) \( x^2 + 3xy + y^2 \)  
(b) \( 6x^2 + 18xy + 13y^2 \)  
(c) \( 37x^2 - 104xy + 73y^2 \)

8. Use a quadratic form to compute continued fractions for the following pairs of numbers:
(a) \( (3 + \sqrt{6})/2 \) and \( (3 - \sqrt{6})/2 \)  
(b) \( (11 + \sqrt{13})/6 \) and \( (11 - \sqrt{13})/6 \)  
(c) \( (14 + \sqrt{7})/9 \) and \( (14 - \sqrt{7})/9 \)

9. For the quadratic form \( x^2 - 14y^2 \) do the following things:
(a) Draw the separator line in the topograph and compute the continued fraction for \( \sqrt{14} \).
(b) Find the smallest positive integer solutions of \( x^2 - 14y^2 = 1 \) and \( x^2 - 14y^2 = -1 \), if these equations have integer solutions.
(c) Find the linear fractional transformation that gives the periodicity translation along the separator line and use this to find a second positive solution of \( x^2 - 14y^2 = 1 \).
(d) Determine the integers \( n \) with \(|n| \leq 12 \) such that the equation \( x^2 - 14y^2 = n \) has an integer solution. (Don’t forget the possibility that there could be solutions \( (x, y) \) that aren’t primitive.)

10. For the quadratic form \( x^2 - 29y^2 \) do the following things:
(a) Draw the separator line and compute the continued fraction for \( \sqrt{29} \).
(b) Find the smallest positive integer solution of \( x^2 - 29y^2 = -1 \).
(c) Find a glide-reflection symmetry of the separator line and use this to find the smallest positive integer solution of \( x^2 - 29y^2 = 1 \).

11. Compute the periodic separator line for the form \( x^2 - 43y^2 \) and use this to find the continued fraction for \( \sqrt{43} \).
We can divide quadratic forms \( Q(x, y) = ax^2 + bxy + cy^2 \) into four broad classes according to the signs of the values \( Q(x, y) \), where as always we restrict \( x \) and \( y \) to integers. We will always assume at least one of the coefficients \( a, b, c \) is nonzero, so \( Q \) is not identically zero, and we will always assume \( (x, y) \) is not \((0, 0)\). There are four possibilities:

(I) \( Q(x, y) \) takes on both positive and negative values but not 0. In this case we call \( Q \) a hyperbolic form.

(II) \( Q(x, y) \) takes on both positive and negative values and also 0. Then we call \( Q \) a 0-hyperbolic form.

(III) \( Q(x, y) \) takes on only positive values or only negative values. Then we call \( Q \) elliptic.

(IV) \( Q \) takes on the value 0 and either positive or negative values, but not both. Then \( Q \) is called parabolic.

The hyperbolic-elliptic-parabolic terminology is motivated in part by what the level curves \( ax^2 + bxy + cy^2 = k \) are, where we now allow \( x \) and \( y \) to take on all real values so that one gets actual curves. The level curves are hyperbolas in cases (I) and (II), and ellipses in case (III). In case (IV), however, the level curves are not parabolas as one might guess, but straight lines. Case (IV) will be the least interesting of the four cases.

There is an easy way to distinguish the four types of forms \( ax^2 + bxy + cy^2 \) in terms of their discriminants \( \Delta = b^2 - 4ac \). As we will show later in the chapter:

(I) If \( \Delta \) is positive but not a square then \( Q \) is hyperbolic.

(II) If \( \Delta \) is positive and a square then \( Q \) is 0-hyperbolic.

(III) If \( \Delta \) is negative then \( Q \) is elliptic.

(IV) If \( \Delta \) is zero then \( Q \) is parabolic.

Discriminants turn out to play a central role in the theory of quadratic forms. A natural question to ask is whether every integer occurs as the discriminant of some form, and this is easy to answer. For a form \( ax^2 + bxy + cy^2 \) we have \( \Delta = b^2 - 4ac \), and this is congruent to \( b^2 \mod 4 \). A square such as \( b^2 \) is always congruent to 0 or 1 mod 4, so the discriminant of a form is always congruent to 0 or 1 mod 4. Conversely, for every integer \( \Delta \) congruent to 0 or 1 mod 4 there exists a form whose discriminant is \( \Delta \) since:

\[
\begin{align*}
x^2 - ky^2 \text{ has discriminant } \Delta &= 4k \\
x^2 + xy - ky^2 \text{ has discriminant } \Delta &= 4k + 1
\end{align*}
\]

Here \( k \) can be positive, negative, or zero. The forms \( x^2 - ky^2 \) and \( x^2 + xy - ky^2 \) are called the principal quadratic forms of these discriminants.
We will analyze each of the four types of forms in turn, but before doing this let us make a few preliminary general comments.

In the arithmetic progression rule for labeling the four regions surrounding an edge of the topograph, we can label the edge by the common increment \( h = (q + r) - p = s - (q + r) \) as in the figure at the right. The edge can be oriented by an arrow showing the direction in which the progression increases by \( h \). Changing the sign of \( h \) corresponds to changing the orientation of the edge. In the special case that \( h \) happens to be 0 the orientation of the edge is irrelevant and can be omitted.

The values of the increment \( h \) along the boundary of a region in the topograph have the interesting property that they also form an arithmetic progression when all these edges are oriented in the same direction, and the amount by which \( h \) increases as we move from one edge to the next is \( 2p \) where \( p \) is the label on the region adjacent to all these edges:

We will call this property the *Second Arithmetic Progression Rule*. To see why it is true, start with the edge labeled \( h \) in the figure, with the adjacent regions labeled \( p \) and \( q \). The original Arithmetic Progression Rule then gives the value \( p + q + h \) in the next region to the right. From this we can deduce that the label on the edge between the regions labeled \( p \) and \( p + q + h \) must be \( h + 2p \) since this is the increment from \( q \) to \( p + (p + q + h) \). Thus the edge label increases by \( 2p \) when we move from one edge to the next edge to the right, so by repeated applications of this fact we see that we have an arithmetic progression of edge labels all along the border of the region labeled \( p \).

Another thing worth noting at this point is something that we will refer to as the *Monotonicity Property*: In the figure at the right, if the three labels \( p \), \( q \), and \( h \) adjacent to an edge are all positive, then so are the three labels for the next two edges in front of this edge, and the new labels are larger than the old labels. It follows that when one continues forward out this part of the topograph, all the labels become monotonically larger the farther one goes. Similarly, when the original three labels are negative, all the labels become larger and larger negative, by the same principle applied to the negative \(-Q(x, y)\) of the original form \(Q(x, y)\).

Next we have a very useful way to compute the discriminant of a form directly
Proposition 5.1. If an edge in the topograph of a form \( Q(x, y) \) is labeled \( h \) with adjacent regions labeled \( p \) and \( q \), then the discriminant of \( Q(x, y) \) is \( h^2 - 4pq \).

Proof: For the given form \( Q(x, y) = ax^2 + bxy + cy^2 \), the regions 1/0 and 0/1 in the topograph are labeled \( a \) and \( c \), and the edge in the topograph separating these two regions has \( h = b \) since the 1/1 region is labeled \( a + b + c \). So the statement of the proposition is correct for this edge. For other edges we proceed by induction, moving farther and farther out the tree. For the induction step suppose we have two adjacent edges labeled \( h \) and \( k \) as in the figure, and suppose inductively that the discriminant equals \( h^2 - 4pq \). We have \( r = p + q + h \), and from the second arithmetic progression rule we know that \( k = h + 2q \). Then we have \( k^2 - 4qr = (h + 2q)^2 - 4q(p + q + h) = h^2 + 4hq + 4q^2 - 4pq - 4q^2 - 4hq = h^2 - 4pq \), which means that the result holds for the edge labeled \( k \) as well.

1. The Four Types of Forms

Elliptic forms have fairly simple qualitative behavior so let us look at these forms first. Recall that we defined a form \( Q(x, y) \) to be elliptic if it takes on only positive or only negative values at all integer pairs \( (x, y) \neq (0,0) \). The positive and negative cases are equivalent since one can switch from one to the other just by putting a minus sign in front of \( Q \). Thus it suffices to consider the case that \( Q \) takes on only positive values, and we will assume we are in this case from now on. We will also generally assume when we look at topographs of elliptic forms that the orientations of the edges are chosen so as to give positive \( h \)-values unless we state otherwise.

For a positive elliptic form \( Q \) let \( p \) be the minimum positive value taken on by \( Q \), so \( Q(x, y) = p \) for some \( (x, y) \neq (0,0) \). Here \( (x, y) \) must be a primitive pair otherwise \( Q \) would take on a smaller positive value than \( p \). Thus there is a region in the topograph of \( Q \) with the label \( p \). All the edges having one endpoint at this region must be oriented away from the region, by the arithmetic progression rule and the assumption that \( p \) is the minimum value of \( Q \). The monotonicity property then implies that all edges farther away from the \( p \) region are also oriented away from the region, and the values of \( Q \) increase steadily as one moves away from the \( p \) region.

For the edges making up the border of the \( p \) region we know that the \( h \)-labels on these edges form an arithmetic progression with increment \( 2p \), provided that we temporarily re-orient these edges so that they all point in the same direction. If some edge bordering the \( p \) region has the label \( h = 0 \) then the topograph has the form...
shown in the first figure below, with the orientations on edges that give positive \( h \)-labels. An example of such a form is \( px^2 + qy^2 \). We call the 0-labeled edge a \textit{source edge} since all other edges are oriented away from this edge.

The other possibility is that no edge bordering the \( p \) region has label \( h = 0 \). Then since the labels on these edges form an arithmetic progression, there must be some vertex where the terms in the progression change sign. Thus when we orient the edges to give positive \( h \)-labels, all three edges meeting at this vertex will be oriented away from the vertex, as in the second figure above. We call this a \textit{source vertex} since all edges in the topograph are oriented away from this vertex.

If the three regions surrounding a source vertex are labeled \( p, q, r \) then the fact that the three edges leading from this vertex all point away from the vertex is equivalent to the three inequalities

\[
p < q + r \quad q < p + r \quad r < p + q
\]

These are called triangle inequalities since they are satisfied by the lengths of the three sides of any triangle. In the case of a source edge one of the inequalities becomes an equality, for example \( r = p + q \) in the earlier figure with a source edge.

As we know, any three integers \( p, q, r \) can be realized as the three labels surrounding a vertex in the topograph of some form. If these are positive integers satisfying the triangle inequalities then this vertex is the source vertex of an elliptic form since these inequalities imply that the three edges at this vertex are oriented away from the vertex, so the monotonicity property guarantees that all values of the form are positive. The situation for source edges is simpler since any two positive integers \( p \) and \( q \) determine an elliptic form with a source edge having adjacent regions labeled \( p \) and \( q \) as in the earlier figure.

Now let us move on to hyperbolic forms, whose topographs have quite a different appearance from the topographs of elliptic forms. Most notably, the topographs of hyperbolic forms always contain a periodic separator line of the sort that we saw in several of the examples in the previous chapter. Here is the general statement:
Theorem 5.2. In the topograph of a hyperbolic form the edges for which the two adjacent regions are labeled by numbers of opposite sign form a line which is infinite in both directions, and the topograph is periodic along this line, with other edges of the topograph leading off the line on both sides.

Proof: For a hyperbolic form $Q$ all regions in the topograph have labels that are either positive or negative, never zero, and there must exist two regions of opposite sign. By moving along a path in the topograph joining two such regions we will somewhere encounter two adjacent regions of opposite sign. Thus there must exist edges whose two adjacent regions have opposite sign. Let us call these edges separating edges.

At an end of a separating edge the value of $Q$ in the next region must be either positive or negative since $Q$ does not take the value 0:

$$
\begin{array}{c}
+ \\
- \\
+ \\
- \\
\end{array}
$$

This implies that exactly one of the two edges at each end of the first separating edge is also a separating edge. Repeating this argument, we see that each separating edge is part of a line of separating edges that is infinite in both directions, and the edges that lead off from this line are not separating edges.

The monotonicity property implies that as we move off this line of separating edges the values of $Q$ are steadily increasing through positive integers on the positive side and steadily decreasing through negative integers on the negative side. In particular this means that there are no other separating edges that are not on the initial separator line, so there is only one separator line.

It remains to prove that the topograph is periodic along the separator line. We can assume all the edges along the separator line are oriented in the same direction by changing the signs of the $h$ values if necessary. For an edge of the separator line labeled $h$ with adjacent regions labeled $p$ and $-q$ with $p > 0$ and $q > 0$, Proposition 5.1 says that $h^2 + 4pq$ is equal to the discriminant $\Delta$. The equation $\Delta = h^2 + 4pq$ with $p$ and $q$ positive implies that $\Delta$ is positive and furthermore that each of $|h|$, $p$, and $q$ is less than $\Delta$. Thus there are only finitely many possible values for $h$, $p$, and $q$ along the separator line since $\Delta$ is a constant depending only on $Q$. It follows that there are only finitely many possible combinations of values $h$, $p$, and $q$ at each edge on the separator line. Since the separator line is infinite, there must then be two edges on the line that have the same values of $h$, $p$, and $q$. Since the topograph is uniquely determined by the three labels $h$, $p$, $q$ at a single edge, the translation of the line along itself that takes one edge to another edge with the same three labels must preserve all the labels on the line. This shows that the separator line is periodic, and in fact the whole topograph is periodic.

There must be edges leading away from the separating line on both the positive
and the negative side, otherwise there would be just a single region on one side of the line and then the second arithmetic progression rule would say that the $h$ labels along the line formed an infinite arithmetic progression with nonzero increment $2p$ where $p$ is the label on the region in question. However, this would contradict the fact that these $h$ labels are periodic.

The qualitative behavior of the topograph of a hyperbolic form away from the separator line fits the pattern we have seen in examples. Since the separator line is periodic the whole topograph is periodic, consisting of repeating sequences of trees leading off from the separator line on each side, with monotonically increasing positive values of the form on each tree on the positive side of the separator line and monotonically decreasing negative values on the negative side, as a consequence of the monotonicity property.

The remaining types of forms to consider are parabolic forms and 0-hyperbolic forms. These turn out to be less interesting, and they play only a minor role in the theory of quadratic forms.

Parabolic and 0-hyperbolic forms are the forms whose topograph contains at least one region labeled 0. By the second arithmetic progression rule, each edge adjacent to a 0 region has the same label $h$, and from this it follows that the labels on the regions adjacent to the 0 region form an arithmetic progression.

Suppose first that $h = 0$. Then the topograph is as shown in the next figure so the form is parabolic, taking on only positive or only negative values away from the 0 region. Notice that the topograph is periodic along the 0 region since it consists of the same tree pattern repeated infinitely often.

A form with this topograph is $Q(x, y) = qx^2$.

The remaining case is that the label $h$ on the edges bordering a 0 region is nonzero. The arithmetic progression of values of $Q$ adjacent to the 0 region is then not constant, so it includes both positive and negative numbers, and hence $Q$ is
Chapter 5 Classification of Quadratic Forms

0-hyperbolic. If the arithmetic progression includes the value 0, this gives a second 0 region adjacent to the first one, and the topograph is as shown at the right. An example of a form with this topograph is \( Q(x, y) = qxy \), with the two 0 regions at \( x/y = 1/0 \) and \( 0/1 \).

If the arithmetic progression of values of \( Q \) adjacent to the 0 region does not include 0, there will be an edge separating the positive from the negative values in the progression. We can extend this separating edge to a line of separating edges as we did with hyperbolic forms, but the extension will eventually have to terminate with a second 0 region, otherwise the reasoning we used in the hyperbolic case would yield two edges along this line having the same \( h \) and the same positive and negative labels on the two adjacent regions, which would force the line to be periodic and hence extend infinitely far in both directions, which is impossible since it began at a 0 region at one end. Thus the topograph contains a finite separator line connecting two 0 regions. An example of such a form is \( Q(x, y) = qxy - py^2 = (qx - py)y \) which has the value 0 at \( x/y = 1/0 \) and at \( x/y = p/q \). Here we must have \(|q| > 1\) for the two 0 regions to be nonadjacent. The separator line must then follow the strip of triangles in the Farey diagram corresponding to the continued fraction for \( p/q \). For example, for \( p/q = 2/5 \) the topograph of the form \( 5xy - 2y^2 = (5x - 2y)y \) is the following:

\[
\begin{array}{c|c|c|c|c|c|c|c}
3q & 2q & q & 2q & 3q \\
0 & 0 & & & \\
-3q & -2q & -q & -2q & -3q
\end{array}
\]

This completes our description of what the topographs of the four types of forms look like. We can also deduce the characterization of each type in terms of the discriminant:

**Proposition 5.3.** The four types of forms are distinguished by their discriminants, which are negative for elliptic forms, positive nonsquares for hyperbolic forms, positive squares for 0-hyperbolic forms, and zero for parabolic forms.

**Proof:** Consider first an elliptic form \( Q \), which we may assume takes on only positive values since changing \( Q \) to \(-Q\) does not change the discriminant. The topograph of \( Q \) contains either a source vertex or a source edge. For a source edge with the label \( h = 0 \) separating regions with positive labels \( p \) and \( q \) the discriminant is \( \Delta = h^2 - 4pq = -4pq \), which is negative. For a source vertex with adjacent regions having
positive labels $p, q, r$, the edge between the $p$ and $q$ regions is labeled $h = p + q - r$ so we have

$$\Delta = h^2 - 4pq = (p + q - r)^2 - 4pq = p^2 + q^2 + r^2 - 2pq - 2pr - 2qr = p(p - q - r) + q(q - p - r) + r(r - p - q)$$

In the last line the three quantities in parentheses are negative by the triangle inequalities, so $\Delta$ is negative.

For a parabolic form the topograph contains a region labeled 0 bordered by edges labeled 0, so $\Delta = h^2 - 4pq = 0$. A 0-hyperbolic form has a region labeled 0 bordered by edges all having the same label $h \neq 0$ so $\Delta = h^2$, a positive square.

For an edge in the separator line for a hyperbolic form the adjacent regions have labels $p$ and $-q$ with $p$ and $q$ positive so $\Delta = h^2 + 4pq$ is positive. To see that $\Delta$ is not a square, suppose the form is $ax^2 + bxy + cy^2$. Here $a$ must be nonzero, otherwise the form would have the value 0 at $(x, y) = (1, 0)$, which is impossible for a hyperbolic form. If the discriminant was a square then the equation $az^2 + bz + c = 0$ would have a rational root $z = x/y$ with $y \neq 0$ by the familiar quadratic formula $z = \frac{-b \pm \sqrt{b^2 - 4ac}}{2a}$. Thus we would have $a(x/y)^2 + b(x/y) + c = 0$ and hence $ax^2 + bxy + cy^2 = 0$ so the form would have the value 0 at a pair $(x, y)$ with $y \neq 0$, which is again impossible for a hyperbolic form.

The presence or absence of periodicity in a topograph has the following consequence:

**Proposition 5.4.** If an equation $Q(x, y) = n$ with $n \neq 0$ has one integer solution $(x, y)$ then it has infinitely many integer solutions when $Q$ is hyperbolic or parabolic, but only finitely many integer solutions when $Q$ is elliptic or 0-hyperbolic.

**Proof:** Consider first the hyperbolic and parabolic cases. Suppose $(x, y)$ is a solution of $Q(x, y) = n$. If $(x, y)$ is a primitive pair, then $n$ appears in the topograph of $Q$ so by periodicity it appears infinitely often, giving infinitely many solutions of $Q(x, y) = n$. If there is nonprimitive solution $(x, y)$ then it is $d$ times a primitive pair $(x', y')$ with $Q(x', y') = n/d^2$. The latter equation has infinitely many solutions $(x', y')$ by what we just showed, hence $Q(x, y) = n$ has infinitely many solutions $(x, y) = (dx', dy')$.

For elliptic and 0-hyperbolic forms there is no periodicity and the monotonicity property implies that each number appears in the topograph at most a finite number of times. Thus $Q(x, y) = n$ can have only finitely many primitive solutions. If it had infinitely many nonprimitive solutions, these would yield infinitely many primitive solutions of equations $Q(x, y) = m$ for a set of numbers $m < n$. However, this is impossible since by induction each equation $Q(x, y) = m$ for a fixed $m < n$ can
have only finitely many primitive solutions so there can only be finitely many primitive solutions for all \( m < n \) together.

\[ \square \]

### 2. Equivalence of Forms

In the pictures of topographs we have drawn, we often omit the fractional labels \( \frac{x}{y} \) for the regions in the topograph since the more important information is often just the values \( Q(x, y) \) of the form. This leads to the idea of considering two quadratic forms to be equivalent if their topographs “look the same” when the labels \( \frac{x}{y} \) are disregarded. For a precise definition, one can say that quadratic forms \( Q_1 \) and \( Q_2 \) are *equivalent* if there is a vertex \( v_1 \) in the topograph of \( Q_1 \) and a vertex \( v_2 \) in the topograph of \( Q_2 \) such that the values of \( Q_1 \) in the three regions surrounding \( v_1 \) are equal to the values of \( Q_2 \) in the three regions surrounding \( v_2 \). For example if the values at \( v_1 \) are 2, 2, 3 then the values at \( v_2 \) should also be 2, 2, 3, in any order, but 2, 3, 3 is regarded as different from 2, 2, 3. Since the three values around a vertex determine all the other values in a topograph, having the same values at one vertex guarantees that the topographs look the same everywhere, if the labels \( \frac{x}{y} \) are omitted.

An alternative definition of equivalence of forms would be to say that two forms are equivalent if there is a linear fractional transformation in \( LF(\mathbb{Z}) \) that takes the topograph of one form to the topograph of the other form. This is really the same as the first definition since there is a vertex of the topograph in the center of each triangle of the Farey diagram and we know that elements of \( LF(\mathbb{Z}) \) are determined by where they send a triangle, so if two topographs each have a vertex surrounded by the same triple of numbers, there is an element of \( LF(\mathbb{Z}) \) taking one topograph to the other, and conversely.

A topograph and its mirror image correspond to equivalent forms since the mirror image topograph has the same three labels around each vertex as at the corresponding vertex of the original topograph. For example, switching the variables \( x \) and \( y \) reflects the circular Farey diagram across its vertical axis and hence reflects the topograph of a form \( Q(x, y) \) to the topograph of the equivalent form \( Q(y, x) \). As another example, the forms \( ax^2 + bxy + cy^2 \) and \( ax^2 - bxy + cy^2 \) are always equivalent since they are related by changing \( (x, y) \) to \( (-x, y) \), reflecting the Farey diagram across its horizontal axis, with a corresponding reflection of the topograph.

Equivalent forms have the same discriminant since the discriminant of a form is determined by the three numbers surrounding any vertex, as these three numbers determine the numbers \( p, q, h \) at each edge abutting the vertex and the discriminant is \( h^2 - 4pq \) for any of these edges. Our next goal will be to see how to compute all the different equivalence classes of forms of a given discriminant. The method for doing this will depend on which of the four types of forms we are dealing with.
Let us look at elliptic forms first to see how to determine all the different equivalence classes for a given discriminant in this case. As usual it suffices to consider only the forms with positive values. At a source vertex or edge in the topograph of a positive elliptic form \( Q \) let the smaller two of the three adjacent values of \( Q \) be \( a \) and \( c \) with \( a \leq c \), and let the edge between them be labeled \( h \geq 0 \). For a source edge we have \( h = 0 \) and for a source vertex we have \( h > 0 \). The third of the three smallest values of \( Q \) is then \( a + c - h \) in either case. The form \( Q \) is equivalent to the form \( ax^2 + hxy + cy^2 \) which has the values \( a \), \( c \), and \( a + h + c \) for \((x, y) = (1, 0), (0, 1)\), and \((1, 1)\). Since \( a \) and \( c \) are the smallest values of \( Q \) we have \( a \leq c \leq a + c - h \), and the latter inequality is equivalent to \( h \leq a \). Summarizing, we have the inequalities \( 0 \leq h \leq a \leq c \).

Thus every positive elliptic form is equivalent to a form \( ax^2 + hxy + cy^2 \) with \( 0 \leq h \leq a \leq c \). An elliptic form satisfying these conditions is called reduced. Two different reduced elliptic forms with the same discriminant are never equivalent since \( a \) and \( c \) are the labels on the two regions in the topograph where the form takes its smallest values, and \( h \) is determined by \( a \), \( c \), and \( \Delta \) via the formula \( \Delta = h^2 - 4ac \) since we assume \( h \geq 0 \).

To avoid dealing with negative numbers let us set \( \Delta = -D \) with \( D > 0 \), so the discriminant equation becomes \( D = 4ac - h^2 \). To find all equivalence classes of forms of discriminant \(-D\) we therefore need to find all solutions of the equation

\[
4ac = h^2 + D \quad \text{with} \quad 0 \leq h \leq a \leq c
\]

This equation implies that \( h \) must have the same parity as \( D \), and we can bound the choices for \( h \) by the inequalities \( 4h^2 \leq 4a^2 \leq 4ac = D + h^2 \) which imply \( 3h^2 \leq D \), or \( h^2 \leq D/3 \). This limits \( h \) to a finite number of possibilities, and for each of these values of \( h \) we just need to find all of the finitely many factorizations of \( h^2 + D \) as \( 4ac \). In particular this shows that there are just finitely many equivalence classes of elliptic forms of a given discriminant.

As an example consider the case \( \Delta = -260 \), so \( D = 260 \). Since \( \Delta \) is even, so is \( h \), and we must have \( h^2 \leq 260/3 \) so \( h \) must be 0, 2, 4, 6, or 8. The corresponding values of \( a \) and \( c \) that are possible can then be computed from the equation \( 4ac = 260 + h^2 \), always keeping in mind the requirement that \( h \leq a \leq c \). The possibilities are shown in the following table:

<table>
<thead>
<tr>
<th>( h )</th>
<th>( ac )</th>
<th>((a, c))</th>
</tr>
</thead>
<tbody>
<tr>
<td>0</td>
<td>65</td>
<td>((1, 65), (5, 13))</td>
</tr>
<tr>
<td>2</td>
<td>66</td>
<td>((2, 33), (3, 22), (6, 11))</td>
</tr>
<tr>
<td>4</td>
<td>69</td>
<td>—</td>
</tr>
<tr>
<td>6</td>
<td>74</td>
<td>—</td>
</tr>
<tr>
<td>8</td>
<td>81</td>
<td>((9, 9))</td>
</tr>
</tbody>
</table>

As a side comment, note that the values of \( ac \) increase successively by 1, 3, 5, 7, \( \cdots \).
This always happens when $\Delta$ is even and the $h$ values are $0, 2, 4, 6, \cdots$. For odd $\Delta$ the values of $h$ are $1, 3, 5, 7, \cdots$ and the increments for $ac$ are $2, 4, 6, 8, \cdots$. (Let it be an exercise for the reader to figure out why these statements are true.)

From the table we see that every positive elliptic form of discriminant $-260$ is equivalent to one of the six reduced forms $x^2 + 65y^2$, $5x^2 + 13y^2$, $2x^2 + 2xy + 33y^2$, $3x^2 + 2xy + 22y^2$, $6x^2 + 2xy + 11y^2$, or $9x^2 + 8xy + 9y^2$, and no two of these reduced forms are equivalent to each other. Here are small parts of the topographs of these forms:

\[
\begin{align*}
x^2 + 65y^2 & \quad 2x^2 + 2xy + 33y^2 & \quad 6x^2 + 2xy + 11y^2 \\
\begin{array}{ccc}
66 & 1 & 65 \\
16 & 5 & 13 \\
66 & 18 & 33 \\
\end{array} & \\
\begin{array}{ccc}
18 & 3 & 33 \\
2 & 37 & 27 \\
6 & 19 & 11 \\
\end{array} & \\
\begin{array}{ccc}
9 & 9 & 26 \\
9 & 10 & 15 \\
\end{array}
\end{align*}
\]

In the first two topographs the central edge is a source edge, and in the last four topographs the lower vertex is a source vertex.

One might wonder what would happen if we continued the table with larger values of $h$ not satisfying $h^2 \leq 260/3$. For example for $h = 10$ we would have $ac = 90$ so the condition $a \leq c$ would force $a$ to be 9 or less, violating the condition $h \leq a$. Larger values of $h$ would run into similar difficulties. The condition $h^2 \leq |\Delta|/3$ saves one the trouble of trying larger values of $h$.

Next we consider hyperbolic forms of a given discriminant $\Delta > 0$. The topograph of a hyperbolic form has a separator line, so for each edge in the separator line we have the edge label $h$ with the adjacent regions labeled $p$ and $-q$ for $p > 0$ and $q > 0$. We can assume $h \geq 0$ by reorienting the edge if necessary. The discriminant equation is $\Delta = h^2 + 4pq$. Since $p$ and $q$ are positive this implies $h^2 < \Delta$ so there are only finitely many possibilities for $h$ along the separator lines of forms of the given discriminant $\Delta$. For each $h$ we then look at the factorizations $\Delta - h^2 = 4pq$. There can be only finitely many of these, so this means there are just finitely many possible combinations of labels $h, p, -q$ and hence only finitely many possible separator lines. Thus the number of equivalence classes of hyperbolic forms of a given discriminant is finite.

As an example, let us determine all the quadratic forms of discriminant 60, up to equivalence. Two obvious forms of discriminant 60 are $x^2 - 15y^2$ and $3x^2 - 5y^2$, whose separator lines consist of periodic repetitions of the following two patterns:
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From the topographs it is apparent that these two forms are not equivalent, and also that the negatives of these two forms, \(-x^2 + 15y^2\) and \(-3x^2 + 5y^2\), give two more inequivalent forms, for a total of four equivalence classes so far. To see whether there are others we use the formula \(\Delta = 60 = h^2 + 4pq\) relating the values \(p\) and \(-q\) along an edge labeled \(h\) in the separator line, with \(p > 0\) and \(q > 0\). The various possibilities are listed in the table below. The equation \(\Delta = h^2 + 4pq\) implies that \(h\) and \(\Delta\) must have the same parity, just as in the elliptic case.

<table>
<thead>
<tr>
<th>(h)</th>
<th>(pq)</th>
<th>((p, q))</th>
</tr>
</thead>
<tbody>
<tr>
<td>0</td>
<td>15</td>
<td>(1, 15), (3, 5), (5, 3), (15, 1)</td>
</tr>
<tr>
<td>2</td>
<td>14</td>
<td>(1, 14), (2, 7), (7, 2), (14, 1)</td>
</tr>
<tr>
<td>4</td>
<td>11</td>
<td>(1, 11), (11, 1)</td>
</tr>
<tr>
<td>6</td>
<td>6</td>
<td>(1, 6), (2, 3), (3, 2), (6, 1)</td>
</tr>
</tbody>
</table>

Each pair of values for \((p, q)\) in the table occurs at some edge along the separator line in one of the two topographs shown above or the negatives of these topographs. Hence every form of discriminant 60 is equivalent to one of these four. If it had not been true that all the possibilities in the table occurred in the topographs of the forms we started with, we could have used these other possibilities for \(h\), \(p\), and \(q\) to generate new topographs and hence new forms, eventually exhausting all the finitely many possibilities.

The procedure in this example works for all hyperbolic forms. One makes a list of all the positive integer solutions of \(\Delta = h^2 + 4pq\), then one constructs separator lines that correspond exactly to the different equivalence classes of forms of discriminant \(\Delta\). Each solution \((h, p, q)\) gives a form \(px^2 + hxy − qy^2\). These are organized into "cycles" corresponding to the pairs \((p, −q)\) occurring along one of the periodic separator lines. Thus in the preceding example with \(\Delta = 60\) the 14 pairs \((p, q)\) in the table give rise to the four cycles along the four different separator lines.

Note that a hyperbolic form \(ax^2 + bxy + cy^2\) belongs to one of these cycles for the discriminant \(\Delta = b^2 − 4ac\) exactly when \(a > 0\) and \(c < 0\) since \(a\) and \(c\) are the numbers \(p\) and \(-q\) lying on opposite sides of an edge of the separator line, when \((x, y) = (1, 0)\) and \((0, 1)\).

If we superimpose the separator line of a hyperbolic form on the associated infinite strip in the Farey diagram, we see that the forms within a cycle correspond to the edges of the Farey diagram that lie in the strip and join one border of the strip to the other. For example, for the form \(3x^2 − 5y^2\) we obtain the following picture, with fans of two triangles alternating with fans of three triangles:
The number of forms within a given cycle can be fairly large in general. The situation can be improved somewhat by considering only the “most important” forms in the cycle, namely the forms that correspond to those edges in the strip that separate pairs of adjacent fans, indicated by heavier lines in the figure. In terms of the topograph itself these are the edges in the separator line whose two endpoints have edges leading away from the separator line on opposite sides. The forms corresponding to these edges are traditionally called the reduced forms within the given equivalence class. In the example of discriminant 60 these are the forms with \((p,q) = (1,6), (6,1), (3,2),\) and \((2,3).\) These are the forms \(x^2+6xy−6y^2, 6x^2+6xy−y^2, 3x^2+6xy−2y^2,\) and \(2x^2 + 6xy − 3y^2.\) In this example there is just one reduced form for each cycle, but in more complicated examples there can be any number of reduced forms in a cycle. Note that the reduced forms do not necessarily give the simplest-looking forms, which in this example were the original forms \(x^2 − 15y^2\) and \(3x^2 − 5y^2\) along with their negatives, or alternatively \(15x^2 − y^2\) and \(5x^2 − 3y^2.\)

For 0-hyperbolic forms it is rather easy to determine all the equivalence classes of forms of a fixed discriminant. As we saw in our initial discussion of 0-hyperbolic forms, their topographs contain two regions labeled 0, and the labels on the regions adjacent to each 0-region form an arithmetic progression with increment given by the label on the edges bordering the 0 region. Previously we called this label \(h\) but now let us change notation and call it \(q.\) We may assume \(q\) is positive by re-orienting the edges if necessary. The discriminant is \(\Delta = q^2\) so both 0 regions must have the same edge label \(q.\) Either one of the two arithmetic progressions determines the form up to equivalence since two successive terms in the progression together with the 0 in the adjacent region give the three values of the form around a vertex in the topograph.

The form \(qxy − py^2\) has discriminant \(q^2\) and has \(-p\) as one term of the arithmetic progression adjacent to the 0-region \(x/y = 1/0,\) namely in the region \(x/y = 0/1.\) Thus every 0-hyperbolic form of discriminant \(q^2\) is equivalent to one of these forms \(qxy − py^2.\) Arithmetic progressions with increment \(q\) can be thought of as congruence classes mod \(q,\) so only the mod \(q\) value of \(p\) affects the arithmetic progression and hence we may assume \(0 ≤ p < q.\) The number of equivalence classes of 0-hyperbolic forms of discriminant \(q^2\) is therefore at most \(q,\) the number of congruence classes mod \(q.\) However, the number of equivalence classes could be smaller
since each form has two 0 regions and hence two arithmetic progressions, which could be the same or different. Since either arithmetic progression determines the form, if the two progressions are the same then the topograph must have a mirror symmetry interchanging the two 0 regions. This always happens for example if the two 0 regions touch, which is the case \( p = 0 \) so the form is \( qxy \) and the mirror symmetry just interchanges \( x \) and \( y \). If we let \( r \) denote the number of forms \( qxy - py^2 \) without mirror symmetry then the number of equivalence classes of 0-hyperbolic forms of discriminant \( q^2 \) is \( q - r \) since each form without mirror symmetry has two different arithmetic progressions giving the same form.

For parabolic forms it is easy to describe what all the different equivalence classes are since we have seen exactly what their topographs look like: There is a single region labeled 0 and all the regions adjacent to this have the same label \( q \), which can be any nonzero integer, positive or negative. The integer \( q \) thus determines the equivalence class, so there is one equivalence class of parabolic forms for each nonzero integer \( q \), with \( qx^2 \) being one form in this equivalence class. Parabolic forms all have discriminant zero, so in this case there are infinitely many different equivalence classes with the same discriminant.

We have now shown how to compute all the equivalence classes of forms of a given discriminant for each of the four types of forms. In particular we have proved the following general fact:

**Theorem 5.6.** There are only a finite number of equivalence classes of forms with a given nonzero discriminant.

### 3. The Class Number

When considering the various equivalence classes of forms of a given discriminant there are further refinements that turn out to be very useful. The first involves forms whose topographs are mirror images of each other. According to the definition we have given, two such forms are regarded as equivalent. However, there is a more refined notion of equivalence in which two forms are considered equivalent only if there is an orientation-preserving transformation in \( LF(\mathbb{Z}) \) taking the topograph of one form to the topograph of the other. In this case the forms are called *properly equivalent*.

To illustrate the distinction between equivalence and proper equivalence, let us look at the earlier example of discriminant \( \Delta = -260 \) where we saw that there were six equivalence classes of forms:
In the first two topographs the central edge is a source edge and in the other four the lower vertex is a source vertex. Whenever there is a source edge the topograph has mirror symmetry across a line perpendicular to the source edge. When there is a source vertex there is mirror symmetry only when at least two of the three surrounding values of the form are equal, as in the third and sixth topographs above, but not the fourth or fifth topographs. Thus the mirror images of the fourth and fifth topographs correspond to two more quadratic forms which are not equivalent to them under any orientation-preserving transformation. To obtain an explicit formula for the mirror image forms we can interchange the coefficients \( a \) and \( c \) in \( ax^2 + bxy + cy^2 \), which corresponds to interchanging \( x \) and \( y \), reflecting the topograph across a vertical line. Alternatively we could change the sign of \( b \), corresponding to changing the sign of either \( x \) or \( y \) and thus reflecting the topograph across a horizontal line.

The net result of all this is that with the more refined notion of proper equivalence there are eight proper equivalence classes of forms of discriminant \(-260\).

For a general discriminant \( \Delta \) each equivalence class of forms of discriminant \( \Delta \) gives rise to two proper equivalence classes except when the class contains forms with mirror symmetry, in which case equivalence and proper equivalence amount to the same thing since every orientation-reversing equivalence can be converted into an orientation-preserving equivalence by composing with a mirror reflection. Here we are using the fact that the only linear fractional transformations that take a topograph to itself and reverse orientation are mirror reflections, as will be shown in the next section when we study symmetries of topographs in more detail.

Multiplying a form by an integer \( d > 1 \) does not change its essential features in any significant way, so it is reasonable when classifying forms to restrict attention just to primitive forms, the forms that are not proper multiples of other forms. In other words, one considers only the forms \( ax^2 + bxy + cy^2 \) for which \( a \), \( b \), and \( c \) have no common divisor greater than 1. The primitivity of a form is detectable just from the numbers appearing in its topograph since all the numbers in the topograph of a nonprimitive form are divisible by some number \( d > 1 \), and conversely if all numbers in the topograph of a form \( ax^2 + bxy + cy^2 \) are divisible by \( d \) then in particular \( a \), \( c \), and \( a + b + c \), the values at \((1,0)\), \((0,1)\), and \((1,1)\), are divisible by \( d \) which implies that \( b \) is also divisible by \( d \) so the whole form is divisible by \( d \). Thus primitivity
is a property of equivalence classes of forms. Multiplying a form by \( d \) multiplies its discriminant by \( d^2 \), so nonprimitive forms of discriminant \( \Delta \) exist exactly when \( \Delta \) is a square times another discriminant. For example, when \( \Delta = -12 = 4(-3) \) one has the primitive form \( x^2 + 3y^2 \) as well as the nonprimitive form \( 2x^2 + 2xy + 2y^2 \) which is twice the form \( x^2 + xy + y^2 \) of discriminant \(-3\).

The number of proper equivalence classes of primitive forms of a given discriminant is called the class number for that discriminant, where in the case of elliptic forms one considers only the forms with positive values. The traditional notation for the class number for discriminant \( \Delta \) is \( h_\Delta \). (This \( h \) has nothing to do with the \( h \) labels on edges in topographs.)

Since we have an algorithm for computing the finite set of equivalence classes of forms of a given discriminant, this leads to an algorithm for computing class numbers. When computing the table of triples \((h, a, c)\) for elliptic forms or \((h, p, q)\) for hyperbolic forms we omit the nonprimitive triples since these correspond to nonprimitive forms. Then we determine which of the remaining forms have mirror symmetry. For elliptic forms these are the cases when one or more of the inequalities \( 0 \leq h \leq a \leq c \) is an equality, as we will see in the next section. For hyperbolic forms mirror symmetries can be detected in the separator line. Forms with mirror symmetry count once when computing the class number and forms without mirror symmetry count twice. However, just having an algorithm to compute the class number \( h_\Delta \) does not make it transparent how \( h_\Delta \) depends on \( \Delta \), and indeed this is a very difficult question which is still only partially understood.

Of special interest are the discriminants for which all forms are primitive. These are called fundamental discriminants. Thus a fundamental discriminant is one which is not a square times a smaller discriminant. For example 8 is a fundamental discriminant even though it is divisible by a square, 4, since the other factor 2 is not the discriminant of any form, as it is not congruent to 0 or 1 mod 4. Technically 1 is a fundamental discriminant according to our definition, but we will exclude this trivial case. Thus fundamental discriminants are never squares, so fundamental discriminants appear only for elliptic and hyperbolic forms. With 1 excluded it is easy to check that the fundamental discriminants \( \Delta \) with \(|\Delta| < 40 \) are 5, 8, 12, 13, 17, 20, 21, 24, 28, 29, 33, 37 and \(-3, -4, -7, -8, -11, -15, -19, -20, -23, -24, -31, -35, -39\).

It is not hard to characterize precisely the discriminants \( \Delta \) that are fundamental. First write \( \Delta = 2^k n \) with \( k \geq 0 \) and \( n \) odd, possibly negative. If any odd square divides \( n \) then we can factor this out of \( \Delta \) and still get a discriminant since odd squares are congruent to 1 mod 4 so multiplying by an odd square does not affect whether a number is 0 or 1 mod 4. The exponent \( k \) in \( 2^k \) can never be 1 since this would imply \( \Delta \equiv 2 \mod 4 \). If \( k \geq 4 \) we can factor powers of 4 out of \( \Delta \) until we have \( k \) equal to 2 or 3 and still have a discriminant. If \( k = 3 \) we cannot factor a 4 out of
Δ since this would give the excluded case \( k = 1 \). If \( k = 2 \) we can factor \( 4 = 2^k \) out of \( \Delta \) exactly when \( n \equiv 1 \mod 4 \). Finally when \( k = 0 \) we have \( \Delta = n \) so we must have \( n \equiv 1 \mod 4 \). Thus fundamental discriminants other than \(-4\) and \(\pm 8\) are of three types:

1. \( \Delta = n \) with \(|n|\) a product of distinct odd primes and with \( n \equiv 1 \mod 4 \).
2. \( \Delta = 4n \) with \(|n|\) a product of distinct odd primes and with \( n \equiv 3 \mod 4 \).
3. \( \Delta = 8n \) with \(|n|\) a product of distinct odd primes.

Every nonsquare discriminant can be factored as \( \Delta = d^2 \Delta' \) where \( \Delta' \) is a fundamental discriminant uniquely determined by \( \Delta \). The number \( d \), which we can assume is positive, is therefore also uniquely determined by \( \Delta \) and is called the conductor of \( \Delta \). Fundamental discriminants are those whose conductor is 1. Conductors play a role when one studies the deeper properties of forms, as we will see in later chapters.

For a nonfundamental discriminant \( d^2 \Delta \) there is a nice relationship between \( h_{d^2 \Delta} \) and \( h_\Delta \) which says that \( h_{d^2 \Delta} \) is a multiple of \( h_\Delta \) and there is a fairly simple formula for what this multiple is. This means that the determination of class numbers reduces largely to the case of fundamental discriminants.

The question of which discriminants have class number 1 has been much studied. This amounts to finding the discriminants for which all primitive forms are equivalent since if all primitive forms are equivalent, they are all equivalent to the principal form which has mirror symmetry so they are all properly equivalent to the principal form.

For elliptic forms the following nine fundamental discriminants have class number 1:

\[
\Delta = -3, -4, -7, -8, -11, -19, -43, -67, -163
\]

In addition there are four more which are not fundamental: \(-12, -16, -27, -28\). It was conjectured by Gauss around 1800 that there are no other negative discriminants of class number 1. Over a century later in the 1930s it was shown that there is at most one more, and then in the 1950s and 60s Gauss’s conjecture was finally proved completely.

Another result from the 1930s is that for each number \( n \) there are only finitely many negative discriminants with class number \( n \). Finding what these discriminants are is a difficult problem, however, and so far this has been done only in the range \( n \leq 100 \).

The situation for positive discriminants with class number 1 is not as well understood. Computations show that there are a large number of positive fundamental discriminants with class number 1, and it seems likely that there are in fact infinitely many. However, this has not been proved and remains one of the most basic unsolved problems about quadratic forms. If one allows nonfundamental discriminants then it is known that there are infinitely many with \( h_\Delta = 1 \), including for example the discriminants \( \Delta = 2^{2k+1} \) for \( k \geq 1 \) and \( \Delta = 5^{2k+1} \) for \( k \geq 0 \).
Returning to the nine negative fundamental discriminants of class number 1, it is easy to check in each case that all forms are equivalent. For example when $\Delta = -163$ we must have $h$ odd with $h^2 \leq 163/3$ so the only possibilities are $h = 1, 3, 5, 7$. From the equation $4ac = 163 + h^2$ the corresponding values of $ac$ are 41, 43, 47, 53 which all happen to be primes, and since $a \leq c$ this forces $a$ to be 1 in each case. But since $h \leq a$ this means $h$ must be 1, and we obtain the single quadratic form $x^2 + xy + 41y^2$.

The corresponding polynomial $x^2 + x + 41$ has a curious property discovered by Euler: For each $x = 0, 1, 2, 3, \ldots, 39$ the value of $x^2 + x + 41$ is a prime number. Here are these forty primes:

$$
$$

Notice that the successive differences between these numbers are 2, 4, 8, 10, $\ldots$. The next number in the sequence after 1601 would be 1681 = $41^2$, not a prime. (Write $x^2 + x + 41$ as $x(x + 1) + 41$ to see why $x = 40$ must give a nonprime value.) A similar thing happens for the other negative fundamental discriminants of class number 1. The nontrivial cases are listed in the table below, where $D = -\Delta$.

<table>
<thead>
<tr>
<th>$D$</th>
<th>$x^2 + x + 2$</th>
<th>2</th>
</tr>
</thead>
<tbody>
<tr>
<td>7</td>
<td>$x^2 + x + 3$</td>
<td>3</td>
</tr>
<tr>
<td>11</td>
<td>$x^2 + x + 5$</td>
<td>5, 7, 11, 17</td>
</tr>
<tr>
<td>19</td>
<td>$x^2 + x + 11$</td>
<td>11, 13, 17, 23, 31, 41, 53, 67, 83, 101</td>
</tr>
<tr>
<td>43</td>
<td>$x^2 + x + 17$</td>
<td>17, 19, 23, 29, 37, 47, 59, 73, 89, 107, 127, 149, 173, 199, 227, 257</td>
</tr>
</tbody>
</table>

It is curious that these lists, including the one for $x^2 + x + 41$, account for all primes less than 100 except 79.

Suppose one asks about the next 40 values of $x^2 + x + 41$ after the value $41^2$ when $x = 40$. The next value, when $x = 41$, is 1763 = 41 · 43, also not a prime. After this the next two values are primes, then comes 2021 = 43 · 47, then four primes, then 2491 = 47 · 53, then six primes, then 3233 = 53 · 61, then eight primes, then 4331 = 61 · 71, then ten primes, then 5893 = 71 · 83. This last number was for $x = 76$, and the next four values are prime as well for $x = 77, 78, 79, 80$, completing the second forty values. But then the pattern breaks down when $x = 81$ where one gets the value 6683 = 41 · 163. Thus, before the breakdown, not only were we getting sequences of 2, 4, 6, 8, 10 primes but the non-prime values were the products of two successive terms in the original sequence of prime values 41, 43, 47, 53, 61, $\ldots$. All this seems quite surprising, even if the nice patterns do not continue forever. A partial explanation can be found in the fact that the polynomial $P(x) = x^2 + x + 41$ satisfies the identity $P(40 + n^2) = P(n - 1)P(n)$ as one can easily check, so when $n = 1, 2, 3, \ldots$ we get $P(41) = P(0)P(1) = 41 · 43$, $P(44) = 43 · 47$, $P(49) = 47 · 53$, $\ldots$. 

\[ P(56) = 53 \cdot 61, \text{ etc. However this does not explain why the intervening values of } P(x) \text{ should be prime. For the first forty values of } P(x) \text{ satisfactory explanations are known for the appearance of so many primes and for why this is related to the class number being 1.} \]

4. Symmetries

We have observed that some topographs are symmetric in various ways. To give a precise meaning to this term, let us say that a symmetry of a form \( Q \) (or its topograph) is a transformation \( T \) in \( LF(\mathbb{Z}) \) that leaves all the values of \( Q \) unchanged, so \( Q(T(x, y)) = Q(x, y) \) for all pairs \( (x, y) \). For example, every hyperbolic form has a periodic separator line, which means there is a symmetry that translates the separator line along itself. If \( T \) is the symmetry translating by one period in either direction, then all the positive and negative powers of \( T \) are also translational symmetries. Strictly speaking, the identity transformation is always a symmetry but we will sometimes ignore this trivial symmetry.

Some hyperbolic forms also have mirror symmetry, where the symmetry is reflection across a line perpendicular to the separator line. This reflector line could contain one of the edges leading off the separator line, or it could be halfway between two consecutive edges leading off the separator line on the same side. Both kinds of symmetry occur along the separator line of the form \( x^2 - 19y^2 \), for example:

\[ \begin{array}{cccccccc}
1 & 6 & 5 & 9 & 5 & 6 & 1 \\
\end{array} \]

Elliptic forms can have mirror symmetries as well, as we saw in the earlier example \( \Delta = -260 \) where two topographs had mirror symmetry across a line perpendicular to an edge and two had symmetry across a line containing an edge.

Traditionally, a form whose topograph has mirror symmetry is called “ambiguous” although there is really nothing about the form that is ambiguous in the usual sense of the word, unless perhaps it is the fact that such a form is indistinguishable from its mirror image.

**Proposition 5.7.** All orientation-reversing symmetries of the topograph of a form are reflections of either of the following two types:

(i) A reflection across a line perpendicular to an edge of the topograph. Forms with this type of symmetry are exactly the forms equivalent to a form \( ax^2 + cy^2 \).

(ii) A reflection across a line containing an edge of the topograph. Forms with this type of symmetry are exactly the forms equivalent to a form \( ax^2 + bxy + ay^2 \). Alternatively, one could take forms \( ax^2 + axy + cy^2 \), or forms \( ax^2 + cxy + cy^2 \).
In particular the principal forms $x^2 - ky^2$ and $x^2 + xy - ky^2$ have mirror symmetry, so there is at least one form with mirror symmetry in each discriminant.

**Proof**: We first check that an orientation-reversing symmetry must be a reflection across a line that either contains or is perpendicular to an edge of the topograph. For an elliptic form each symmetry must take the source vertex or edge to itself since this is where the smallest values of the form occur. In the case of a source edge, if an orientation-reversing symmetry $T$ does not interchange the two ends of the edge then $T$ must be reflection across a line containing this edge. If the two ends are interchanged then $T$ must reflect across a line perpendicular to the edge. For a source vertex, $T$ must take this vertex to itself and reflect across a line containing one of the three adjacent edges.

For a hyperbolic form each symmetry must take the separator line to itself since this line consists of the edges that separate positive from negative values of the form. Now it is a simple fact that a symmetry of a line $L$ that is divided into a sequence of edges, say of length 1, extending to infinity in both directions, must be either a translation along $L$ by some integer distance in either direction, or a reflection of $L$ fixing either a vertex of $L$ or the midpoint of an edge of $L$ and interchanging the two halves of $L$ on either side of the fixed point. This can be seen as follows. Symmetries are assumed to take vertices to vertices, so suppose the symmetry $T$ of $L$ sends a vertex $v$ to the vertex $T(v)$. Then if $T$ preserves the orientation of $L$ it must be a translation along $L$ by the distance from $v$ to $T(v)$ as one can see by considering what $T$ does to the two edges adjacent to $v$, then to the next two adjacent edges on either side, then the next two edges, and so on. If $T$ reverses the orientation of $L$ then either $T(v) = v$ or $T$ fixes the midpoint of the segment from $v$ to $T(v)$ since it sends this segment to a segment of the same length with one end at $T(v)$ but extending back toward $v$ since $T$ reverses orientation of $L$. Thus $T$ fixes a point of $L$ in either case, and it follows that $T$ must reflect $L$ across this fixed point, as one can again see by considering the edge or edges containing the fixed point, then the next two edges, and so on. If the distance from $v$ to $T(v)$ is an even integer the midpoint between $v$ and $T(v)$ will be a vertex and if it odd the midpoint will be a midpoint of an edge.

Returning to the situation of a symmetry $T$ of the topograph of a hyperbolic form that takes the separator line $L$ to itself and reverses orientation of the whole topograph, then $T$ must interchange the two ends of $L$ since $T$ preserves positive and negative values of the form. Thus $T$ is a reflection of $L$ and therefore a reflection of the whole plane across a line perpendicular to $L$. This reflector line must either contain one of the edges leading off the separator line or it must cross an edge of the separator line at its midpoint.

For a 0-hyperbolic form, if the two regions labeled 0 in the topograph have a border edge in common then an orientation-reversing symmetry must be a reflection across this edge. If the two regions do not have a common border edge they are joined
by a finite separator line and an orientation-reversing symmetry must take this line to itself, reflecting it across a perpendicular line that either contains an edge of the topograph or passes through the midpoint of an edge.

A parabolic form has a single 0 region in its topograph, so the border of this region must taken to itself by any symmetry. If the symmetry reverses orientation must be a reflection of this line just as in the hyperbolic case.

Now let us verify the remaining statements in items (i) and (ii) of the proposition using the figures below. In the first figure if a reflector line crosses the central edge horizontally then the upper and lower regions have the same label so $b = 0$. If we position the topograph so that the left and right regions correspond to the slopes $1/0$ and $0/1$ then the form is $ax^2 + cy^2$. Conversely, every form $ax^2 + cy^2$ has mirror symmetry across the line crossing the edge between the $a$ and $c$ regions.

The other case is mirror symmetry across a line containing an edge of the topograph. We can arrange that this edge is the central edge in the first figure above, with the left and right regions again corresponding to $1/0$ and $0/1$. Then $a = c$ so the form is $ax^2 + bxy + ay^2$ as shown in the second figure. This form is equivalent to a form $Ax^2 + Axy + Cy^2$ as we can see by moving the $1/0$ region from the left side of the figure to the top, keeping the $0/1$ region on the right. This gives the form $(2a + b)x^2 + (2a + b)xy + ay^2$ with the first two coefficients equal. Alternatively we could move the $1/0$ region to the bottom of the figure to get the form $(2a + b)x^2 + (2a + b)xy + ay^2$. For a form $Ax^2 + Axy + Cy^2$ we can switch $x$ and $y$ to get another equivalent form with second and third coefficients equal. The third figure above shows that all forms $Ax^2 + Axy + Cy^2$ arise in this way from forms $ax^2 + bxy + ay^2$ since the labels $A, A, C$ adjacent to the upper right edge determine all the other labels. A similar statement holds for forms $Ax^2 + Cxy + Cy^2$, by interchanging $x$ and $y$.

Here is a chart showing all six of the equivalent forms we just described with two of the three coefficients equal, where we abbreviate the form $ax^2 + bxy + cy^2$ to $[a, b, c]$:

<table>
<thead>
<tr>
<th>$2a + b, 2a + b, a$</th>
<th>$[a, b, a]$</th>
<th>$[a, 2a + b, 2a + b]$</th>
</tr>
</thead>
<tbody>
<tr>
<td>$2a - b, 2a - b, a$</td>
<td>$[a, -b, a]$</td>
<td>$[a, 2a - b, 2a - b]$</td>
</tr>
</tbody>
</table>
We could also start with a form \([A, A, C]\) and express the other five forms in terms of it:

\[
\begin{array}{ccc}
[4C - A, 4C - A, C] & [C, 2C - A, C] & [C, 4C - A, 4C - A]
\end{array}
\]

In each chart the three columns correspond to the three types of forms with two coefficients equal, while the two rows give the two forms of each type obtained by changing the sign of \(b\). Reflecting topographs across a horizontal axis thus corresponds to reflecting the chart across a horizontal axis, replacing either \(x\) or \(y\) by its negative. The same is true for reflections across a vertical axis, interchanging \(x\) and \(y\).

These charts show that each form of type \([a, b, a]\), \([A, A, C]\), or \([A, C, C]\) is paired with exactly one other form of the same type, the form in the same column of the chart. Forms \(ax^2 + cy^2\) also come in pairs, where \(ax^2 + cy^2\) is paired with \(cx^2 + ay^2\). There is one special situation when the two forms in a pair coincide, namely when \(b = 0\) in the first chart, so the form \([a, b, a]\) is \([a, 0, a]\) = \(a[1, 0, 1]\). This is also the situation where the two forms of a pair \([a, 0, c]\) and \([c, 0, a]\) coincide.

For an elliptic form \(ax^2 + bxy + cy^2\) that is reduced, so \(0 \leq b \leq a \leq c\), it is easy to recognize exactly when mirror symmetries occur. As one can see from the first figure above, there are symmetries when at least one of these three inequalities becomes an equality. When \(b = 0\) one has a source edge with a mirror symmetry across the perpendicular line. When \(b = a\) we have \(a - b + c = c\) so there is a mirror symmetry across the lower right edge. And when \(a = c\) one has mirror symmetry across the central edge. Since \(a\) and \(c\) are the two smallest labels on regions in the topograph, we see that reduced forms \(ax^2 + bxy + ay^2\) occur when the smaller two of the three labels at the source vertex are equal, and reduced forms \(ax^2 + axy + cy^2\) occur when the larger two labels are equal, at 0/1 and \(-1/1\).

Certain combinations of equalities in \(0 \leq b \leq a \leq c\) are also possible. If \(b = 0\) and \(a = c\) the form is \(a(x^2 + y^2)\) and there are mirror symmetries across the source edge as well as across the line perpendicular to this edge. The composition of these two reflections is a 180 degree rotational symmetry about the midpoint of the edge. Another possibility is that \(b = a = c\) so the form is \(a(x^2 + xy + y^2)\) and all three values of the form around the lower vertex in the diagram are equal. In this case there is a 120 degree rotational symmetry about this source vertex as well as mirror symmetries across the three adjacent edges. These are the only combinations that can occur since we must have \(a > 0\) so \(0 = b = a\) is impossible.

For reduced elliptic forms this exhausts all the possible symmetries since if we have strict inequalities \(0 < b < a < c\) then the values of the form in the four regions shown in the first figure above are all distinct.

If the separator line of a hyperbolic form has a mirror symmetry then because of periodicity there has to be at least one reflector line in each period, but in fact there
are exactly two reflector lines in each period. To see this, let $T$ be the translation by one period and let $R_1$ be a reflection across a reflector line $L_1$. If $L_2$ is the line halfway between $L_1$ and $T(L_1)$, then $T(R_1(L_2)) = L_2$.

Since $TR_1$ reverses orientation and fixes $L_2$ while preserving the positive and negative sides of the separator line, it must be a reflection $R_2$ across $L_2$. Thus there are at least two reflector lines in each period. There cannot be more than two since if $R_1$ and $R_2$ are the reflections across two adjacent reflector lines $L_1$ and $L_2$ then the composition $R_2R_1$, first reflecting by $R_1$ then by $R_2$, is orientation-preserving and sends $L_1$ to $R_2(R_1(L_1)) = R_2(L_1)$ so this composition is a symmetry translating the separator line by twice the distance between $L_1$ and $L_2$.

The distance between $L_1$ and $L_2$ must then be half the length of the period, otherwise if the translation $R_2R_1$ were some power $T^n$ of the basic periodicity translation $T$ with $|n| > 1$, there would be fewer than two reflector lines in a period.

Let us define the symmetric class number, $h_\Delta^s$ to be the number of equivalence classes of primitive forms of discriminant $\Delta$ with mirror symmetry. Recall that equivalence is the same as proper equivalence for forms with mirror symmetry. The ordinary class number $h_\Delta$ is thus $h_\Delta^s$ plus twice the number of equivalence classes of primitive forms without mirror symmetry. We have $h_\Delta \geq h_\Delta^s$, and in fact $h_\Delta$ is always an integer multiple of $h_\Delta^s$ as we will see in Proposition 7.17.

In contrast with $h_\Delta$ it is possible to compute $h_\Delta^s$ explicitly. Here is the result for elliptic and hyperbolic forms:

**Theorem 5.8.** If $\Delta$ is a nonsquare discriminant and $k$ is the number of distinct prime divisors of $\Delta$ then $h_\Delta^s = 2^{k-1}$ except in the following cases:

(i) If $\Delta = 4(4m + 1)$ then $h_\Delta^s = 2^{k-2}$.

(ii) If $\Delta = 32m$ then $h_\Delta^s = 2^k$.

For example, for the discriminants $\Delta = 60 = 3 \cdot 4 \cdot 5$ and $\Delta = -260 = -4 \cdot 5 \cdot 13$ that we looked at in the previous section the number of distinct prime divisors is $k = 3$ so the theorem says there are $2^2 = 4$ equivalence classes of mirror symmetric forms in these two cases since the exception situations in (i) and (ii) do not occur here and all
forms of these two discriminants are primitive. This agrees with what the topographs showed.

The proof of the theorem will be somewhat lengthy since there are a number of different cases to consider. Fortunately most of the complication disappears in the final answer.

Proof: By Proposition 5.7 every form with mirror symmetry is equivalent to a form \( ax^2 + cy^2 \) or \( ax^2 + axy + cy^2 \). The strategy will be to count how many of these special forms there are that are primitive with discriminant \( \Delta \), then determine which of these special forms are equivalent.

For counting the special forms \( ax^2 + cy^2 \) and \( ax^2 + axy + cy^2 \) we may assume \( a > 0 \) since \( a \) is the value of the form when \( (x, y) = (1, 0) \) and for elliptic forms we only consider those with positive values, while for hyperbolic forms we are free to change a form to its negative so it suffices to count only those with \( a > 0 \) and then double the result.

Case 1: Forms \( ax^2 + cy^2 \). Then \( \Delta = -4ac = 4\delta \) for \( \delta = -ac \). Primitivity of the form is equivalent to \( a \) and \( c \) being coprime. The only way to have coprime factors \( a \) and \( c \) of \( \delta = -ac \) is to take an arbitrary subset of the distinct primes dividing \( \delta \) and let \( a \) be the product of these primes each raised to the same power as in \( \delta \) (so \( a = 1 \) when we choose the empty subset). The number of such subsets is \( 2^{k'} \) where \( k' \) is the number of distinct prime divisors of \( \delta \), so there are \( 2^{k'} \) primitive forms \( ax^2 + cy^2 \) with \( a > 0 \).

Case 2: Forms \( ax^2 + axy + cy^2 \) with \( \Delta \) odd. We have \( \Delta = a^2 - 4ac \) so \( \Delta \) and \( a \) have the same parity. From \( \Delta = a(a - 4c) \) we see that \( a \) divides \( \Delta \). We claim that each divisor \( a \) of \( \Delta \) gives rise to a form \( ax^2 + axy + cy^2 \) of discriminant \( \Delta \). Solving \( \Delta = a^2 - 4ac \) for \( c \) gives \( c = (a^2 - \Delta)/4a \). The numerator is divisible by 4 since \( a \) and \( \Delta \) are odd and hence \( a^2 \) and \( \Delta \) are both 1 mod 4, making the numerator 0 mod 4. The numerator is also divisible by \( a \) if \( a \) divides \( \Delta \). Since 4 and \( a \) are coprime when \( a \) is odd it follows that 4\( a \) divides the numerator so \( c \) is an integer and we get a form \( ax^2 + axy + cy^2 \) of discriminant \( \Delta \). This form is primitive exactly when \( a \) and \( c \) are coprime. This is equivalent to saying that the two factors of \( \Delta = a(a - 4c) \) are coprime since any divisor of \( a \) and \( c \) must divide the two factors, and conversely any divisor of the two factors must divide \( a \) and \( 4c \), hence also \( c \) since this divisor of the odd number \( a \) must be odd. As in Case 1, the only way to obtain a factorization \( \Delta = a(a - 4c) \) with the two factors coprime is to take an arbitrary subset of the distinct primes dividing \( \Delta \) and let \( a \) be the product of these primes each raised to the same power as in \( \Delta \). The number of such subsets is \( 2^k \) so this is the number of primitive forms \( ax^2 + axy + cy^2 \) with \( a > 0 \) when \( \Delta \) is odd.

There remain the forms \( ax^2 + axy + cy^2 \) with \( \Delta = 4\delta \). Again \( \Delta \) and \( a \) have the same parity since \( \Delta = a^2 - 4ac \), so \( a \) is even, say \( a = 2d \). From \( \Delta = a^2 - 4ac \) we
then have $\delta = d^2 - 2dc = d(d - 2c)$.

**Case 3:** Forms $ax^2 + axy + cy^2$ with $\Delta = 4\delta$ and $a = 2d$ for odd $d$. By primitivity $c$ must be odd. The two factors of $\delta = d(d - 2c)$ are odd and must be distinct mod 4 since $c$ is odd. Thus one factor is 1 mod 4 and the other is 3 mod 4, so $\delta \equiv 3 \pmod{4}$, say $\delta = 4m + 3$. We claim that when $\delta = 4m + 3$, each divisor $d$ of $\delta$ gives rise to a form $ax^2 + axy + cy^2$ with $a = 2d$. To show this, note first that $d$ must be odd since it divides $\delta$ which is odd. Solving $\delta = d(d - 2c)$ for $c$ gives $c = (d^2 - \delta)/2d$. Since $d$ and $\delta$ are odd, the numerator of $(d^2 - \delta)/2d$ is even hence divisible by the 2 in the denominator. The numerator is also divisible by the $d$ in the denominator if $d$ divides $\delta$. Since $d$ is odd, this implies that $2d$ divides the numerator, so $c$ is an integer for each divisor $d$ of $\delta$. In fact $c$ is an odd integer since the numerator $d^2 - \delta$ is 2 mod 4 and so $cd = (d^2 - \delta)/2$ is odd, forcing $c$ to be odd. For the form $ax^2 + axy + cy^2$ to be primitive means that $a$ and $c$ are coprime. Since $c$ is odd and $a = 2d$ this is equivalent to $c$ and $d$ being coprime. This in turn is equivalent to the two factors of $\delta = d(d - 2c)$ being coprime since $c$ and $d$ are odd. Thus when $\delta = 4m + 3$ we get a primitive form $ax^2 + axy + cy^2$ for each choice of a subset of the distinct prime divisors of $\delta$ since this determines $d$ as before, and $d$ determines $c$ and $a$. The number of primitive forms $ax^2 + axy + cy^2$ is then $2^k'$ when $\Delta$ is even and $a = 2d$ with $d$ odd, where $k'$ is the number of distinct prime divisors of $\delta$ as in Case 1.

**Case 4:** Forms $ax^2 + axy + cy^2$ with $\Delta$ even and $a = 2d$ for even $d$, say $d = 2e$. Then $\delta = d(d - 2c) = 4e(e - c)$. Since $c$ is odd by primitivity of the form, the two factors $e$ and $e - c$ of $\delta = 4e(e - c)$ have opposite parity, hence $\delta$ must be divisible by 8, say $\delta = 8m$. We need to determine which choices of $e$ and $c$ yield primitive forms $ax^2 + axy + cy^2$. Let $\delta' = \delta/4 = 2m$ so the equation $\delta = 4e(e - c)$ becomes $\delta' = e(e - c)$. Thus $e$ must divide $\delta'$. We have $c = e - (\delta'/e)$ and this will be an integer if $e$ divides $\delta'$. From the equation $c = e - (\delta'/e)$ we see that any divisor of two of the three terms will divide the third. In particular, $c$ and $e$ will be coprime exactly when $e$ and $\delta'/e$ are coprime. Since $\delta' = e \cdot (\delta'/e)$ this means we want to choose $e$ by choosing some subset of the distinct prime divisors of $\delta'$ and letting $e$ be the product of these primes raised to the same powers as in $\delta'$. Then $e$ and $\delta'/e$ will be coprime and of opposite parity since they are not both even and their product $\delta'$ is even. Their difference $c = e - (\delta'/e)$ will then be odd. Also, $c$ and $e$ will be coprime so $c$ and $a = 4e$ will be coprime, making the form $ax^2 + axy + cy^2$ primitive. The number of distinct prime divisors of $\delta'$ is the same as for $\delta = 4\delta'$ since $\delta'$ is even. Thus in Case 4 the number of primitive forms $ax^2 + axy + cy^2$ with $a > 0$ is $2^{k'}$. Note that $k' = k$ when $\delta$ is even and $k' = k - 1$ when $\delta$ is odd. By combining the four cases above and remembering to double the number of forms when $\Delta > 0$ to account for negative coefficients of $x^2$, we then obtain the following table for the number of forms of either of the types $ax^2 + cy^2$ or $ax^2 + axy + cy^2$ for each
discriminant:

<table>
<thead>
<tr>
<th>( \Delta )</th>
<th>odd ( 4\delta , \delta = 4m + 1 )</th>
<th>odd ( 4\delta , \delta = 4m + 3 )</th>
</tr>
</thead>
<tbody>
<tr>
<td>Cases</td>
<td>(2)</td>
<td>(1) &amp; (3)</td>
</tr>
</tbody>
</table>

<table>
<thead>
<tr>
<th>( \Delta )</th>
<th>even ( 4\delta , \delta = 4m + 1 )</th>
<th>( 4\delta , \delta = 4m + 3 )</th>
</tr>
</thead>
<tbody>
<tr>
<td>Cases</td>
<td>(1) &amp; (4)</td>
<td>(1)</td>
</tr>
</tbody>
</table>

Comparing the results in the table with the statement of the theorem, we see that the proof will be finished when we show that under the relation of equivalence the special forms split up into pairs when \( \Delta < 0 \) and into groups of four when \( \Delta > 0 \). To do this we can use forms \( ax^2 + cy^2 \) instead of \( ax^2 + axy + cy^2 \) since there is a one-to-one correspondence between the forms of these two types, as we saw at the end of the proof of the preceding proposition.

Two easy cases that can be disposed of first are \( \Delta = -3 \) and \( \Delta = -4 \). Here all forms are equivalent and are primitive, and \( k = 1 \), so the theorem is true since the exceptional cases (i) and (ii) do not apply.

The figure below shows how a reflector line intersects the topograph for the forms \( ax^2 + cy^2 \) and \( ax^2 + bxy + ay^2 \).

In the first case we may assume \( a \neq c \), otherwise primitivity would imply \( a = c = 1 \) with \( \Delta = -4 \). In the second case we may assume \( b \neq 0 \) for the same reason, and we can also assume \( a \neq \pm b \) otherwise primitivity would imply the form is \( x^2 \pm xy + y^2 \) with \( \Delta = -3 \).

Consider first elliptic forms, with \( \Delta < 0 \). Here we have a source edge in the middle of the first figure above for the forms \( ax^2 + cy^2 \), while for the forms \( ax^2 + bxy + ay^2 \) we have a source vertex, either the upper or lower vertex in the second figure since the source vertex must be at one end of the edge contained in the line of reflection. No form of the first type is equivalent to one of the second type since a form cannot have both a source edge and a source vertex, so we can treat the two types separately. For the first type we can interchange \( a \) and \( c \) to get an equivalent form, but no other
equivalences are possible since $a$ and $c$ are the two smallest values of the form. For the second type with a source vertex we can change $b$ to $-b$ to get two equivalent forms, moving the source vertex from one end of the central edge to the other, but there are no other equivalences since the labels adjacent to a source vertex are the smallest values of the form and we can distinguish the labels $a$ from the other label $2a \pm b$ at the source vertex by the fact that two of the three adjacent labels are $a$. (We have already taken care of the case that all three labels at the source vertex are equal, so $\Delta = -3$.) This finishes the proof of the theorem when $\Delta < 0$.

When $\Delta > 0$ the reflector line in the first figure above crosses the central edge so this edge must be contained in the separator line, with the labels $a$ and $c$ of opposite sign. We can interchange $a$ and $c$ to get two equivalent forms, and we can eliminate this ambiguity by requiring $a > 0$. In the second figure the central edge is contained in the reflector line so the separator line contains either the upper two edges or the lower two edges. We can interchange $b$ and $-b$ to get two equivalent forms, and this twofold ambiguity can be eliminated by requiring the separator line to contain the upper two edges, adjacent to the region labeled $2a + b$. Thus we have divided the special forms into pairs of equivalent forms with a preferred form in each pair. These pairs are further paired to give groups of four since there are two reflector lines in each period along the separator line. The preferred forms at these two reflector lines cannot be the same, otherwise there would be a translational symmetry taking one reflector line to an adjacent reflector line, contradicting the fact that there are two reflector lines in each period. There can be no further equivalences between the special forms since the equivalence class determines the separator line and the reflector lines and hence also the labels on the regions adjacent to the intersection points of separator and reflector lines. Thus when $\Delta > 0$ the number of equivalence classes of special forms is one-quarter of the number of special forms and the proof is complete.

In the first part of the proof we used forms $ax^2 + axy + cy^2$ rather than the more obviously symmetric forms $ax^2 + bxy + ay^2$ since the discriminants in the two cases are $\Delta = a^2 - 4ac = a(a - 4c)$ and $\Delta = b^2 - 4a^2 = (b + 2a)(b - 2a)$, with the first factorization being a little simpler and easier to analyze than the second.

We illustrate the theorem with an example, the first negative discriminant with four distinct prime divisors, $\Delta = -420 = -3 \cdot 4 \cdot 5 \cdot 7$. This is a case when $\Delta = 4(4m+3)$ so the theorem says there are $2^3 = 8$ equivalence classes of symmetric primitive forms. If we compute all the reduced forms for $\Delta = -420$ by the method earlier in the chapter we get the following table, with the letter $b$ replacing $h$ so we are finding solutions of $b^2 + 420 = 4ac$ with $0 \leq b \leq a \leq c$. 

\[
\begin{array}{c|c|c|c|c}
 a & b & c & \text{Equivalence Class} \\
\hline
 1 & 0 & 0 & 1 \\
 0 & 1 & 0 & 2 \\
 0 & 0 & 1 & 3 \\
 1 & 1 & 0 & 4 \\
 1 & 0 & 1 & 5 \\
 0 & 1 & 1 & 6 \\
 1 & 1 & 1 & 7 \\
\end{array}
\]
<table>
<thead>
<tr>
<th>$b$</th>
<th>$ac$</th>
<th>$(a,c)$</th>
<th>$[a,b,c]$</th>
<th>equivalent forms</th>
</tr>
</thead>
<tbody>
<tr>
<td>0</td>
<td>105</td>
<td>$(1,105)$</td>
<td>$[1,0,105]$</td>
<td>$[105,0,1]$</td>
</tr>
<tr>
<td></td>
<td>(3,35)</td>
<td>[3,0,35]</td>
<td>$[35,0,3]$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>(5,21)</td>
<td>[5,0,21]</td>
<td>$[21,0,5]$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>(7,15)</td>
<td>[7,0,15]</td>
<td>$[15,0,7]$</td>
<td></td>
</tr>
<tr>
<td>2</td>
<td>106</td>
<td>(2,53)</td>
<td>$[2,2,53]$</td>
<td>$[53,104,53],[210,210,53]$</td>
</tr>
<tr>
<td>4</td>
<td>109</td>
<td>—</td>
<td></td>
<td></td>
</tr>
<tr>
<td>6</td>
<td>114</td>
<td>(6,19)</td>
<td>$[6,6,19]$</td>
<td>$[19,32,19],[70,70,19]$</td>
</tr>
<tr>
<td>8</td>
<td>121</td>
<td>(11,11)</td>
<td>$[11,8,11]$</td>
<td>$[14,14,11],[30,30,11]$</td>
</tr>
<tr>
<td>10</td>
<td>130</td>
<td>(10,13)</td>
<td>$[10,10,13]$</td>
<td>$[13,16,13],[42,42,13]$</td>
</tr>
</tbody>
</table>

Thus all forms of discriminant $-420$ are symmetric and primitive. The first four have $b = 0$ so these arise in Case 1 in the proof of the theorem where we set $\Delta = 4\delta$, so $\delta = -3 \cdot 5 \cdot 7$ and we get a form $[a,0,c]$ for each divisor $a$ of $\delta$. This gives the first four entries in the $[a,b,c]$ column of the table. Reversing $a$ and $c$ gives equivalent forms, the four corresponding entries in the last column. The remaining four forms in the $[a,b,c]$ column are the remaining four reduced forms. These have $b$ nonzero and are instances of forms $[a,a,c]$ and $[a,b,a]$, with three of type $[a,a,c]$ and one of type $[a,b,a]$. According to Case 3 in the proof of the theorem the numbers $a$ in the forms $[a,a,c]$ should be twice the numbers $a$ in the forms $[a,0,c]$, and they are: $2 = 2 \cdot 1$, $6 = 2 \cdot 3$, $10 = 2 \cdot 5$, $14 = 2 \cdot 7$, $30 = 2 \cdot 15$, $42 = 2 \cdot 21$, $70 = 2 \cdot 35$, and $210 = 2 \cdot 105$.

**Corollary 5.9.** For fundamental nonsquare discriminants $\Delta$ the symmetric class number $h_\Delta^s$ is 1 only when $\Delta = -4, \pm 8$ and $\pm p$ for odd primes $p$, with $p \equiv 1 \mod 4$ when $\Delta > 0$ and $p \equiv 3 \mod 4$ when $\Delta < 0$. More generally, if nonfundamental nonsquare discriminants are allowed then the only cases when $h_\Delta^s = 1$ are $\Delta = -4, \pm 8, -16$ as well as $\pm p^{2k+1}$ and $\pm 4p^{2k+1}$ for odd primes $p$ with $p \equiv 1 \mod 4$ when $\Delta > 0$ and $p \equiv 3 \mod 4$ when $\Delta < 0$.

**Proof:** Consider first the case $\Delta > 0$. If we are not in one of the exceptional cases (i) and (ii) in Theorem 5.8 then $\Delta$ must have just one distinct prime divisor so it must be a power of a prime, in fact an odd power since it is not a square. Thus for $p$ odd we have $\Delta = p^{2k+1}$ and we must have $p \equiv 1 \mod 4$ in order to have $\Delta \equiv 1 \mod 4$. For odd powers of $p = 2$ the only possibility is $\Delta = 8$ since $\Delta$ cannot be 2 and odd powers beyond 8 are of the form $\Delta = 32m$, the exceptional case (ii) where $h_\Delta^s$ is always even, so this is ruled out as well. In the exceptional case (i) we have $\Delta = 4(4m + 1)$ with $4m + 1$ a prime power $p^{2k+1}$ with $p \equiv 1 \mod 4$.

When $\Delta < 0$ the reasoning is similar, the main difference being that $-p^{2k}$ and $-4p^{2k}$ are ruled out not because squares are excluded but because $p^{2k}$ is always $1 \mod 4$ when $p$ is odd, so $-p^{2k}$ is $3 \mod 4$. This rules out $-p^{2k}$ as a discriminant, and it rules out $-4p^{2k}$ being an exceptional case $\Delta = 4(4m + 1)$.

Requiring $\Delta$ to be a fundamental discriminant eliminates the cases $\Delta = -16$ and $\pm 4p^{2k+1}$ and restricts the exponent in $\pm p^{2k+1}$ to be 1. $\Box$
We have mentioned the fact that $h_\Delta$ is always a multiple of $h_\Delta^s$, which will be proved in Proposition 7.17. When $h_\Delta^s = 1$ this tells us nothing about $h_\Delta$, but we will also prove that $h_\Delta^s = 1$ exactly when $h_\Delta$ is odd. Thus Corollary 5.9 above gives a way to determine whether $h_\Delta$ is even or odd. In the examples we have looked at so far $h_\Delta$ has been either 1 or even, but odd numbers besides 1 can also occur as class numbers. Here is a table with some examples for negative discriminants, so we are finding the solutions of $h^2 + |\Delta| = 4ac$ with $0 \leq h \leq a \leq c$ as usual.

<table>
<thead>
<tr>
<th>$\Delta$</th>
<th>$h$</th>
<th>$ac$</th>
<th>$(a,c)$</th>
<th>$h_\Delta$</th>
</tr>
</thead>
<tbody>
<tr>
<td>$-23$</td>
<td>1</td>
<td>6</td>
<td>$(1,6),(2,3)$</td>
<td>3</td>
</tr>
<tr>
<td>$-47$</td>
<td>1</td>
<td>12</td>
<td>$(1,12),(2,6),(3,4)$</td>
<td>5</td>
</tr>
<tr>
<td></td>
<td>3</td>
<td>14</td>
<td>—</td>
<td></td>
</tr>
<tr>
<td>$-71$</td>
<td>1</td>
<td>18</td>
<td>$(1,18),(2,9),(3,6)$</td>
<td>7</td>
</tr>
<tr>
<td></td>
<td>3</td>
<td>20</td>
<td>$(4,5)$</td>
<td></td>
</tr>
<tr>
<td>$-199$</td>
<td>1</td>
<td>50</td>
<td>$(1,50),(2,25),(5,10)$</td>
<td>9</td>
</tr>
<tr>
<td></td>
<td>3</td>
<td>52</td>
<td>$(4,13)$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>5</td>
<td>56</td>
<td>$(7,8)$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>7</td>
<td>62</td>
<td>—</td>
<td></td>
</tr>
<tr>
<td>$-167$</td>
<td>1</td>
<td>42</td>
<td>$(1,42),(2,21),(3,14),(6,7)$</td>
<td>11</td>
</tr>
<tr>
<td></td>
<td>3</td>
<td>44</td>
<td>$(4,11)$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>5</td>
<td>48</td>
<td>$(6,8)$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>7</td>
<td>54</td>
<td>—</td>
<td></td>
</tr>
<tr>
<td>$-191$</td>
<td>1</td>
<td>48</td>
<td>$(1,48),(2,24),(3,16),(4,12),(6,8)$</td>
<td>13</td>
</tr>
<tr>
<td></td>
<td>3</td>
<td>50</td>
<td>$(5,10)$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>5</td>
<td>54</td>
<td>$(6,9)$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>7</td>
<td>60</td>
<td>—</td>
<td></td>
</tr>
<tr>
<td>$-239$</td>
<td>1</td>
<td>60</td>
<td>$(1,60),(2,30),(3,20),(4,15),(5,12),(6,10)$</td>
<td>15</td>
</tr>
<tr>
<td></td>
<td>3</td>
<td>62</td>
<td>—</td>
<td></td>
</tr>
<tr>
<td></td>
<td>5</td>
<td>66</td>
<td>$(6,11)$</td>
<td></td>
</tr>
<tr>
<td></td>
<td>7</td>
<td>72</td>
<td>$(8,9)$</td>
<td></td>
</tr>
</tbody>
</table>

The examples in the table are all fundamental discriminants, and in each case they are the first negative discriminant with the given class number.

Besides the cases when $h_\Delta^s = 1$ another nice situation is when $h_\Delta = h_\Delta^s$ so all primitive forms of discriminant $\Delta$ have mirror symmetry. We call such discriminants fully symmetric. As we will see in the following chapters, forms with fully symmetric discriminants have very special properties. A table at the end of the book lists the 101 known negative discriminants that are fully symmetric, ranging from $-3$ to $-7392$. Of these, 65 are fundamental discriminants, the largest being $-5460$. Since 5460 factors as $3 \cdot 4 \cdot 5 \cdot 7 \cdot 13$ with five distinct prime factors, Theorem 5.8 says that $h_\Delta^s = 2^4 = 16$. This is in fact the largest value of $h_\Delta^s$ among the 101 discriminants in the list. Computer calculations have extended to much larger negative discriminants.
without finding any more that are fully symmetric. It has not yet been proved that no more exist, although it is known that there are at most two more. For positive discriminants there are probably infinitely many that are fully symmetric since it is likely that there are already infinitely many with $h_\Delta = 1$.

Among the examples of hyperbolic forms we have considered there were some whose topograph had a “symmetry” which was a glide-reflection along the separator line that had the effect of changing each value to its negative rather than preserving the values. These are not actual symmetries according to the definition we have given, so let us call such a transformation that takes each value of a form to its negative a skew symmetry. (Compare this with skew-symmetric matrices in linear algebra which equal the negative of their transpose.)

A skew symmetry must take the separator line to itself while interchanging the two sides of the separator line, so it either translates the separator line along itself and hence is a glide-reflection, or it reflects the separator line, interchanging its two ends as well as the two sides of the separator line, making it a 180 degree rotation about a point of the separator line. Examples of forms with this sort of skew symmetry occurred in Chapter 4, the forms $x^2 - 13y^2$ and $10x^2 - 29y^2$.

The figures below shows forms whose separator lines have all the possible combinations of symmetries and skew symmetries.

The first form has all four types: translations, mirror symmetries, glide-reflections, and rotations. The next three forms have only one type of symmetry or skew symmetry besides translations, while the last form has only translational symmetries and no
mirror symmetries or skew symmetries. It is not possible to have two of the three
types of non-translational symmetries and skew symmetries without having the third
since the composition of two of these three types gives the third type. One can see
this by considering the effect of a symmetry or skew symmetry on the orientation of
the plane and the orientation of the separator line. The four possible combinations
distinguish the four types of transformations according to the following chart, where
+ denotes orientation-preserving and − denotes orientation-reversing.

<table>
<thead>
<tr>
<th></th>
<th>plane orientation</th>
<th>line orientation</th>
</tr>
</thead>
<tbody>
<tr>
<td>translation</td>
<td>+</td>
<td>+</td>
</tr>
<tr>
<td>rotation</td>
<td>+</td>
<td>−</td>
</tr>
<tr>
<td>glide reflection</td>
<td>−</td>
<td>+</td>
</tr>
<tr>
<td>reflection</td>
<td>−</td>
<td>−</td>
</tr>
</tbody>
</table>

A rotational skew symmetry is a rotation about the midpoint of an edge of the
separator line where the two adjacent regions have labels $a$ and $-a$. If the edge
separating these two regions has label $b$ then the form associated to this edge is
$ax^2 + bxy - ay^2$. Conversely any form $ax^2 + bxy - ay^2$ whose discriminant $\Delta = b^2 + 4a^2$ is not a square (although it is the sum of two squares) will be a hyperbolic form
having a rotational skew symmetry, as one can see in the figure.

Note that the form $ax^2 + bxy - ay^2$ will be one of the reduced
forms in the equivalence class of the given form since the two edges
leading off the separator line at the ends of the edge labeled $b$ do
so on opposite sides of the separator line. Thus rotational skew symmetries can be
detected by looking just at the reduced forms. The same is true for mirror symmetries
and glide-reflection skew symmetries, but for these one must look at the arrangement
of the whole cycle of reduced forms rather than just the individual reduced forms.

For rotational skew symmetries there are two rotation points along the separator
line in each period, just as reflector lines occur in pairs in each period. The proof is
similar.

5. Charting All Forms

We have used the Farey diagram to study individual quadratic forms through
their topographs, but the diagram also appears in another way when one seeks a
global picture of all forms simultaneously, as we will now see.

Quadratic forms are defined by formulas $ax^2 + bxy + cy^2$, and our point of view
will be to regard the coefficients $a$, $b$, and $c$ as parameters that vary over all integers
independently. It is natural to consider the triples $(a, b, c)$ as points in 3-dimensional
Euclidean space $\mathbb{R}^3$, and more specifically as points in the integer lattice $\mathbb{Z}^3$ consisting
of points $(a, b, c)$ whose coordinates are integers. We will exclude the origin $(0, 0, 0)$
since this corresponds to the trivial form that is identically zero. Instead of using the
traditional $(x, y, z)$ as coordinates for $\mathbb{R}^3$ we will use $(a, b, c)$, but since $a$ and $c$
play a symmetric role as the coefficients of the squared terms $x^2$ and $y^2$ in a form $ax^2 + bxy + cy^2$ we will position the $a$ and $c$ axes in a horizontal plane, with the $b$ axis vertical, perpendicular to the $ac$ plane. Here is a figure showing the location of a few forms:

Along each ray starting at the origin and passing through a lattice point $(a, b, c)$ there are infinitely many lattice points $(ka, kb, kc)$ for positive integers $k$. If $a$, $b$, and $c$ have a common divisor greater than 1 we can first cancel this common divisor to get a primitive triple $(a, b, c)$ corresponding to a primitive form $ax^2 + bxy + cy^2$, with all the other lattice points on the ray through $(a, b, c)$ being the positive integer multiples of this. Thus primitive forms correspond exactly to rays from the origin passing through lattice points. These are the same as rays passing through points $(a, b, c)$ with rational coordinates since denominators can always be eliminated by multiplying the coordinates by the least common multiple of the denominators.

Since the discriminant $\Delta = b^2 - 4ac$ plays such an important role in the classification of forms, let us see how this fits into the picture in $(a, b, c)$ coordinates. When $b^2 - 4ac$ is zero we have the special class of parabolic forms, and the points in $\mathbb{R}^3$ satisfying the equation $b^2 - 4ac = 0$ form a double cone with the common vertex of the two cones at the origin, as shown in the figure below. The double cone intersects the $ac$ plane in the $a$ and $c$ axes. The central axis of the double cone itself is the line $a = c$ in the $ac$ plane. Points $(a, b, c)$ inside either cone have $b^2 - 4ac < 0$ so the lattice points inside the cones correspond to elliptic forms. Positive elliptic forms have $a > 0$ and $c > 0$ so they lie inside the cone projecting to the first quadrant of the $ac$ plane. We call this the positive cone. Inside the other cone are the negative elliptic forms, those with $a < 0$ and $c < 0$. Outside the cones is a single region consisting of points with $b^2 - 4ac > 0$ so the lattice points here correspond to hyperbolic forms and 0-hyperbolic forms.

If one slices the positive cone via a vertical plane perpendicular to the axis of the
cone such as the plane $a + c = 1$, then the intersection of the cone with this plane is an ellipse which we denote $E$.

The top and bottom points of $E$ are $(a, b, c) = (\frac{1}{2}, \pm 1, \frac{1}{2})$ so its height is 2. The left and right points of $E$ are $(1, 0, 0)$ and $(0, 0, 1)$ so its width is $\sqrt{2}$. Thus $E$ is somewhat elongated vertically. If we wanted, we could compress the vertical coordinate to make $E$ a circle, but there is no special advantage to doing this.

When we project a lattice point $(a, b, c)$ corresponding to a primitive positive elliptic form along the ray to the origin passing through $(a, b, c)$, this ray intersects the plane $a + c = 1$ in the point $(a/(a + c), b/(a + c), c/(a + c))$ since the sum of the first and third coordinates of this point is 1. This point lies inside the ellipse $E$ and has rational coordinates. Conversely, every point inside $E$ with rational coordinates is the radial projection of a unique primitive positive elliptic form, obtained by multiplying the coordinates of the point by the least common multiple of their denominators. Thus the rational points inside $E$ parametrize primitive positive elliptic forms. We shall use the notation $[a, b, c]$ to denote both the form $ax^2 + bxy + cy^2$ and the corresponding rational point $(a/(a + c), b/(a + c), c/(a + c))$ inside $E$. The figure below shows some examples, including a few parabolic forms on $E$ itself.
In this figure the lines radiating out from the points \([1, 0, 0]\) and \([0, 0, 1]\) consist of the points \([a, b, c]\) with a fixed ratio \(a/b\) or \(b/c\). The ratios \(a/c\) are fixed along vertical lines. Two out of three of these ratios determine the third since \(a \cdot b = a/c\).

Of special interest are the reduced primitive elliptic forms \([a, b, c]\), which are those satisfying \(0 \leq b \leq a \leq c\) where \(a, b,\) and \(c\) have no common divisor. These correspond to the rational points in the shaded triangle in the figure above, with vertices \([1, 1, 1]\), \([1, 0, 1]\), and \([0, 0, 1]\). The edges of the triangle correspond to one of the three inequalities \(0 \leq b \leq a \leq c\) becoming an equality, so \(b = 0\) for the lower edge, \(a = c\) for the vertical edge, and \(a = b\) for the hypotenuse. Thus the three edges correspond to the reduced forms with mirror symmetry, the forms \([a, 0, c]\) for the bottom edge, \([a, b, a]\) for the left edge, and \([a, a, c]\) for the diagonal edge. Points in the interior of the triangle correspond to forms without mirror symmetry.

Just as rational points inside the ellipse \(E\) correspond to primitive positive elliptic forms, the rational points on \(E\) itself correspond to primitive positive parabolic forms. As we know, every parabolic form is equivalent to the form \(ax^2\) for some nonzero integer \(a\). For this to be primitive means that \(a = \pm 1\), so every positive primitive parabolic form is equivalent to \(x^2\). Equivalent forms are those that can be obtained from each other by a change of variable replacing \((x, y)\) by \((px + qy, rx + sy)\) for
some integers \(p, q, r, s\) satisfying \(ps - qr = \pm 1\). For the form \(x^2\) this means that the primitive positive parabolic forms are the forms \((px + qy)^2 = p^2x^2 + 2pqxy + q^2y^2\) for any pair of coprime integers \(p\) and \(q\). In \([a, b, c]\) notation this is \([p^2, 2pq, q^2]\), defining a point on the ellipse \(E\).

More concisely, we could label the rational point on \(E\) corresponding to the form \((px + qy)^2\) just by the fraction \(p/q\). Thus at the left and right sides of \(E\) we have the fractions 1/0 and 0/1 corresponding to the forms \(x^2\) and \(y^2\), while at the top and bottom of \(E\) we have 1/1 and \(-1/1\) corresponding to \((x + y)^2\) and \((x - y)^2 = (-x + y)^2\).

Note that changing the signs of both \(p\) and \(q\) does not change the form \((px + qy)^2\) or the fraction \(p/q\). In the first quadrant of the ellipse the fractions \(p/q\) increase monotonically from 0/1 to 1/1 since the ratio \(b/c\) equals \(2p/q\) and \(b\) is increasing while \(c\) is decreasing so \(2p/q\) is increasing, and hence also \(p/q\). Similarly in the second quadrant the values of \(p/q\) increase from 1/1 to 1/0 since we have \(b/a = 2q/p\) which decreases as \(b\) decreases and \(a\) increases. In the lower half of the ellipse we have just the negatives of the values in the upper half since the sign of \(b\) has changed from plus to minus.

This labeling of the rational points of \(E\) by fractions \(p/q\) seems very similar to the labeling of vertices in the circular Farey diagram. As we saw in Section 1.4, if the Farey diagram is drawn with 1/0 at the top of the unit circle in the \(xy\) plane, then the point on the unit circle labeled \(p/q\) has coordinates \((x, y) = (2pq/(p^2 + q^2), (p^2 - q^2)/(p^2 + q^2))\). After rotating the circle to put 1/0 on the left side by replacing \((x, y)\) by \((-y, x)\) this becomes \(( (q^2 - p^2)/(p^2 + q^2), 2pq/(p^2 + q^2))\). Here the \(y\)-coordinate \(2pq/(p^2 + q^2)\) is the same as the \(b\)-coordinate of the point of \(E\) labeled \(p/q\), namely the point \((a, b, c) = (p^2/(p^2 + q^2), 2pq/(p^2 + q^2), q^2/(p^2 + q^2))\). Since the vertical coordinates of points in either the left or right half of the unit circle or the ellipse \(E\) determine the horizontal coordinates uniquely, this means that the labeling of points...
Let us return now to the general picture of how forms \( ax^2 + bxy + cy^2 \) are represented by points \((a, b, c)\) in \( \mathbb{R}^3 \). As we know, a change of variables by a linear transformation \( T \) sending \((x, y)\) to \( T(x, y) = (px + qy, rx + sy)\), where \( p, q, r, s \) are integers with \( ps - qr = \pm 1 \), transforms each form into another equivalent form. To see the effect of this change of variables on the coefficients \((a, b, c)\) of a form \( Q(x, y) = ax^2 + bxy + cy^2 \) we do a simple calculation:

\[
Q(px + qy, rx + sy) = a(px + qy)^2 + b(px + qy)(rx + sy) + c(rx + sy)^2
= (ap^2 + bpr + c^2r^2)x^2 + (2apq + bps + bqr + 2crs)xy + (aq^2 + bqs + cs^2)y^2
\]

This means that the \((a, b, c)\) coordinates of points in \( \mathbb{R}^3 \) are transformed according to the formula

\[
T^*(a, b, c) = (p^2a + prb + r^2c, 2pqa + (ps + qr)b + 2rsq, aq^2 + bsq + s^2c)
\]

For fixed values of \( p, q, r, s \) this \( T^* \) is a linear transformation of the variables \( a, b, c \). Its matrix is

\[
\begin{pmatrix}
p^2 & pr & r^2 \\
2pq & ps + qr & 2rs \\
q^2 & qs & s^2
\end{pmatrix}
\]

Since \( T^* \) is a linear transformation, it takes lines to lines and planes to planes, but \( T^* \) also has another special geometric property. Since equivalent forms have the same discriminant, this means that each surface defined by an equation \( b^2 - 4ac = k \) for \( k \) a constant is taken to itself by \( T^* \). In particular, the double cone \( b^2 - 4ac = 0 \) is taken to itself, and in fact each of the two cones separately is taken to itself since one cone consists of positive parabolic forms and the other cone of negative parabolic forms (as one can see just by looking at the coefficients \( a \) and \( c \)), and positive parabolic forms are never equivalent to negative parabolic forms. When \( k > 0 \) the surface \( b^2 - 4ac = k \) is a hyperboloid of one sheet and when \( k < 0 \) it is a hyperboloid of two sheets. In the case of two sheets the lattice points on one sheet give positive elliptic forms and those on the other sheet give negative elliptic forms.

Since \( T^* \) takes lines through the origin to lines through the origin and it takes the double cone \( b^2 - 4ac = 0 \) to itself, this means that \( T^* \) gives a transformation of the ellipse \( E \) to itself, taking rational points to rational points since rational points on \( E \) correspond to lattice points on the cones. Regarding \( E \) as the boundary circle of the Farey diagram, we know that linear fractional transformations give symmetries of the Farey diagram, also taking rational points on the boundary circle to rational boundary points. And in fact, the transformation of this circle defined by \( T^* \) is exactly one of these linear fractional transformations. This is because \( T^* \) takes the parabolic form \((dx + ey)^2\) to the form \((dp + er)x + (dq + es)y)^2 = ((dp + er)x + (dq + es)y)^2
so in the fractional labeling of points of $E$ this says $T^*(d/e) = (pd + re)/(qd + se)$ which is a linear fractional transformation. If we write this using the variables $x$ and $y$ instead of $d$ and $e$ it would be $T^*(x/y) = (px + ry)/(qx + sy)$. This is not quite the same as the linear fractional transformation $T(x/y) = (px + qy)/(rx + sy)$ defined by the original change of variables $T(x,y) = (px + qy, rx + sy)$, but rather $T^*$ is obtained from $T$ by transposing the matrix of $T$, interchanging the off-diagonal terms $q$ and $r$.

Via radial projection, the transformation $T^*$ determines a transformation not just of $E$ but of the interior of $E$ in the plane $a + c = 1$ as well. This transformation, which we still call $T^*$ for simplicity, takes lines inside $E$ to lines inside $E$ since $T^*$ takes planes through the origin to planes through the origin. This leads us to consider a “linear” version of the Farey diagram in which each circular arc of the original Farey diagram is replaced by a straight line segment joining the two endpoints of the circular arc. These line segments divide the interior of $E$ into triangles, just as the original Farey diagram divides the disk into curvilinear triangles. The transformation $T^*$ takes each of these triangles onto another triangle, analogous to the way that linear fractional transformations provide symmetries of the original Farey diagram.

Suppose we divide each triangle of the linear Farey diagram into six smaller triangles as in the figure at the right. The transformation $T^*$ takes each of these small triangles onto another small triangle since it takes lines to lines. One of these small triangles is the triangle defined by the inequalities $0 \leq b \leq a \leq c$ that we considered earlier. The fact that every positive primitive elliptic form is equivalent to exactly one reduced form, corresponding to a rational point in this special triangle, is now visible geometrically as the fact that there is always exactly one transformation $T^*$ taking a given small triangle in the subdivided linear Farey diagram to this one special small triangle.

Elliptic forms whose topograph contains a source edge are equivalent to forms $ax^2 + cy^2$ so these are the forms corresponding to rational points on the edges of the linear Farey diagram, shown in black in the figure above. These are the forms whose topograph has a symmetry reflecting across a line perpendicular to the source edge. (This line is just the edge in the Farey diagram containing the given form.) The other type of reflectional symmetry in the topograph of an elliptic form is reflection across an edge of the topograph. Forms with this sort of symmetry correspond to rational
points in the blue and green edges in the preceding figure, the edges we added to subdivide the Farey diagram into the smaller triangles. The blue and green edges are distinguished by whether the two equal values of the form in the three regions surrounding the source vertex occur for the smallest value of the form (blue edges) or the next-to-smallest value (green edges). Note that the blue edges form the dual tree of the Farey diagram.

Let us now turn our attention to hyperbolic and 0-hyperbolic forms, which correspond to integer lattice points that lie outside the two cones. As a preliminary observation, note that for a point \((a, b, c)\) outside the double cone there are exactly two planes in \(\mathbb{R}^3\) that are tangent to the double cone and pass through \((a, b, c)\). Each of these planes is tangent to the double cone along a whole line through the origin. The two tangent planes through \((a, b, c)\) are determined by their intersection with the plane \(a + c = 1\), which consists of two lines tangent to the ellipse \(E\). These two lines can either intersect or be parallel. The latter possibility occurs when the point \((a, b, c)\) lies in the plane \(a + c = 0\), so the two tangent planes intersect in a line in this plane.

As a simple example, if the point \((a, b, c)\) we start with happens to lie on the \(b\) axis, then the tangent planes are the \(ab\) plane and the \(bc\) plane. These intersect the plane \(a + c = 1\) in the two vertical tangent lines to the ellipse \(E\).

Our goal will be to show the following:

**Proposition 5.10.** Let \(Q(x, y) = ax^2 + bxy + cy^2\) be a form of positive discriminant, either hyperbolic or 0-hyperbolic. Then the two points where the tangent lines to \(E\) determined by \((a, b, c)\) touch \(E\) are the points diametrically opposite the two points that are the endpoints of the separator line in the topograph of \(Q\) in the case that \(Q\) is hyperbolic, or the two points labeling the regions in the topograph of \(Q\) where \(Q\) takes the value zero in the case that \(Q\) is 0-hyperbolic.

**Proof:** We begin with a few preliminary remarks that will allow us to treat both the hyperbolic and 0-hyperbolic cases in the same way. A form \(Q(x, y) = ax^2 + bxy + cy^2\) of positive discriminant can always be factored as \((px + qy)(rx + sy)\) with \(p, q, r, s\) real numbers since if \(a = 0\) we have the factorization \(y(bx + cy)\) and if \(a \neq 0\) then the associated quadratic equation \(ax^2 + bx + c = 0\) has positive discriminant so it has two distinct real roots \(\alpha\) and \(\beta\), leading to the factorization \(ax^2 + bxy + cy^2 =\)
which can be rewritten as $(px + qy)(rx + sy)$ by incorporating $a$ into either factor. If $Q$ is hyperbolic then the discriminant is not a square and hence the factorization $(px + qy)(rx + sy)$ will involve coefficients that are quadratic irrationals. If $Q$ is 0-hyperbolic then the discriminant is a square so the roots $\alpha$ and $\beta$ are rational and we obtain a factorization of $Q$ as $(px + qy)(rx + sy)$ with rational coefficients. In fact we can take $p, q, r, s$ to be integers in this case since we know every 0-hyperbolic form is equivalent to a form $y(bx + cy)$ so we can obtain the given form $Q$ from $y(bx + cy)$ by replacing $x$ and $y$ by certain linear combinations $dx + ey$ and $fx + gy$ with integer coefficients $d, e, f, g$.

The points where the tangent planes touch the double cone correspond to forms of discriminant zero, with coefficients that may not be integers or even rational. A simple way to construct two such forms from a given form $Q = (px + qy)(rx + sy)$ is just to take the squares of the two linear factors, so we obtain the two forms $(px + qy)^2$ and $(rx + sy)^2$, each of discriminant zero. We will show that each of these two forms lies on the line of tangency for one of the two tangent planes determined by $Q$.

To do this for the case of $(px + qy)^2$ we consider the line $L$ in $\mathbb{R}^3$ passing through the two points corresponding to the forms $(px + qy)(rx + sy)$ and $(px + qy)^2$. We claim that $L$ consists of the forms

$$Q_t = (px + qy)[(1 - t)(rx + sy) + t(px + qy)]$$

as $t$ varies over all real numbers. When $t = 0$ or $t = 1$ we obtain the two forms $Q_0 = (px + qy)(rx + sy)$ and $Q_1 = (px + qy)^2$ so these forms lie on $L$. Also, we can see that the forms $Q_t$ do form a straight line in $\mathbb{R}^3$ by rewriting the formula for $Q_t$ in $(a, b, c)$ coordinates, where it becomes:

$$(a, b, c) = (pr(1 - t) + p^2t, (ps + qr)(1 - t) + 2pqt, qs(1 - t) + q^2t)$$

This defines a line since $p, q, r, s$ are constants, so each coordinate is a linear function of $t$. Since the forms $Q_t$ factor as the product of two linear factors, they have non-negative discriminant for all $t$. This means that $L$ does not go into the interior of either cone. It also does not pass through the origin since if it did, it would have to be a subset of the double cone since it contains the form $Q_1$ which lies in the double cone. From these facts we deduce that $L$ must be a tangent line to the double cone. Hence the plane containing $L$ and the origin must be tangent to the double cone along the line containing the origin and $Q_1$. The same reasoning shows that the other tangent plane that passes through $(px + qy)(rx + sy)$ intersects the double cone along the line containing the origin and $(rx + sy)^2$.

The labels of the points of $E$ corresponding to the forms $(px + qy)^2$ and $(rx + sy)^2$ are $p/q$ and $r/s$ according to the convention we have adopted. On the other hand, when the form $(px + qy)(rx + sy)$ is hyperbolic the ends of the separator line in its topograph are at the two points where this form is zero, which occur when $x/y$
is \(-q/p\) and \(-s/r\). These are the negative reciprocals of the previous two points \(p/q\) and \(r/s\) so they are the diametrically opposite points in \(E\). Similarly when \((px + qy)(rx + sy)\) is 0-hyperbolic the vertices of the Farey diagram where it is zero are at \(-q/p\) and \(-s/r\), again diametrically opposite \(p/q\) and \(r/s\).

It might have been nicer if the statement of the previous proposition did not involve passing to diametrically opposite points, but to achieve this we would have had to use a different rule for labeling the points of \(E\), with the label \(p/q\) corresponding to the form \((qx - py)^2\) instead of \((px + qy)^2\). This 180 degree rotation of the labels would put the negative labels in the upper half of \(E\) rather than the lower half, which doesn’t seem like such a good idea.

Next let us investigate how hyperbolic and 0-hyperbolic forms are distributed over the lattice points outside the double cone \(b^2 - 4ac = 0\). This is easier to visualize if we project such points radially into the plane \(a + c = 1\). This only works for forms \(ax^2 + bxy + cy^2\) with \(a + c > 0\), but the forms with \(a + c < 0\) are just the negatives of these so they give nothing essentially new. The forms with \(a + c = 0\) will be covered after we deal with those with \(a + c > 0\).

Forms with \(a + c > 0\) that are hyperbolic or 0-hyperbolic correspond via radial projection to points in the plane \(a + c = 1\) outside the ellipse \(E\). As we have seen, each such point determines a pair of tangent lines to \(E\) intersecting at the given point.

For a 0-hyperbolic form \((px + qy)(rx + sy)\) the points of tangency in \(E\) have rational labels \(p/q\) and \(r/s\). We know that every 0-hyperbolic form is equivalent to a form \(y(rx + sy)\) with \(a = 0\), so \(p/q = 0/1\) and one line of tangency is the vertical line tangent to \(E\) on the right side. The form \(y(rx + sy)\) corresponds to the point \((0,r,s)\) in the plane \(a = 0\) tangent to the double cone. Projecting radially into the vertical tangent line to \(E\), we obtain the points \((0,r/s,1)\), where \(r/s\) is an arbitrary rational number. Thus 0-hyperbolic forms are dense in this vertical tangent line to \(E\). Choosing any rational number \(r/s\), the other tangent line for the form \(y(rx + sy)\) is tangent to \(E\) at the point labeled \(r/s\).

An arbitrary 0-hyperbolic form \((px + qy)(rx + sy)\) is obtained from one with \(p/q = 0/1\) by applying a linear fractional transformation \(T\) taking \(0/1\) to \(p/q\), so the vertical tangent line to \(E\) at \(0/1\) is taken to the tangent line at \(p/q\), and the dense set of 0-hyperbolic forms in the vertical tangent line is taken to a dense set of 0-hyperbolic forms in the tangent line at \(p/q\). Thus we see that the 0-hyperbolic forms in the plane \(a + c = 1\) consist of all the rational points on all the tangent lines to \(E\) at rational points \(p/q\) of \(E\).

In the case of a hyperbolic form \(ax^2 + bxy + cy^2\) with \(a + c > 0\) the two tangent lines intersect \(E\) at a pair of conjugate quadratic irrationals, the negative reciprocals of the roots \(\alpha\) and \(\overline{\alpha}\) of the equation \(ax^2 + bx + c = 0\). Since \(\alpha\) determines \(\overline{\alpha}\) uniquely, one tangent line determines the other uniquely, unlike the situation for 0-hyperbolic
forms whose rational tangency points \( p/q \) and \( r/s \) can be varied independently. A consequence of this uniqueness for hyperbolic forms is that each of the two tangent lines contains only one rational point, the intersection point of the two lines, since any other rational point would correspond to another form having one of its tangent lines the same as for \( ax^2 + bxy + cy^2 \) and the other tangent line different, contradicting the previous observation that each tangent line for a hyperbolic form determines the other. (The hypothetical second form would also be hyperbolic since the common tangency point for the two forms is not a rational point on \( E \).)

The points in the plane \( a + c = 1 \) that correspond to 0-hyperbolic forms are dense in the region of this plane outside \( E \) since for an arbitrary point in this region we can first take the two tangent lines to \( E \) through this point and then take a pair of nearby lines that are tangent at rational points of \( E \) since points in \( E \) with rational labels are dense in \( E \). It is also true that points in the plane \( a + c = 1 \) that correspond to hyperbolic forms are dense in the region outside \( E \). To see this we can proceed in two steps. First consider the case of a point in this region whose two tangent lines to \( E \) are tangent at irrational points of \( E \). These two irrational points are the endpoints of an infinite strip in the Farey diagram that need not be periodic. However we can approximate this strip by a periodic strip by taking a long finite segment of the infinite strip and then repeating this periodically at each end. This means that the given point in the region outside \( E \) lies arbitrarily close to points corresponding to hyperbolic forms. Finally, a completely arbitrary point in the region outside \( E \) can be approximated by points whose tangent lines to \( E \) touch \( E \) at irrational points since irrational numbers are dense in real numbers.

It remains to consider hyperbolic and 0-hyperbolic forms \((px + qy)(rx + sy)\) corresponding to points \((a, b, c)\) in the plane \( a + c = 0 \). Such a form determines a line through the origin in this plane, and the tangent planes to the double cone that intersect in this line intersect the plane \( a + c = 1 \) in two parallel lines tangent to \( E \) at two diametrically opposite points \( p/q \) and \( -q/p \). Thus the form is \((px + qy)(qx - py)\), up to a constant multiple. If \( p/q \) is rational this is a 0-hyperbolic form. Examples are:

- \( xy \) with vertical tangents to \( E \) at 1/0 and 0/1.
- \( x^2 - y^2 = (x + y)(x - y) \) with horizontal tangents to \( E \) at 1/1 and \(-1/1\).
- \( 2x^2 - 3xy - 2y^2 = (2x + y)(x - 2y) \) with parallel tangents at 2/1 and \(-1/2\).

If \( p/q \) and \(-q/p \) are conjugate quadratic irrationals then we have a hyperbolic form \( ax^2 + bxy + cy^2 = a(x - \alpha)(x - \overline{\alpha}) \) where \( \alpha \overline{\alpha} = -1 \) since \( c = -a \) when \( a + c = 0 \). Thus \( \alpha \) and \( \overline{\alpha} \) are negative reciprocals of each other that are interchanged by 180
degree rotation of $E$. As examples we have:

$$x^2 + xy - y^2 = \left(x - \frac{-1 + \sqrt{5}}{2}y\right)\left(x - \frac{-1 - \sqrt{5}}{2}y\right)$$

$$2x^2 + xy - 2y^2 = 2\left(x - \frac{-1 + \sqrt{17}}{4}y\right)\left(x - \frac{-1 - \sqrt{17}}{4}y\right)$$

One can consider a pair of parallel tangent lines to $E$ as the limit of a pair of intersecting tangents where the point of intersection moves farther and farther away from $E$ in a certain direction which becomes the direction of the pair of parallel tangents.
Exercises

1. Find a hyperbolic quadratic form whose periodic separator line has the following pattern:

```
  |   |   |   |   |   |   |   |   |
  |   |   |   |   |   |   |   |   |
  |   |   |   |   |   |   |   |   |
  |   |   |   |   |   |   |   |   |
  |   |   |   |   |   |   |   |   |
  |   |   |   |   |   |   |   |   |
```

2. (a) Find two elliptic forms \( ax^2 + cy^2 \) that have the same discriminant but take on different sets of values. Draw enough of the topographs of the two forms to make it apparent that they do not have exactly the same sets of values. Include the source vertex or source edge in the topographs. (Remember that the topograph only shows the values \( Q(x, y) \) for primitive pairs \((x, y)\).)

(b) Do the same thing with hyperbolic forms \( ax^2 + cy^2 \). Include the separator lines in their topographs.

3. (a) Show the quadratic form \( Q(x, y) = 92x^2 - 74xy + 15y^2 \) is elliptic by computing its discriminant.

(b) Find the source vertex or edge in the topograph of this form.

(c) Using the topograph of this form, find all the integer solutions of \( 92x^2 - 74xy + 15y^2 = 60 \), and explain why your list of solutions is a complete list. (There are exactly four pairs of solutions \( \pm (x, y) \), three of which will be visible in the topograph.)

4. (a) Show that if a quadratic form \( Q(x, y) = ax^2 + bxy + cy^2 \) can be factored as a product \((Ax + By)(Cx + Dy)\) with \( A, B, C, D \) integers, then \( Q \) takes the value 0 at some pair of integers \((x, y) \neq (0, 0)\), hence \( Q \) must be either 0-hyperbolic or parabolic. Show also, by a direct calculation, that the discriminant of this form is a square.

(b) Find a 0-hyperbolic form \( Q(x, y) \) such that \( Q(1, 5) = 0 \) and \( Q(7, 2) = 0 \) and draw a portion of the topograph of \( Q \) that includes the two regions where \( Q = 0 \).

5. Determine the number of equivalence classes of quadratic forms of discriminant \( \Delta = 120 \) and list one form from each equivalence class.

6. Do the same thing for \( \Delta = 61 \).

7. (a) Find the smallest positive nonsquare discriminant for which there is more than one equivalence class of forms of that discriminant. (In particular, show that all smaller discriminants have only one equivalence class.)

(b) Find the smallest positive nonsquare discriminant for which there are two inequivalent forms of that discriminant, neither of which is simply the negative of the other.

8. (a) For positive elliptic forms of discriminant \( \Delta = -D \), verify that the smallest value of \( D \) for which there are at least two inequivalent forms of discriminant \( -D \) is \( D = 12 \).
(b) If we add the requirement that all forms under consideration are primitive, then what is the smallest \( D \)?

9. Determine all the equivalence classes of positive elliptic forms of discriminants \(-67\), \(-104\), and \(-347\).

10. (a) Determine all the equivalence classes of 0-hyperbolic forms with discriminant 49.
(b) Determine which equivalence class in part (a) each of the forms \( Q(x, y) = 7xy - p y^2 \) for \( p = 0, 1, 2, 3, 4, 5, 6 \) belongs to.

11. Show that if a form takes the same value on two adjacent regions of its topograph, then these regions are both adjacent to the source vertex or edge when the form is elliptic, or both lie along the separator line when the form is hyperbolic.

12. Show that a positive nonsquare number is the sum of two squares at least one of which is even if and only if it is the discriminant of some hyperbolic form whose topograph has a rotational skew symmetry.

13. In this extended exercise the goal will be to show that the only negative even discriminants with class number 1 are \(-4\), \(-8\), \(-12\), \(-16\), and \(-28\). (Note that of these, only \(-4\) and \(-8\) are fundamental discriminants.) The strategy will be to exhibit an explicit reduced primitive form \( Q \) different from the principal form \( x^2 + dy^2 \) for each discriminant \(-4d\) with \( d > 4 \) except \( d = 7 \). This will be done by breaking the problem into several cases, where in each case a form \( Q \) will be given and you are to show that this form has the desired properties, namely it is of discriminant \(-4d\), primitive, reduced, and different from the principal form. You should also check that the cases considered cover all possibilities.

(a) Suppose \( d \) is not a prime power. Then it can be factored as \( d = ac \) where \( 1 < a < c \) and \( a \) and \( c \) are coprime. In this case let \( Q \) be the form \( ax^2 + cy^2 \).
(b) The form \( ax^2 + 2xy + cy^2 \) will work provided that \( d + 1 \) factors as \( d + 1 = ac \) where \( a \) and \( c \) are coprime and \( 1 < a < c \). If \( d \) is odd, for example a power of an odd prime, then \( d + 1 \) is even so it has such a factorization \( d + 1 = ac \) unless \( d + 1 = 2^n \).
(c) If \( d = 2^n \) the cases we need to consider are \( n \geq 3 \) since we assume \( d > 4 \). When \( n = 3 \) take \( Q \) to be \( 3x^2 + 2xy + 3y^2 \) and when \( n \geq 4 \) take \( Q \) to be \( 4x^2 + 4xy + (2^{n-2} + 1)y^2 \).
(d) When \( d + 1 = 2^n \) the cases of interest are \( n \geq 3 \). When \( n = 3 \) we have \( d = 7 \) which is one of the allowed exceptions with class number 1. When \( n = 4 \) we have \( d = 15 \) and \( 3x^2 + 5y^2 \) works as in part (a). When \( n = 5 \) we have \( d = 31 \) and we take the form \( 5x^2 + 4xy + 7y^2 \). When \( n \geq 6 \) we use the form \( 8x^2 + 6xy + (2^{n-3} + 1)y^2 \).

14. Show that the number of equivalence classes of forms of discriminant 45 with mirror symmetry is not a power of 2 if nonprimitive as well as primitive forms are allowed. (Compare this with Theorem 5.8.)
15. Show that the class number for discriminant $\Delta = q^2 > 1$ is $\varphi(q)$ where $\varphi(q)$ is the number of positive integers less than $q$ and coprime to $q$.

16. Show that the topograph of a primitive 0-hyperbolic form $qxy - py^2$ has mirror symmetry exactly when $p^2 \equiv 1 \pmod{q}$, and has rotational skew symmetry exactly when $p^2 \equiv -1 \pmod{q}$. (See the discussion in Chapter 2 about the relation between the continued fraction for $p/q$ and the continued fraction obtained by reversing the order of the terms.)
Chapter 6. Representations by Quadratic Forms

With the various things we have learned about quadratic forms so far, let us return to the basic representation problem of determining what values a given form \( Q(x, y) = ax^2 + bxy + cy^2 \) can take on when \( x \) and \( y \) are integers, or in other words, which numbers can be represented in the form \( ax^2 + bxy + cy^2 \). Remember that it suffices to restrict attention to the values in the topograph since these are the values for primitive pairs \((x, y)\), and to get all other values one just multiplies the values in the topograph by arbitrary squares. With this in mind we will adopt the following convention in the rest of the book:

*When we say that a form \( Q \) represents a number \( n \) we mean that \( n = Q(x, y) \) for some primitive pair of integers \((x, y) \neq (0, 0)\).*

This differs from the traditional terminology in which any solution of \( n = Q(x, y) \) is called a representation of \( n \), without requiring \((x, y)\) to be a primitive pair, and when \((x, y)\) is primitive it is called a proper or primitive representation of \( n \). However, since we will rarely consider the case that \((x, y)\) is not a primitive pair, it will save many words not to have to insert the extra modifier for every representation.

We will focus on forms that are either elliptic or hyperbolic, as these are the most interesting cases.

1. Three Levels of Complexity

In this section we will look at a series of examples to try to narrow down what sort of answer one could hope to obtain for the representation problem. The end result will be a reasonable guess that will be verified in the rest of this chapter and the next one, at least for fundamental discriminants. For nonfundamental discriminants there is sometimes a small extra wrinkle that seems to be rather subtle and more difficult to analyze.

As a first example let us try to find a general pattern in the values of the form \( x^2 + y^2 \). In view of the symmetry of the topograph for this form it suffices to look just in the first quadrant of the topograph. A piece of this quadrant is shown in the figure at the right, somewhat distorted to squeeze more numbers into the picture. What is shown is all the numbers in the topograph that are less than 100. At first glance it may be hard to detect any patterns here. Both even and odd numbers occur, but none of the even numbers are divisible by 4 so they are all twice an odd number, and in fact an odd number that appears in the topograph. Considering the odd numbers, one notices they are all congruent to 1 mod 4 and not 3 mod 4, which is the other
possibility for odd numbers. On the other hand, not all odd numbers congruent to 1 mod 4 appear in the topograph. Up to 100, the ones that are missing are 9, 21, 33, 45, 49, 57, 69, 77, 81, and 93. Each of these has at least one prime factor congruent to 3 mod 4, while all the odd numbers that do appear have all their prime factors congruent to 1 mod 4. Conversely, all products of primes congruent to 1 mod 4 are in the topograph.

This leads us to guess that the following statements might be true:

**Conjecture.** The numbers that appear in the topograph of \( x^2 + y^2 \) are precisely the numbers \( n = 2^a p_1 p_2 \cdots p_k \) where \( a \leq 1 \) and each \( p_i \) is a prime congruent to 1 mod 4. Consequently the values of the quadratic form \( Q(x,y) = x^2 + y^2 \) as \( x \) and \( y \) range over all integers (not just the primitive pairs) are exactly the numbers \( n = m^2 p_1 p_2 \cdots p_k \) where \( m \) is an arbitrary integer and each \( p_i \) is either 2 or a prime congruent to 1 mod 4.

In both statements the index \( k \) denoting the number of prime factors \( p_i \) is allowed to be zero as well as any positive integer. The restriction \( a \leq 1 \) in the first statement disappears in the second statement where even powers of 2 can be absorbed into the factor \( m^2 \).

We will prove the conjecture later in the chapter. A weaker form of the conjecture can be proved just by considering congruences mod 4 as follows. An even number squared is congruent to 0 mod 4 and an odd number squared is congruent to 1 mod 4, so \( x^2 + y^2 \) must be congruent to 0, 1, or 2 mod 4. Moreover, the only way that \( x^2 + y^2 \) can be 0 mod 4 is for both \( x \) and \( y \) to be even, which cannot happen for primitive pairs. Thus all numbers in the topograph must be congruent to 1 or 2 mod 4. This says that the odd numbers in the topograph are congruent to 1 mod 4 and the even numbers are each twice an odd number.

However, these simple observations say nothing about the role played by primes and prime factorizations, nor do they include any positive assertions about which numbers actually are represented by \( x^2 + y^2 \). It definitely takes more work to show for example that every prime \( p = 4k+1 \) can be represented as the sum of two squares.

Let us look at a second example to see whether the same sorts of patterns occur, this time for the form \( Q(x,y) = x^2 + 2y^2 \). Here is a portion of its topograph showing all values less than 100, with the lower half of the topograph omitted since it is just the mirror image of the upper half:
Again the even values are just the doubles of the odd values. The odd prime values are 3, 11, 17, 19, 41, 43, 59, 67, 73, 83, 89, 97 and the other odd values are all the products of these primes. The odd prime values are not determined by their values mod 4 in this case, but instead by their values mod 8 since the primes we just listed are exactly the primes less than 100 that are congruent to 1 or 3 mod 8. Apart from this change, the answer to the representation problem for $x^2 + 2y^2$ is completely analogous to the answer for $x^2 + y^2$. Namely, the numbers represented by $x^2 + 2y^2$ are the numbers $n = 2^a p_1 p_2 \cdots p_k$ with $a \leq 1$ and each $p_i$ a prime congruent to 1 or 3 mod 8. Using congruences mod 8 we could easily prove the weaker statement that all numbers represented by $x^2 + 2y^2$ must be congruent to 1, 2, 3, or 6 mod 8, so all odd numbers in the topograph must be congruent to 1 or 3 mod 8 and all even numbers must be twice an odd number.

These two examples were elliptic forms, but the same sort of behavior can occur for hyperbolic forms as we see in the next example, the form $x^2 - 2y^2$. The negative values of this form happen to be just the negatives of the positive values, so we need only show the positive values in the topograph:

Here the primes that occur are 2 and primes congruent to ±1 modulo 8. The non-prime values that occur are the products of primes congruent to ±1 modulo 8 and twice these products. Again there is a weaker statement that can be proved using just congruences mod 8.
In these three examples the guiding principle was to look at prime factorizations and at primes modulo certain numbers, the numbers 4, 8, and 8 in the three cases. Notice that these numbers are just the absolute values of the discriminants \(-4\), \(-8\), and 8. Looking at primes modulo \(|\Delta|\) turns out to be a key idea for all quadratic forms.

Another example of the same sort is the form \(x^2 + xy + y^2\) of discriminant \(-3\). This time it is the prime 3 that plays a special role rather than 2.

We only have to draw one-sixth of the topograph because of all the symmetries. Notice that all the values are odd, so the prime 2 plays no role here. Since the discriminant is \(-3\) we are led to consider congruences mod 3. The primes in the topograph are 3 and the primes congruent to 1 mod 3 (which in particular excludes the prime 2), namely the primes 7, 13, 19, 31, 37, 43, 61, 67, 73, 79, 97. The nonprime values are the products of these primes with the restriction that the prime 3 never has an exponent greater than 1. This is analogous to the prime 2 never having an exponent greater than 1 in the preceding examples. In all four examples the “special” primes whose exponents are restricted are just the prime divisors of the discriminant. This is a general phenomenon, that primes dividing the discriminant behave differently from primes that do not divide the discriminant.

A special feature of the discriminants \(-4\), \(-8\), 8, and \(-3\) is that in each case all forms of that discriminant are equivalent. We will see that the representation problem always has the same type of answer for discriminants with a single equivalence class of forms.

Before going on to the next level of complexity let us digress to describe a nice property that forms of the first level of complexity have. As we know, if an equation \(Q(x,y) = n\) has an integer solution \((x,y)\) then so does \(Q(x,y) = m^2 n\) for any integer \(m\). The converse is not always true however. For example the equation \(2x^2 + 7y^2 = 9\) has the solution \((x,y) = (1,1)\) but \(2x^2 + 7y^2 = 1\) obviously has no solution with \(x\) and \(y\) integers. Nevertheless, this converse property does hold for forms such as those in the preceding four examples where the numbers \(n\) for which
$Q(x, y) = n$ has an integer solution are exactly the numbers that can be factored as $n = m^2 p_1 p_2 \cdots p_k$ for primes $p_i$ satisfying certain conditions and $m$ an arbitrary integer. This is because if a number $n$ has a factorization of this type then we can cancel any square factor of $n$ and the result still has a factorization of the same type.

Let us apply this “square-cancellation” property in the case of the form $x^2 + y^2$ to determine the numbers $n$ such that the circle $x^2 + y^2 = n$ contains a rational point, and hence, as in Chapter 0, an infinite dense set of rational points. Suppose first that the circle $x^2 + y^2 = n$ contains a rational point, so after putting the two coordinates over a common denominator the point is $(x, y) = (\frac{a}{c}, \frac{b}{c})$. The equation $x^2 + y^2 = n$ then becomes $a^2 + b^2 = c^2 n$. This means that the equation $x^2 + y^2 = c^2 n$ has an integer solution. Then the square-cancellation property implies that the original equation $x^2 + y^2 = n$ has an integer solution. Thus we see that if there are rational points on the circle $x^2 + y^2 = n$ then there are integer points on it. This is not something that is true for all quadratic curves, as shown by the example of the ellipse $2x^2 + 7y^2 = 1$ which has rational points such as $(\frac{1}{3}, \frac{1}{3})$ but no integer points.

From the solution to the representation problem for $x^2 + y^2$ we deduce that the circle $x^2 + y^2 = n$ contains rational points exactly when $n = m^2 p_1 p_2 \cdots p_k$ where $m$ is an arbitrary integer and each $p_i$ is either 2 or a prime congruent to 1 mod 4. The first few values of $n$ satisfying this condition are 1, 2, 4, 5, 8, 9, 10, 13, 16, 17, 18, 20, \cdots.

Now let us look at some examples with a second level of complexity. First consider the form $x^2 - 10y^2$ whose positive values less than 100 are shown in the following topograph:

$$Q_1(x, y) = x^2 - 10y^2$$

There is no need to show any more of the negative values since these will just be the negatives of the positive values. The prime values less than 100 are 31, 41, 71, 79, 89. These are the primes congruent to ±1 or ±9 modulo 40, the discriminant. However, in contrast to what happened in the previous examples, there are many nonprime values that are not products of these prime values. The prime factors of these nonprime values are 2, 3, 5, 13, 37, 43, none of which occur in the topograph. Rather miraculously, these prime values are realized instead by another form $2x^2 - 5y^2$ having the same discriminant as $x^2 - 10y^2$. Here is the topograph of this companion form:
Again the negative values are just the negatives of the positive values. The prime values this form takes on are 2 and 5, which are the prime divisors of the discriminant 40, along with primes congruent to ±3 and ±13 modulo 40, namely 3, 13, 37, 43, 53, 67, 83.

The congruence conditions for primes not dividing the discriminant are in fact satisfied by all numbers coprime to the discriminant that appear in the topographs. Thus for $Q_1$ all numbers in the topograph not divisible by 2 or 5 are congruent to ±1 or ±9 mod 40, while for $Q_2$ the numbers in the topograph not divisible by 2 or 5 are congruent to ±3 and ±13 mod 40. However, for nonprimes these conditions do not always guarantee representability. For example, 49 is coprime to 40 and congruent to 9 mod 40 but 49 does not appear in the topograph of $Q_1$.

Apart from the primes 2 and 5 that divide the discriminant 40, the possible values of primes modulo 40 are ±1, ±3, ±7, ±9, ±11, ±13, ±17, ±19 since even numbers and multiples of 5 are excluded. There are 16 different congruence classes here, and exactly half of them, 8, are realized by one or the other of the two forms $x^2 - 10y^2$ and $2x^2 - 5y^2$, with 4 classes realized by each form. The other 8 congruence classes are not realized by any form of discriminant 40 since every form of discriminant 40 is equivalent to one of the two forms $x^2 - 10y^2$ or $2x^2 - 5y^2$, as is easily checked by the methods from the previous chapter.

This turns out to be a general phenomenon valid for all elliptic and hyperbolic forms: If one excludes the primes that divide the discriminant, then the prime values of quadratic forms of that discriminant are exactly the primes in half of the possible congruence classes modulo the discriminant. This will be proved in Corollary 6.13.

After further examination of the topographs of the two forms $Q_1 = x^2 - 10y^2$ and $Q_2 = 2x^2 - 5y^2$ it seems that the following statements should be true:

**Conjecture.** The positive numbers represented by either $Q_1$ or $Q_2$ are exactly the products $2^a 5^b p_1 p_2 \cdots p_k$ where $a, b \leq 1$ and each $p_i$ is a prime congruent to ±1, ±3, ±9, or ±13 mod 40. Furthermore, one can determine which form will represent such a product by the rule that if the number of terms in the product that are represented by $Q_2$ (namely 2, 5, and $p_i \equiv \pm 3$ or $\pm 13$ mod 40) is even, then the number is represented by $Q_1$ and if it is odd then the number is represented by $Q_2$. 
For example, the topograph of $Q_1$ contains the even powers of 3 while the topograph of $Q_2$ contains the odd powers. Another consequence is that the even values in one topograph are just the doubles of the odd values in the other topograph.

This characterization of numbers represented by these two forms also implies that no number is represented by both $Q_1$ and $Q_2$. However, for some discriminants it is possible for two non-equivalent forms of that discriminant to represent the same nonzero number, as we will see.

The preceding Conjecture will be proved piece by piece as we gradually develop the necessary general theory. We will focus first on determining which primes are represented by $Q_1$ and which by $Q_2$, where Proposition 6.10 will provide the final answer. For nonprimes the first sentence of the Conjecture will follow from Theorem 6.19, while the second sentence will use results from Chapter 7.

One small part of the Conjecture is simple and can be proved right now, the restriction on the exponents of primes dividing the discriminant.

**Proposition 6.1.** Suppose the prime $p$ divides the discriminant $\Delta$ but not the conductor of $\Delta$. Then no form of discriminant $\Delta$ represents $p^2$ or any number divisible by $p^2$.

Recall that the conductor for discriminant $\Delta$ is the largest positive number $d$ such that $\Delta = d^2 \Delta'$ for some discriminant $\Delta'$. This $\Delta'$ is then a fundamental discriminant. In the preceding examples $\Delta$ was itself a fundamental discriminant so the conductor was 1 and thus Proposition 6.1 applies to all primes dividing the discriminant. Primes dividing the conductor can sometimes be genuine exceptions. For example the form $x^2 + 7y^2$ of discriminant $-28$ with conductor 2 represents 8, 16, 32, and 64 when $(x, y)$ is $(1, 1), (3, 1), (5, 1),$ and $(1, 3)$. In fact $x^2 + 7y^2$ represents all higher powers of 2 as well, as we will see in Chapter 8.

**Proof:** Suppose a form $Q$ of discriminant $\Delta$ represents a number $p^2m$ with $m$ an integer. Then $Q$ is equivalent to a form $p^2mx^2 + bxy + cy^2$ with $\Delta = b^2 - 4p^2mc$.

Suppose that $p$ is a prime dividing $\Delta$ but not the conductor. Then $p$ divides $b^2 - 4p^2mc$ so $p$ divides $b^2$ and hence also $b$, so $p^2$ divides $b^2$. Therefore $p^2$ also divides $\Delta$ and we have $\Delta = p^2 \Delta'$ for some integer $\Delta'$. Consider first the case that $p$ is odd. Then $p^2 \equiv 1 \mod 4$ so $\Delta \equiv \Delta' \mod 4$. This means that $\Delta'$ is also a discriminant, so by the definition of the conductor, $p$ divides the conductor, contradicting what we have assumed. This contradiction shows that $Q$ cannot represent any number $p^2m$.

In the case that $p = 2$, the assumption that $p$ divides $\Delta$ means that $\Delta$ is even and hence so is $b$. The discriminant equation is now $\Delta = b^2 - 4\cdot 2^2mc$ so $\Delta \equiv b^2 \mod 16$. The only squares of even numbers mod 16 are 0 and 4, as one sees by checking $0^2$, $(\pm 2)^2$, $(\pm 4)^2$, $(\pm 6)^2$, and $8^2$, so $\Delta$ is either $16k = 4(4k)$ or $16k + 4 = 4(4k + 1)$. In both cases $\Delta$ is 4 times a discriminant so 2 divides the conductor and we have the same contradiction as before. 

$\square$
There is another way of formulating the second half of the preceding Conjecture that is quite illuminating and turns out to be generalizable to all discriminants, though the generalization is not as simple as this special case. Restricting attention just to numbers that are represented by at least one of the two forms, then the second half of the Conjecture can be reformulated as the following three statements:

1. The product of two numbers represented by $Q_1$ is again represented by $Q_1$.
2. The product of two numbers represented by $Q_2$ is represented by $Q_1$.
3. The product of a number represented by $Q_1$ with a number represented by $Q_2$ is represented by $Q_2$.

Here we must assume that the products in question are not divisible by $2^2$ or $5^2$, as in the Conjecture.

An abbreviated way of writing the statements (1)-(3) is by the formulas $Q_1Q_1 = Q_1$, $Q_2Q_2 = Q_1$, and $Q_1Q_2 = Q_2$. One can see that these are formally the same as the rules for addition of integers mod 2: $0 + 0 = 0$, $1 + 1 = 0$, and $0 + 1 = 1$. The two formulas $Q_1Q_1 = Q_1$ and $Q_1Q_2 = Q_2$ say that $Q_1$ serves as an identity element “1” for this multiplication operation, and then the formula $Q_2Q_2 = Q_1$ can be interpreted as saying that $Q_2$ is equal to its own inverse, so $Q_2 = Q_2^{-1}$.

Let us look at another example where the representation problem has an answer that is qualitatively similar to the preceding example but just a little more complicated, the case of discriminant $-84$. Here there are twice as many equivalence classes of forms, four instead of two, with topographs shown below.

\[
Q_1(x, y) = x^2 + 21y^2
\]
\[
Q_2(x, y) = 3x^2 + 7y^2
\]
\[
Q_3(x, y) = 2x^2 + 2xy + 11y^2
\]
\[
Q_4(x, y) = 5x^2 + 4xy + 5y^2
\]
The primes dividing the discriminant $-84$ are 2, 3, and 7, and these primes are each represented by one of the forms. For the remaining primes, we will show later in the chapter that the primes represented by each form are as follows:

- For $Q_1$ the primes $p \equiv 1, 25, 37 \mod 84$.
- For $Q_2$ the primes $p \equiv 19, 31, 55 \mod 84$.
- For $Q_3$ the primes $p \equiv 11, 23, 71 \mod 84$.
- For $Q_4$ the primes $p \equiv 5, 17, 41 \mod 84$.

This agrees with what is shown in the four topographs above, and one could expand the topographs to get further evidence that these are the right answers. The non-primes in the topographs appear to be the products $2^a3^b7^c p_1 \cdots p_k$ with $a, b, c \leq 1$ and each $p_i$ one of the other primes represented by $Q_1, Q_2, Q_3$, or $Q_4$.

One can work out hypothetical rules for multiplying the forms by considering how products of two primes are represented. For example, 3 is represented by $Q_2$ and 11 is represented by $Q_3$, while their product $3 \cdot 11 = 33$ is represented by $Q_4$, so we might guess that $Q_2 Q_3 = Q_4$. Some other products that give the same conclusion are $3 \cdot 2 = 6$, $3 \cdot 23 = 69$, $7 \cdot 2 = 14$, $7 \cdot 11 = 77$, $31 \cdot 2 = 62$, etc. In the same way one can determine tentative rules for all the products $Q_i Q_j$. One finds:

- The principal form $Q_1$ acts as the identity, so $Q_1 Q_i = Q_i$ for each $i$.
- $Q_i Q_i = Q_1$ for each $i$ so each $Q_i$ equals its own inverse.
- The product of any two out of $Q_2$, $Q_3$, $Q_4$ is equal to the third.

These multiplication rules are formally identical to how one would add pairs $(m, n)$ of integers mod 2 by adding their two coordinates separately. The form $Q_1$ corresponds to the pair $(0, 0)$ and the first of the three rules above becomes the formula $(0, 0) + (m, n) = (m, n)$. The forms $Q_2$, $Q_3$, and $Q_4$ correspond to $(1, 0)$, $(0, 1)$, and $(1, 1)$ in any order, and then the second rule above becomes $(m, n) + (m, n) = (0, 0)$ which is valid for addition mod 2, while the third rule becomes the fact that the sum of any two of $(1, 0)$, $(0, 1)$, and $(1, 1)$ is equal to the third if we do addition mod 2.

The multiplication rules determine which form represents a given number $n$ by replacing each prime in the prime factorization of $n$ by the form $Q_i$ that represents it, then multiplying out the resulting product using the three multiplication rules, keeping in mind that 2, 3, and 7 can never occur with an exponent greater than 1. For example, for $n = 70 = 2 \cdot 5 \cdot 7$ we get the product $Q_3 Q_4 Q_2$ which equals $Q_1$ and so 70 is represented by $Q_1$, as the topograph shows. For $n = 66 = 2 \cdot 3 \cdot 11$ we get $Q_3 Q_2 Q_3 = Q_2$ and 66 is represented by $Q_2$. In general, for a number $n = 2^a3^b7^c p_1 \cdots p_k$ we can determine which form represents $n$ by the following steps. First compute the number $q_i$ of prime factors of $n$ represented by $Q_i$. Next compute the sum $q_1(0, 0) + q_2(1, 0) + q_3(0, 1) + q_4(1, 1) = (a_2 + a_4, a_3 + a_4)$ where $(0, 0), (1, 0), (0, 1), (1, 1)$ correspond to $Q_1, Q_2, Q_3, Q_4$ respectively. The resulting sum $(m, n)$ then tells which form represents $n$. 

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An interesting feature of all the forms at the first or second level of complexity that we have examined so far is that their topographs have mirror symmetry. This is actually a general phenomenon: Whenever all the forms of a given discriminant have mirror symmetry, then one can determine which primes are represented by each form just in terms of congruence conditions modulo the discriminant. And in fact this is the only time when congruences modulo the discriminant determine how primes are represented, at least if one restricts attention just to primitive forms. This will be shown in Corollary 6.18 later in the chapter. In Chapter 5 we called discriminants for which all primitive forms have mirror symmetry \textit{fully symmetric} discriminants, and we observed that they are unfortunately rather rare, with only 101 negative discriminants known to have this property, and probably no more.

Now we move on to the third level of complexity, illustrated by the case \( \Delta = -56 \) where there are three equivalence classes of forms:

\[
Q_1(x, y) = x^2 + 14y^2
\]

\[
Q_2(x, y) = 2x^2 + 7y^2
\]

\[
Q_3(x, y) = 3x^2 + 2xy + 5y^2
\]
Notice that the topograph of the form $Q_3 = 3x^2 + 2xy + 5y^2$ does not have mirror symmetry, so $Q_3$ counts twice when determining the class number for discriminant $-56$, which is therefore 4 rather than 3.

In the examples at the first two levels of complexity it was possible to determine which numbers are represented by a given form by looking at primes and which congruence classes they fall into modulo the discriminant. The primes represented by a given form were exactly the primes in certain congruence classes modulo the discriminant, and no number could be represented by two nonequivalent forms. These nice properties no longer hold in the case $\Delta = -56$, however. Here the primes 23 and 79 are congruent modulo 56, and yet 23 is represented by $Q_1 = x^2 + 14y^2$ since $Q_1(3,1) = 23$, while 79 is represented by $Q_2 = 2x^2 + 7y^2$ since $Q_2(6,1) = 79$. Also, some non-primes are represented by both $Q_1$ and $Q_2$. For example, $Q_1(1,1) = 15$ and $Q_2(2,1) = 15$.

Apart from the primes 2 and 7 that divide the discriminant $-56$, all other primes belong to the following 24 congruence classes modulo 56, corresponding to odd numbers less than 56 not divisible by 7:

$$\begin{array}{cccccccccccccccccccc}
\end{array}$$

The six congruence classes whose prime elements are represented by $Q_1$ or $Q_2$ are indicated by underlines, and the six congruence classes whose prime elements are represented by $Q_3$ are indicated by overlines. Primes not represented by any of the three forms are in the remaining 12 congruence classes.

The new thing that happens in this example is that one cannot tell whether a prime is represented by $Q_1$ or $Q_2$ just by considering congruence classes mod the discriminant. We saw this for the pair of primes 23 and 79, and another such pair visible in the topographs is 71 and 127. By extending the topographs we could find many more such pairs. One might try using congruences modulo some other number besides 56, but it is known that this does not help.

Congruences mod 56 suffice to tell which primes are represented by $Q_3$, but there is a different sort of novel behavior involving $Q_3$ when we look at representing products of primes. To illustrate this, observe that the primes 3 and 5 are represented by $Q_3$ but their product 15 is represented by both $Q_1$ and $Q_2$. This means there is some ambiguity about whether the product $Q_3Q_3$ should be $Q_1$ or $Q_2$. The same thing happens in fact for any pair of coprime numbers represented by $Q_3$.

For other products $Q_iQ_j$ there seems to be no ambiguity. The principal form $Q_1$ acts as the identity for multiplication, while $Q_2Q_2 = Q_1$ and $Q_2Q_3 = Q_3$, although this last formula is somewhat odd since it seems to imply that $Q_3$ does not have a multiplicative inverse since if it did, we could multiply by this inverse to get that $Q_2 = Q_1$, the identity for multiplication.

It turns out that there is a way out of these difficulties, discovered by Gauss. The troublesome form $Q_3$ is different from the other forms in this example and in the
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preceding examples in that it does not have mirror symmetry. Thus the equivalence class of \( Q_3 \) splits into two proper equivalence classes, with \( Q_3 \) having a mirror image form \( Q_4 = 3x^2 - 2xy + 5y^2 \) obtained from \( Q_3 \) by changing the sign of either \( x \) or \( y \) and hence changing the coefficient of \( xy \) to its negative. Using \( Q_4 \) we can then resolve the ambiguous product \( Q_3Q_3 \) by setting \( Q_3Q_3 = Q_2 = Q_4Q_4 \) and \( Q_3Q_4 = Q_1 \) so that \( Q_4 \) is the inverse of \( Q_3 \). This means that each \( Q_i \) has its inverse given by the mirror image topograph since \( Q_1 \) and \( Q_2 \) have mirror symmetry and equal their own inverses. The rigorous justification for the formulas \( Q_3Q_3 = Q_2 = Q_4Q_4 \) and \( Q_3Q_4 = Q_1 \) will come in Chapter 7, but for the moment one can check that these formulas are at least consistent with the topographs.

Since \( Q_3^2 = Q_2 \) we have \( Q_3^4 = Q_2^2 = Q_1 \). Multiplying the equation \( Q_3^4 = Q_1 \) by \( Q_4 \), the inverse of \( Q_3 \), gives \( Q_3^3 = Q_4 \). Thus all four proper equivalence classes of forms are powers of the single form \( Q_3 \) since \( Q_3^2 = Q_2 \), \( Q_3^3 = Q_4 \), and \( Q_3^4 = Q_1 \). This is corroborated by the representations of powers of 3 since 3 is represented by \( Q_3 \), \( 3^2 = Q_2 \), \( 3^3 = Q_4 \), and \( 3^4 = Q_1 \). Products of powers \( Q_3^i \) are computed by adding exponents mod 4 since \( Q_3^4 \) is the identity. Thus multiplication of the four forms is formally identical with addition of integers mod 4. The earlier doubtful formula \( Q_2Q_3 = Q_3 \) is resolved into the two formulas \( Q_2Q_3 = Q_4 \) and \( Q_2Q_4 = Q_3 \), which become \( Q_3^2Q_3 = Q_3^3 \) and \( Q_3^2Q_3^3 = Q_3^5 = Q_3 \).

The appearance of the same number in two different topographs is easy to explain now that we have two forms \( Q_3 \) and \( Q_4 \) representing exactly the same numbers. For example, to find all appearances of the number 15 = 3 · 5 in the topographs we observe that its prime factors 3 and 5 appear in the topographs of both \( Q_3 \) and \( Q_4 \) so 15 will appear in the topographs of \( Q_3Q_3 = Q_2 \), \( Q_3Q_4 = Q_1 \), and \( Q_4Q_4 = Q_2 \), although this last formula gives no new representations.

The procedure for finding which forms represent a number \( n = 2^a7^b p_1 \cdots p_k \) with \( a, b \leq 1 \) and primes \( p_i \) different from 2 or 7 is to replace each prime factor in this product by the form or forms \( Q_i \) that represent it, then multiply out the resulting product of forms \( Q_i \), which is most easily done by expressing each \( Q_i \) as the appropriate power of \( Q_3 \). There is also an extra condition that will be justified in Chapter 7: Whenever a prime \( p_i \) represented by \( Q_3 \) appears more than once in the prime factorization of \( n \), we should replace all of its appearances by \( Q_3 \) or all by \( Q_3^{-1} = Q_4 \). For example, the forms representing 18 = 2 · 3^2 are just the products \( Q_2Q_3^2 = Q_1 \) and \( Q_2Q_4^2 = Q_1 \) and not \( Q_2Q_3Q_4 = Q_2 \), as one can see in the topographs. Similarly, 9 = 3 · 3 is represented only by \( Q_3^2 = Q_2 = Q_4^2 \) and not by \( Q_3Q_4 = Q_1 \).

We will show in Chapter 7 that the set of proper equivalence classes of primitive forms of fixed discriminant always has a multiplication operation compatible with multiplying values of forms of that discriminant in the way illustrated by the preceding examples. This multiplication operation gives this set the structure of a group, that is, a set with an associative multiplication operation for which there is an element of
the set that functions as an identity for the multiplication, and such that each element
of the set has a multiplicative inverse in the set whose product with the given element
is the identity element. The set of proper equivalence classes of primitive forms with
this group structure is called the class group for the given discriminant. The identity
element is the class of the principal form, and the inverse of a class is obtained by
taking the mirror image topograph.

The class group has the additional property that the multiplication is commuta-
tive. This makes its algebraic structure much simpler than the typical noncommuta-
tive group. An example of a noncommutative group that we have seen is the group $LF(\mathbb{Z})$ of linear fractional transformations, where the multiplication comes from mul-
tiplication of $2 \times 2$ matrices, or equivalently, composition of the transformations.

For a given discriminant, if the numbers represented by two primitive forms can-
not be distinguished by congruences mod the discriminant, then these two forms are
said to belong to the same genus. Thus in the preceding example of discriminant $-56$
the two forms $Q_1$ and $Q_2$ are of the same genus while $Q_3$ is of a different genus from
$Q_1$ and $Q_2$, so there are two different genera (“genera” is the plural of “genus”).

Equivalent forms always belong to the same genus since their topographs contain
exactly the same numbers. The first two of the three levels of complexity we have
described correspond to the discriminants where there is only one equivalence class
in each genus. As we stated earlier, this desirable situation is also characterized by
the condition that all primitive forms of the given discriminant have mirror symmetry.
For larger discriminants there can be large numbers of genera and large numbers of
equivalence classes within a genus. However, for a fixed discriminant there are always
the same number of proper equivalence classes within each genus, as we will show in
Corollary 7.18. This is illustrated by the case $\Delta = -56$ where one genus consists of
$Q_1$ and $Q_2$ and the other genus consists of $Q_3$ and $Q_4$.

We just saw an example where two non-equivalent forms of the same discriminant
can both represent the same number. However, this does not happen for representa-
tions of 1 or primes or powers of most primes:

**Proposition 6.2.** If $Q_1$ and $Q_2$ are two forms of the same discriminant that both
represent the same prime $p$ or both represent 1, then $Q_1$ and $Q_2$ are equivalent.
The same conclusion holds when $Q_1$ and $Q_2$ both represent the same power $p^k$ of
an odd prime $p$ that does not divide the discriminant.

The last statement is also true for $p = 2$ but the proof is more difficult so we will
wait until the next chapter to deduce this from a more general result, Theorem 7.8.

For an example showing that the second part of the proposition can fail for primes
dividing the discriminant we can take the forms $Q_1 = x^2 + xy + 25y^2$ and $Q_2 =
5x^2 + xy + 5y^2$ of discriminant $-99$, which both represent $3^3$ since $Q_1(1,1) =
27 = Q_2(1,2)$. They also both represent $3^4 = 81$ as $Q_1(7,1)$ and $Q_2(1,−4)$. These
two forms are not equivalent since they are distinct reduced elliptic forms. In this example the prime $p$ in fact divides the conductor which is 3 since $-99 = 3^2(-11)$ and $-11$ is a fundamental discriminant. Primes that divide the discriminant but not the conductor are actually never exceptional for this proposition since for such primes $p$ the only power $p^k$ that is represented by a form of the given discriminant is $p$ itself, by Proposition 6.1. We saw this in all the earlier examples, where the discriminants were fundamental so the conductors were 1.

*Proof:* Suppose that $Q$ is a form representing a number $p$ that is either 1 or a prime. The topograph of $Q$ then has a region labeled $p$, and we have seen that the $h$-labels on the edges adjacent to this $p$-region form an arithmetic progression with increment $2p$ when these edges are all oriented in the same direction. We have the discriminant formula $\Delta = h^2 - 4pq$ where $h$ is the label on one of these edges and $q$ is the value of $Q$ for the region on the other side of this edge. Since $p$ is nonzero the equation $\Delta = h^2 - 4pq$ determines $q$ in terms of $\Delta$ and $h$. This implies that $\Delta$ and the arithmetic progression determine the form $Q$ up to equivalence since the progression determines $p$, and any $h$-value in the progression then determines the $q$-value corresponding to this $h$-value, so $Q$ is equivalent to $px^2 + hxy + qy^2$.

In the case that $p = 1$ the increment in the arithmetic progressions is $2p = 2$ so the two possible progressions of $h$-values adjacent to the $p$-region are the even numbers and the odd numbers. We know that $h$ has the same parity as $\Delta$, so $\Delta$ determines which of the two progressions we have. As we saw in the preceding paragraph, this implies that the form is determined by $\Delta$, up to equivalence.

Now we consider the case that $p$ is prime. Let $Q_1$ and $Q_2$ be two forms of the same discriminant $\Delta$ both representing $p$. For $Q_1$ choose an edge in its topograph adjacent to the $p$-region, with $h$-label $h_1$ and $q$-label $q_1$. For the form $Q_2$ we similarly choose an edge with associated labels $h_2$ and $q_2$. Both $h_1$ and $h_2$ have the same parity as $\Delta$. We have $\Delta = h_1^2 - 4pq_1 = h_2^2 - 4pq_2$ and hence $h_1^2 = h_2^2 \mod 4p$. This implies $h_1^2 \equiv h_2^2 \mod p$, so $p$ divides $h_1^2 - h_2^2 = (h_1 + h_2)(h_1 - h_2)$. Since $p$ is prime, it must divide one of the two factors and hence we must have $h_1 \equiv \pm h_2 \mod p$. By changing the orientations of the edges in the topograph for $Q_1$ or $Q_2$ if necessary, we can assume that $h_1 \equiv h_2 \mod p$.

If $p$ is odd we can improve this congruence to $h_1 \equiv h_2 \mod 2p$ since we know that $h_1 - h_2$ is divisible by both $p$ and 2 (since $h_1$ and $h_2$ have the same parity), hence $h_1 - h_2$ is divisible by $2p$. The congruence $h_1 \equiv h_2 \mod 2p$ implies that the arithmetic progression of $h$-values adjacent to the $p$-region for $Q_1$ is the same as for $Q_2$ since $2p$ is the increment for both progressions. By what we showed earlier, this implies that $Q_1$ and $Q_2$ are equivalent.

When $p = 2$ this argument needs to be modified slightly. We still have $h_1^2 \equiv h_2^2 \mod 4p$ so when $p = 2$ this becomes $h_1^2 \equiv h_2^2 \mod 8$. Since $2p = 4$ the four possible arithmetic progressions of $h$-values are $h \equiv 0, 1, 2, \text{ or } 3 \mod 4$. We can
interchange the possibilities 1 and 3 just by reorienting the edges, leaving only the possibilities \( h \equiv 0, 1, \text{ or } 2 \mod 4 \). Since \( h_1 \) and \( h_2 \) have the same parity, this takes care of the case that \( h_1 \) and \( h_2 \) are odd. The remaining two cases \( h \equiv 0, 2 \mod 4 \) are distinguished from each other by the congruence \( h_1^2 \equiv h_2^2 \mod 8 \) since \( (4k)^2 \equiv 0 \mod 8 \) and \( (4k + 2)^2 \equiv 4 \mod 8 \).

Finally we have the case that \( Q_1 \) and \( Q_2 \) both represent the power \( p^k \) of an odd prime \( p \) not dividing \( \Delta \). Following the line of proof above we see that \( p^k \) divides \( h_1^2 - h_2^2 = (h_1 + h_2)(h_1 - h_2) \). If \( p^k \) divides either factor we can proceed exactly as before to show that \( Q_1 \) and \( Q_2 \) are equivalent since we assume \( p \) is odd, hence also \( p^k \). If \( p^k \) does not divide either factor then both factors are divisible by \( p \) (we can assume \( k > 1 \)), hence \( p \) divides their sum \( 2h_1 \). Since \( p \) is odd this implies that \( p \) divides \( h_1 \), and so \( p \) divides \( \Delta = h_1^2 - 4p^kq_1 \). Thus if \( p \) does not divide \( \Delta \) then the case that \( p^k \) divides neither \( h_1 + h_2 \) nor \( h_1 - h_2 \) does not arise. \( \square \)

The same argument shows another interesting fact:

**Proposition 6.3.** If the topograph of a form has two regions with the same label \( n \) where \( n \) is either 1, a prime, or a power of an odd prime not dividing the discriminant, then there is a symmetry of the topograph that takes one region labeled \( n \) to the other. Similarly, for positive discriminants and for the same numbers \( n \), if there is one region labeled \( n \) and another labeled \( -n \) then there is a skew symmetry taking one region to the other.

**Proof:** Suppose first that there are two regions having the same label \( n \). As we saw in the proof of the preceding proposition, each of these regions is adjacent to an edge with the same label \( h \) and hence the labels \( q \) across these edges are also the same. This means there is a symmetry taking one region labeled \( n \) to the other. The other case is that one region is labeled \( n \) and the other \( -n \). This means that the given form \( Q \) and its negative \( -Q \) each have a region labeled \( n \) so there is an equivalence from \( Q \) to \( -Q \) taking the \( n \) region for \( Q \) to the \( n \) region for \( -Q \). This equivalence can be regarded as a skew symmetry of \( Q \) taking the \( n \) region to the \( -n \) region. \( \square \)

The examples in this section show the significance of primes in certain congruence classes for solving the representation problem. In the examples there seems to be no shortage of primes in each of the relevant congruence classes. For example, for the form \( x^2 + y^2 \) the primes represented, apart from 2, seem to be the primes congruent to 1 mod 4, the primes of the form \( 4k + 1 \) starting with 5, 13, 17, 29, 37, 41, 53, \( \ldots \). The other possibility for odd primes is the sequence 3, 7, 11, 19, 23, 31, 43, 47, \( \ldots \), primes of the form \( 4k - 1 \).

Such sequences form arithmetic progressions \( an + b \) for fixed positive integers \( a \) and \( b \) and varying \( n \) = 0, 1, 2, 3, \( \ldots \). It is natural to ask whether there are infinitely many primes in each arithmetic progression \( an + b \). For this to be true an obvious
restriction is that \( a \) and \( b \) should be coprime since any common divisor of \( a \) and \( b \) will divide every number \( an + b \), so there could be at most one prime in the progression.

A famous theorem of Dirichlet from 1837 asserts that every arithmetic progression \( an + b \) with \( a \) and \( b \) coprime contains an infinite number of primes. This can be rephrased as saying that within each congruence class of numbers \( x \equiv b \) mod \( a \) there are infinitely many primes whenever \( a \) and \( b \) are coprime. Dirichlet’s theorem actually says more, that primes are approximately equally distributed among the various congruence classes mod \( a \) for a fixed \( a \). For example, there are approximately as many primes \( p = 4n + 1 \) as there are primes \( p = 4n - 1 \).

Dirichlet’s Theorem is not easy to prove, and a proof would require methods quite different for anything else in this book so we will not be giving a proof. However a few special cases of Dirichlet’s Theorem can be proved by elementary arguments. The simplest case is the arithmetic progression \( 3, 7, 11, \cdots \) of numbers \( n = 4n - 1 \), using a variant of Euclid’s proof that there are infinitely many primes. Recall how Euclid’s argument goes. Suppose that \( p_1, \cdots, p_k \) is a finite list of primes, and consider the number \( N = p_1 \cdots p_k + 1 \). This must be divisible by some prime \( p \), but \( p \) cannot be any of the primes \( p_i \) since dividing \( p_i \) into \( N \) gives a remainder of 1. Thus no finite list of primes can be complete and hence there must be infinitely many primes.

To adapt this argument to primes of the form \( 4n - 1 \), suppose that \( p_1, \cdots, p_k \) is a finite list of such primes, and consider the number \( N = 4p_1 \cdots p_k - 1 \). The prime divisors of \( N \) must be odd since \( N \) is odd. If all these prime divisors were of the form \( 4n + 1 \) then \( N \) would be a product of numbers of the form \( 4n + 1 \) hence \( N \) itself would have this form, contradicting the fact that \( N \) has the form \( 4n - 1 \). Hence \( N \) must have a prime factor \( p = 4n - 1 \). This \( p \) cannot be any of the primes \( p_i \) since dividing \( p_i \) into \( N \) gives a remainder of \( -1 \). Thus no finite list of primes \( 4n - 1 \) can be a complete list.

This argument does not work for primes \( p = 4n + 1 \) since a number \( N = 4p_1 \cdots p_k + 1 \) can be a product of primes of the form \( 4n - 1 \) so one could not deduce that \( N \) had a prime factor \( p = 4n + 1 \).

However, the quadratic form \( x^2 + y^2 \) can be used to show there are infinitely many primes \( p = 4n + 1 \). In Corollary 6.7 we show that for each discriminant \( \Delta \) there are infinitely many primes represented by at least one form of discriminant \( \Delta \). (This form will be unique up to equivalence by Proposition 6.2.) In the case \( \Delta = -4 \) all forms are equivalent to the form \( x^2 + y^2 \), so this form must represent infinitely many primes. None of these primes can be of the form \( 4n - 1 \) since all values of \( x^2 + y^2 \) are congruent to 0, 1, or 2 mod 4, as squares are always 0 or 1 mod 4. Thus there must be infinitely many primes \( p = 4n + 1 \).

The same arguments work also for primes \( p = 3n + 1 \) and \( p = 3n - 1 \). For \( p = 3n - 1 \) one argues just as for \( 4n - 1 \), using numbers \( N = 3p_1 \cdots p_k - 1 \). For
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$p = 3n + 1$ one uses the form $x^2 + xy + y^2$ of discriminant $-3$. Here again all forms of this discriminant are equivalent so Corollary 6.7 says that $x^2 + xy + y^2$ represents infinitely many primes. All values of $x^2 + xy + y^2$ are congruent to 0 or 1 mod 3 as one can easily check by listing the various possibilities for $x$ and $y$ mod 3. Thus there are infinitely many primes $p = 3n + 1$.

We can try these arguments for arithmetic progressions $5n \pm 1$ and $5n \pm 2$ but there are problems. The Euclidean argument we have given fails in each case for much the same reason that it failed for primes $p = 4n + 1$. For the approach via quadratic forms we would use the form $x^2 + xy - y^2$ of discriminant 5. This is the only form of this discriminant, up to equivalence, so Corollary 6.7 implies that it represents infinitely many primes. The methods in the next section will show that the primes represented by this form are the primes $p = 5n \pm 1$, so there are infinitely many primes $p = 5n + 1$ or $p = 5n - 1$ but we cannot be more specific than this. Dirichlet's Theorem says there are infinitely primes of each type, and in fact there are fancier forms of the Euclidean argument that prove this, but these Euclidean arguments do not work for the other cases $p = 5n \pm 2$.

We have just seen three quadratic forms that represent infinitely many primes, for discriminants $-4$, $-3$, and 5, and Corollary 6.7 provides other examples for each discriminant with class number 1. (Nonprimitive forms obviously cannot represent infinitely many primes, so these forms can be ignored.) For discriminants with larger class numbers Corollary 6.7 only implies that there is at least one form representing infinitely many primes. However there is another hard theorem of Dirichlet which does say that each primitive form of nonsquare discriminant represents infinitely many primes.

2. Representing Primes

Based on the examples in the previous section, a reasonable strategy for attacking the problem of determining which numbers are represented by the various forms of a given discriminant would be first to determine which primes are represented, then figure out which products of these primes are represented. In this section we focus on the first step of this program, representing primes.

One special case is easy:

**Proposition 6.4.** For each prime $p$ that divides the discriminant $\Delta$ there exists a form of discriminant $\Delta$ that represents $p$.

**Proof:** Consider first the case $p = 2$, so we are looking at forms with even discriminant. A form representing 2 must be equivalent to a form $2x^2 + bxy + cy^2$. The simplest possibility would be a form $2x^2 + cy^2$, which for convenience we write as $2x^2 - ky^2$. This has discriminant $8k$, so we are finished if $\Delta = 8k$ for some $k$. Since $\Delta$ is even it must be a multiple of 4 so the only remaining possibility is $\Delta = 8k + 4$ for some $k$. In this case the form $2x^2 + 2xy - ky^2$ works.
This suggests the possibility that perhaps for an odd prime \( p \) dividing \( \Delta \) one of the forms \( px^2 - ky^2 \) or \( px^2 + pxy - ky^2 \) will work. Certainly these forms represent \( p \), so the only question is whether \( k \) can be chosen so that one of the two forms has the given discriminant \( \Delta \). For \( px^2 - ky^2 \) the discriminant is \( \Delta = 4pk \) so in this case \( \Delta \equiv 0 \mod 4 \). Solving \( \Delta = 4pk \) for \( k \), we have \( k = \Delta/4p \). This is an integer when \( \Delta \equiv 0 \mod 4 \) since this says 4 divides \( \Delta \), and \( p \) divides \( \Delta \) by hypothesis, so \( 4p \) divides \( \Delta \) since 4 and \( p \) are coprime when \( p \) is odd.

There remains the case \( \Delta \equiv 1 \mod 4 \). For the form \( px^2 + pxy - ky^2 \) we have \( \Delta = p^2 + 4pk \) so \( k = (\Delta - p^2)/4p \). If \( p \) divides \( \Delta \) then \( p \) divides the numerator \( \Delta - p^2 \). Also 4 divides the numerator since \( p^2 \equiv 1 \mod 4 \) when \( p \) is odd, and \( \Delta \equiv 1 \mod 4 \) by assumption. Thus \( 4p \) divides the numerator so \( k \) is an integer and we have a form \( px^2 + pxy - ky^2 \) representing \( p \).

Before beginning the analysis for primes not dividing the discriminant we first reformulate the problem of representability in a fixed discriminant as a congruence condition. Suppose a number \( n \) is represented by some form \( Q(x,y) \) of discriminant \( \Delta \), so \( n \) appears in the topograph of \( Q \). If we look at an edge of the topograph bordering a region labeled \( n \) then we obtain an equation \( \Delta = h^2 - 4nk \) where \( h \) is the label on the edge and \( k \) is the label on the region on the opposite side of this edge. The key observation is then that the equation \( \Delta = h^2 - 4nk \) says precisely that \( \Delta \) is congruent to \( h^2 \) modulo \( 4n \). Notice that if \( n \) is negative, “modulo \( 4n \)” means the same thing as “modulo \( 4|n| \)” since being divisible by a number \( d \) is equivalent to being divisible by \(-d\) when we are considering both positive and negative numbers.

This in fact gives an exact criterion for representability in a given discriminant:

**Proposition 6.5.** For a fixed discriminant \( \Delta \) there exists a form of discriminant \( \Delta \) that represents \( n \) if and only if \( \Delta \) is congruent to a square modulo \( 4n \).

**Proof:** As we saw above, if we have a form of discriminant \( \Delta \) representing \( n \) then by looking at the topograph we get an equation \( \Delta = h^2 - 4nk \) for some integers \( h \) and \( k \), and this equation says that \( \Delta \) is the square of \( h \) modulo \( 4n \). Conversely, suppose that \( \Delta \) is the square of some integer \( h \) modulo \( 4n \). This means that \( h^2 - \Delta \) is an integer times \( 4n \), or in other words \( h^2 - \Delta = 4nk \) for some \( k \). This equation can also be written as \( \Delta = h^2 - 4nk \). The three numbers \( n \), \( h \), and \( k \) can be used to construct a form whose topograph contains an edge with these three labels, for example \( nx^2 + hxy + ky^2 \) which has these three labels at the \( 1/0, 0/1 \) edge. The discriminant of this form has the desired value \( \Delta = h^2 - 4nk \), and the form represents \( n \) since \( n \) appears as the label on a region in the topograph.

**Corollary 6.6.** If a number \( n \) is represented in discriminant \( \Delta \) then so is every divisor of \( n \).
Thus if we find which primes are represented in discriminant \( \Delta \) and then which products of these primes are represented, we will have found all numbers that are represented.

**Proof:** This is based on a simple observation that we will use repeatedly:

If a congruence \( a \equiv b \mod n \) then it holds mod \( d \) for each divisor \( d \) of \( n \).

This is true because if \( n \) divides \( a - b \) then so does \( d \) for each divisor \( d \) of \( n \).

Thus if \( \Delta \equiv h^2 \mod 4n \), for some number \( h \), then \( \Delta \equiv h^2 \mod 4d \) whenever \( d \) divides \( n \) since \( 4d \) then divides \( 4n \).

\[ \square \]

**Corollary 6.7.** For each discriminant \( \Delta \) the set of primes represented in discriminant \( \Delta \) is infinite.

**Proof:** Let \( Q(x,y) = x^2 + bxy + cy^2 \) be the principal form of discriminant \( \Delta \), and let \( p_1, \ldots, p_k \) be any finite set of primes, possibly with repetitions. The number \( n = Q((1,1,\ldots,1)) = 1 + bp_1 \cdots p_k + c(p_1 \cdots p_k)^2 \) is represented by \( Q \) since \((1,1,\ldots,1)\) is a primitive pair. If \( k \) is large enough we will have \( |n| > 1 \) since \( |c(p_1 \cdots p_k)^2| \) will be much larger than \( |1 + bp_1 \cdots p_k| \). Let \( p \) be any prime divisor of \( n \). By the preceding corollary \( p \) is represented by some form of discriminant \( \Delta \). Also, \( p \) is different from any of the primes \( p_i \) since dividing \( p_i \) into \( n = 1 + bp_1 \cdots p_k + c(p_1 \cdots p_k)^2 \) gives a remainder of 1, whereas \( p \) divides \( n \) evenly. Thus we have shown that for any finite list of primes there is another prime not on the list that is represented in discriminant \( \Delta \), so the set of primes represented in discriminant \( \Delta \) must be infinite.

\[ \square \]

Let us see what the preceding proposition implies about representing small numbers \( n \). For \( n = 1 \) it says that there is a form of discriminant \( \Delta \) representing 1 if and only if \( \Delta \) is a square modulo 4. The squares modulo 4 are 0 and 1, and we already know that discriminants of forms are always congruent to 0 or 1 modulo 4. So we conclude that for every possible value of the discriminant there exists a form that represents 1. This isn’t really new information, however, since the principal form \( x^2 + dy^2 \) or \( x^2 + xy + dy^2 \) represents 1 and there is a principal form for each discriminant.

In the next case \( n = 2 \) the possible values of the discriminant modulo \( 4n = 8 \) are 0,1,4,5, and the squares modulo 8 are 0,1,4 since \( 0^2 = 0 \), \( (\pm 1)^2 = 1 \), \( (\pm 2)^2 = 4 \), \( (\pm 3)^2 \equiv 1 \), and \( (\pm 4)^2 \equiv 0 \). Thus 2 is not represented by any form of discriminant \( \Delta \equiv 5 \mod 8 \), but for all other values of the discriminant there is a form representing 2. We already saw explicit forms doing this for discriminants \( \Delta = 8k \) and \( 8k + 4 \) in Proposition 6.4, the forms \( 2x^2 - ky^2 \) and \( 2x^2 + 2xy - ky^2 \), and for \( \Delta = 8k + 1 \) the form \( 2x^2 + xy - ky^2 \) works.

Moving on to the next case \( n = 3 \), the discriminants mod 12 are 0,1,4,5,8,9 and the squares modulo 12 are 0,1,4,9 since \( 0^2 = 0 \), \( (\pm 1)^2 = 1 \), \( (\pm 2)^2 = 4 \), \( (\pm 3)^2 = 9 \), and for \( \Delta = 8k + 3 \) the form \( 2x^2 - xy - ky^2 \) works.
9, \((\pm 4)^2 \equiv 4\), \((\pm 5)^2 \equiv 1\), and \((\pm 6)^2 \equiv 0\). The excluded discriminants are thus those congruent to 5 or 8 mod 12. Again explicit forms are easily given, the forms \(3x^2 + hxy - ky^2\) with \(\Delta = 12k + h^2\) for \(h = 0, 1, 2, 3\).

We could continue in this direction, exploring which discriminants have forms that represent a given number, but this is not really the question we want to answer, which is to start with a given discriminant, or even a given form, and decide which numbers it represents. The sort of answer we are looking for, based on the various examples we looked at earlier, is also a different sort of congruence condition, with congruence modulo the discriminant rather than congruence modulo \(4n\). So there is more work to be done before we would have the sort of answer we want. Nevertheless, the representability criterion in the preceding proposition is the starting point.

For analyzing the question of which odd primes \(p\) not dividing the discriminant \(\Delta\) are represented by a form of discriminant \(\Delta\) the following result will be useful.

**Lemma 6.8.** When \(n\) is odd, the condition that a discriminant \(\Delta\) is congruent to a square mod \(4n\) is equivalent to \(\Delta\) being a square mod \(n\).

**Proof:** If \(\Delta \equiv h^2 \mod 4n\) then certainly \(\Delta \equiv h^2 \mod n\), whether \(n\) is odd or even. For the less obvious converse, suppose that \(\Delta \equiv h^2 \mod n\). We can assume that \(h\) has the same parity as \(\Delta\) since if it doesn’t, we can replace \(h\) by \(h + n\) which has the opposite parity from \(h\) since \(n\) is odd, and then note that \((h + n)^2 \equiv h^2 \mod n\). Since we always have \(\Delta \equiv 0\) or 1 mod 4, we must have \(\Delta = 4k\) or \(\Delta = 4k + 1\), and since \(h\) has the same parity as \(\Delta\) we have \(h^2 = 4l\) or \(h^2 = 4l + 1\) in these two cases. In either case \(\Delta - h^2 = 4(k - l)\) so \(\Delta - h^2\) is divisible by 4. It is also divisible by \(n\) by the assumption that \(\Delta \equiv h^2 \mod n\). Since \(n\) is odd, this implies that \(\Delta - h^2\) is divisible by \(4n\), so \(\Delta \equiv h^2 \mod 4n\), which finishes the proof of the converse. \(\square\)

At this point it will be convenient to introduce some shorthand notation. For \(p\) an odd prime and \(a\) an integer not divisible by \(p\), define the Legendre symbol \((\frac{a}{p})\) by

\[
(\frac{a}{p}) = \begin{cases} 
+1 & \text{if } a \text{ is a square mod } p \\
-1 & \text{if } a \text{ is not a square mod } p 
\end{cases}
\]

Using this notation we can say:

An odd prime \(p\) that does not divide \(\Delta\) is representable by some form of discriminant \(\Delta\) if and only if \((\frac{\Delta}{p}) = 1\).

It will therefore be useful to know how to compute \((\frac{a}{p})\). The following four basic properties of the Legendre symbol will make this a feasible task:

1. \((\frac{ab}{p}) = (\frac{a}{p})(\frac{b}{p})\).
2. \((\frac{-1}{p}) = +1\) if \(p \equiv 1 \mod 4\) and \((\frac{-1}{p}) = -1\) if \(p \equiv 3 \mod 4\).
3. \((\frac{2}{p}) = +1\) if \(p \equiv \pm 1 \mod 8\) and \((\frac{2}{p}) = -1\) if \(p \equiv \pm 3 \mod 8\).
(4) If \( p \) and \( q \) are distinct odd primes then \( \left( \frac{p}{q} \right) = \left( \frac{q}{p} \right) \) unless \( p \) and \( q \) are both congruent to 3 mod 4, in which case \( \left( \frac{p}{q} \right) = -\left( \frac{q}{p} \right) \).

Property (1), applied repeatedly, reduces the calculation of \( \left( \frac{a}{p} \right) \) to the calculation of \( \left( \frac{q}{p} \right) \) for the various prime factors \( q \) of \( a \), along with \( \left( \frac{-1}{p} \right) \) when \( a \) is negative. Note that \( \left( \frac{q^2}{p} \right) = +1 \) so we can immediately reduce to the case that \( |a| \) is a product of distinct primes. Property (2) will be useful in dealing with negative discriminants, and property (3) will be used for certain even discriminants.

Property (4), which is by far the most subtle of the four properties, is called Quadratic Reciprocity. Its proof is considerably more difficult than for the other three properties and will be given in the last section of this chapter. Proofs of the first three properties will be obtained along the way.

For a quick illustration of the usefulness of these properties let us see how they can be used to compute the values of Legendre symbols. Suppose for example that one wanted to know whether 78 was a square mod 89. The naive approach would be to list the squares of all the numbers \( \pm 1, \pm 2, \ldots, \pm 44 \) and see whether any of these was congruent to 78 mod 89, but this would be rather tedious. Since 89 is prime we can instead evaluate \( \left( \frac{78}{89} \right) \) using the basic properties of Legendre symbols. First we factor 78 to get \( \left( \frac{78}{89} \right) = \left( \frac{2}{89} \right) \left( \frac{3}{89} \right) \left( \frac{13}{89} \right) \). By property (3) we have \( \left( \frac{2}{89} \right) = +1 \) since 89 \( \equiv 1 \mod 8 \). Next we apply reciprocity to get \( \left( \frac{3}{89} \right) = \left( \frac{89}{3} \right) \) and \( \left( \frac{13}{89} \right) = \left( \frac{89}{13} \right) \) since 89 \( \equiv 1 \mod 4 \). Next use the fact that \( \left( \frac{a}{p} \right) \) depends only on the value of \( a \mod p \) to reduce \( \left( \frac{89}{3} \right) \) to \( \left( \frac{2}{3} \right) \) and \( \left( \frac{89}{13} \right) \) to \( \left( \frac{11}{13} \right) \). Using property (3) again we have \( \left( \frac{2}{3} \right) = -1 \) (confirming the obvious fact that 2 is not a square mod 3). For \( \left( \frac{11}{13} \right) \) reciprocity says this equals \( \left( \frac{13}{11} \right) \). This reduces to \( \left( \frac{2}{11} \right) = -1 \). Summarizing, we have \( \left( \frac{78}{89} \right) = \left( \frac{2}{89} \right) \left( \frac{3}{89} \right) \left( \frac{13}{89} \right) = (+1)(-1)(-1) = +1 \) so 78 is a square mod 89. However, this method does not actually produce a number \( x \) such that \( x^2 \equiv 78 \mod 89 \).

Later in the chapter we will also see how to determine whether a number is a square mod a nonprime, showing how this reduces to the prime case.

Returning now to quadratic forms, let us see what the basic properties of Legendre symbols tell us about which primes are represented by the forms discussed at the beginning of the chapter. In the first four cases the class number is 1 so we will be determining which primes are represented by the given form.

**Example:** \( x^2 + y^2 \) with \( \Delta = -4 \). This form obviously represents 2, and it represents an odd prime \( p \) exactly when \( \left( \frac{-4}{p} \right) = +1 \). Using the first of the four properties we have \( \left( \frac{-4}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right) \left( \frac{p}{p} \right) = \left( \frac{-1}{p} \right) \), and the second property says this is +1 exactly for primes \( p = 4k + 1 \). Thus we see the primes represented by \( x^2 + y^2 \) are 2 and the primes \( p = 4k + 1 \).
Example: \( x^2 + 2y^2 \) with \( \Delta = -8 \). The only prime dividing \( \Delta \) is 2, and it is representable. For odd primes \( p \) we have \( \left( \frac{-8}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right)^3 = \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right). \) In the four cases \( p \equiv 1, 3, 5, 7 \mod 8 \) this is, respectively, \((+1)(+1), (-1)(-1), (+1)(-1), \) and \((-1)(+1).\) We conclude that the primes representable by the form \( x^2 + 2y^2 \) are 2 and primes congruent to 1 or 3 mod 8.

Example: \( x^2 - 2y^2 \) with \( \Delta = 8 \). We have \( \left( \frac{8}{p} \right) = \left( \frac{2}{p} \right)^3 = \left( \frac{2}{p} \right) \) so from property (3) we conclude that the primes represented by \( x^2 - 2y^2 \) are 2 and \( p \equiv \pm 1 \mod 8. \)

Example: \( x^2 + xy + y^2 \) with \( \Delta = -3. \) The only prime dividing the discriminant is 3 and it is represented. The prime 2 is not represented since \( \Delta \equiv 5 \mod 8. \) For primes \( p > 3 \) we can evaluate \( \left( \frac{-3}{p} \right) \) using quadratic reciprocity, which says that \( \left( \frac{-3}{p} \right) = \left( \frac{3}{p} \right) \) if \( p = 4k + 1 \) and \( \left( \frac{-3}{p} \right) = -\left( \frac{3}{p} \right) \) if \( p = 4k + 3. \) Thus when \( p = 4k + 1 \) we have \( \left( \frac{-3}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{3}{p} \right) = \left( \frac{3}{p} \right) \) and when \( p = 4k + 3 \) we have \( \left( \frac{-3}{p} \right) = (-1) \left( -\left( \frac{3}{p} \right) \right) \) so we get \( \left( \frac{-3}{p} \right) \) in both cases. Since \( \left( \frac{3}{p} \right) \) only depends on \( \mod 3 \), we get \( \left( \frac{-3}{p} \right) = +1 \) if \( p \equiv 1 \mod 3 \) and \( \left( \frac{-3}{p} \right) = -1 \) if \( p \equiv 2 \mod 3. \) (Since \( p \neq 3 \) we do not need to consider the possibility \( p \equiv 0 \mod 3. \)) The conclusion of all this is that the primes represented by \( x^2 + xy + y^2 \) are 3 and the primes \( p \equiv 1 \mod 3. \)

Example: \( \Delta = 40. \) Here all forms are equivalent to either \( x^2 - 10y^2 \) or \( 2x^2 - 5y^2. \) The primes dividing 40 are 2 and 5 so these are represented by one form or the other, and in fact both are represented by \( 2x^2 - 5y^2 \) as the topographs showed. For other primes \( p \) we have \( \left( \frac{40}{p} \right) = \left( \frac{2}{p} \right)^3 \left( \frac{5}{p} \right) = \left( \frac{2}{p} \right) \left( \frac{5}{p} \right). \) The factor \( \left( \frac{2}{p} \right) \) depends only on \( p \mod 8 \) and \( \left( \frac{5}{p} \right) \) depends only on \( p \mod 5, \) so their product depends only on \( p \mod 40. \) The following table lists all the possibilities for congruence classes mod 40 not divisible by 2 or 5:

<table>
<thead>
<tr>
<th>( \frac{2}{p} )</th>
<th>1 3 7 9 11 13 17 19 21 23 27 29 31 33 37 39</th>
</tr>
</thead>
<tbody>
<tr>
<td>( \frac{5}{p} )</td>
<td>+1 -1 -1 +1 -1 -1 -1 +1 -1 -1 +1 +1 +1 -1 +1</td>
</tr>
</tbody>
</table>

The product \( \left( \frac{2}{p} \right) \left( \frac{5}{p} \right) \) is +1 in exactly the eight cases \( p \equiv 1, 3, 9, 13, 27, 31, 37, 39 \mod 40. \) We conclude that these are the eight congruence classes containing primes (other than 2 and 5) represented by one of the two forms \( x^2 - 10y^2 \) and \( 2x^2 - 5y^2. \) This agrees with our earlier observations based on the topographs. However, we have yet to verify our earlier guesses as to which congruence classes are represented by which form. We will see how to do this shortly.

In the examples above we were able to express \( \left( \frac{\Delta}{p} \right) \) in terms of Legendre symbols \( \left( \frac{-1}{p} \right), \left( \frac{2}{p} \right), \) and \( \left( \frac{5}{p} \right) \) for odd primes \( p \) dividing \( \Delta. \) The following result shows that this can be done for all \( \Delta. \)

Proposition 6.9. Let the nonzero integer \( \Delta \) be factored as \( \Delta = \epsilon 2^s p_1 \ldots p_k \) for \( \epsilon = \pm 1, s \geq 0, \) and each \( p_i \) an odd prime. (We allow \( k = 0 \) when \( \Delta = \epsilon 2^s. \)) Then for odd
primes \( p \) not dividing \( \Delta \) the Legendre symbol \(( \frac{\Delta}{p} )\) has the value given in the table below.

<table>
<thead>
<tr>
<th>( \Delta )</th>
<th>(( \frac{\Delta}{p} ))</th>
</tr>
</thead>
<tbody>
<tr>
<td>( 2^{2l}(4m + 1) )</td>
<td>( \left( \frac{p}{p_1} \right) \cdots \left( \frac{p}{p_k} \right) )</td>
</tr>
<tr>
<td>( 2^{2l}(4m + 3) )</td>
<td>( \left( \frac{-1}{p} \right) \left( \frac{p}{p_1} \right) \cdots \left( \frac{p}{p_k} \right) )</td>
</tr>
<tr>
<td>( 2^{2l+1}(4m + 1) )</td>
<td>( \left( \frac{2}{p} \right) \left( \frac{p}{p_1} \right) \cdots \left( \frac{p}{p_k} \right) )</td>
</tr>
<tr>
<td>( 2^{2l+1}(4m + 3) )</td>
<td>( \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right) \left( \frac{p}{p_1} \right) \cdots \left( \frac{p}{p_k} \right) )</td>
</tr>
</tbody>
</table>

**Proof:** We have \(( \frac{\Delta}{p} ) = \left( \frac{\xi}{p} \right) \left( \frac{\omega}{p} \right)^s \left( \frac{p_1}{p} \right) \cdots \left( \frac{p_k}{p} \right)\). Quadratic reciprocity implies that \( \left( \frac{p_1}{p} \right) \cdots \left( \frac{p_k}{p} \right) = \left( \frac{\omega}{p} \right)^{\varepsilon} \left( \frac{p}{p} \right) \cdots \left( \frac{p}{p} \right) \) where \( \omega \) is +1 or −1 according to whether there are an even or an odd number of factors \( p_i \equiv 3 \mod 4 \). Thus we have \(( \frac{\Delta}{p} ) = \left( \frac{\xi}{p} \right) \left( \frac{\omega}{p} \right)^s \left( \frac{p_1}{p} \right) \cdots \left( \frac{p_k}{p} \right) \). The exponent \( s \) in this formula can be replaced by 0 or 1 according to whether \( s \) is even or odd. In the first and third rows of the table the odd part of \( \Delta \) is \( 4m + 1 \) so we have \( \varepsilon = \omega \) hence \(( \frac{\xi}{p} \left( \frac{\omega}{p} \right) = 1 \). In the second and fourth rows the factor \( 4m + 1 \) is replaced by \( 4m + 3 \) and we have \( \varepsilon = -\omega \), hence \(( \frac{\xi}{p} \left( \frac{-1}{p} \right) \) .

To get more information on how Legendre symbols can be used to tell which forms represent which primes we will use the following key result:

**Proposition 6.10.** Let \( Q \) be a form of discriminant \( \Delta \) and let \( p \) be an odd prime dividing \( \Delta \). Then the Legendre symbol \(( \frac{n}{p} )\) has the same value for all numbers \( n \) in the topograph of \( Q \) that are not divisible by \( p \).

Before proving this let us see how it applies in the case \( \Delta = 40 \) with \( p = 5 \). According to the proposition, the value of \(( \frac{\Delta}{p} )\) must be constant in the topograph of each of the forms \( x^2 - 10y^2 \) and \( 2x^2 - 5y^2 \), leaving aside values of the forms that are divisible by 5. To determine the value of \(( \frac{n}{5} )\) for each form it therefore suffices to compute it for a single number \( n \) in the topograph not divisible by 5. The simplest thing is just to compute it for \((x, y) = (1, 0) \) or \((0, 1) \). Choosing \((1, 0) \), for \( x^2 - 10y^2 \) we have \(( \frac{1}{5} ) = +1 \) and for \( 2x^2 - 5y^2 \) we have \(( \frac{2}{5} ) = -1 \). The proposition then implies that all values \( n \) in the topograph of \( x^2 - 10y^2 \) not divisible by 5 have \(( \frac{n}{5} ) = +1 \), hence \( n \equiv \pm 1 \mod 5 \), while for \( 2x^2 - 5y^2 \) we have \(( \frac{n}{5} ) = -1 \), hence \( n \equiv \pm 2 \mod 5 \). We have seen that the product \(( \frac{\xi}{p} \left( \frac{\omega}{p} \right) \) must equal +1 for primes \( p \equiv 2, 5 \) represented by either form, so for \( x^2 - 10y^2 \) this product must be \((+1)(+1) \) while for \( 2x^2 - 5y^2 \) it must be \((-1)(-1) \). Thus from the earlier table of values for \(( \frac{\xi}{p} \) and \(( \frac{\omega}{p} \) we are able to see exactly which primes each of these two forms represents, namely \( x^2 - 10y^2 \) represents primes \( p \equiv 1, 9, 31, 39 \mod 40 \) while \( 2x^2 - 5y^2 \) represents primes \( p \equiv 3, 13, 27, 37 \mod 40 \).

**Proof of the Proposition:** For an edge in the topograph labeled \( h \) with adjacent regions labeled \( n \) and \( k \) we have \( \Delta = h^2 - 4nk \). If \( p \) is a prime dividing \( \Delta \) this implies that
$4nk \equiv h^2 \mod p$. Thus if neither $n$ nor $k$ is divisible by $p$ and $p$ is odd we have $(\frac{4nk}{p}) = +1$. Since $(\frac{4nk}{p}) = (\frac{h}{p})(\frac{n}{p})(\frac{k}{p})$ and $(\frac{h}{p}) = +1$ this implies $(\frac{n}{p}) = (\frac{k}{p})$. In other words, the symbol $(\frac{n}{p})$ takes the same value on any two adjacent regions of the topograph of $Q$ labeled by numbers not divisible by $p$. To finish the proof we will use the following fact:

Lemma 6.11. Given a form $Q$ and a prime $p$ dividing the discriminant of $Q$, then any two regions in the topograph of $Q$ where the value of $Q$ is not divisible by $p$ can be connected by a path passing only through such regions.

Assuming this, the proposition easily follows since we have seen that the value of $(\frac{n}{p})$ is the same for any two adjacent regions with label not divisible by $p$. $\square$

Proof of the Lemma: Let us call regions in the topograph of $Q$ whose label is not divisible by $p$ good regions, and the other regions bad regions. We can assume that at least one region is good, otherwise there is nothing to prove. What we will show is that no two bad regions can be adjacent. Thus a path in the topograph from one good region to another cannot pass through two consecutive bad regions, and if it does pass through a bad region then a detour around this region allows this bad region to be avoided, creating a new path passing through one fewer bad region as in the following figure:

By repeating this detouring process as often as necessary we eventually obtain a path avoiding bad regions entirely, still starting and ending at the same two given good regions.

To see that no two adjacent regions are bad, suppose this is false, so there are two adjacent regions whose $Q$ values $n$ and $k$ are both divisible by $p$. If the edge separating these two regions is labeled $h$ then we have an equation $\Delta = h^2 - 4nk$, and since we assume $p$ divides $\Delta$ this implies that $p$ divides $h$ as well as $n$ and $k$. Thus the form $nx^2 + hxy + ky^2$ is equal to $p$ times another form. This implies that all regions in the topograph are bad. However we assumed this was not the case, so we conclude that there are no adjacent bad regions. $\square$

A useful observation is that the value of $(\frac{n}{p})$ for numbers $n$ in the topograph of a form $ax^2 + bxy + cy^2$ with discriminant divisible by $p$ can always be determined just by looking at the coefficients $a$ and $c$. This is because $a$ and $c$ appear in adjacent regions of the topograph, so if both these coefficients were divisible by $p$, this would
classes of forms: different prime factors. For this discriminant there are four different equivalence classes of forms: \( Q_1 = x^2 + 21y^2 \), \( Q_2 = 3x^2 + 7y^2 \), \( Q_3 = 2x^2 + 2xy + 11y^2 \), and \( Q_4 = 5x^2 + 4xy + 5y^2 \). The topographs of these forms were shown earlier in the chapter. To see which odd primes are represented in discriminant \( \Delta \) as indicated in the table, since the product \( 2\Delta \) does not divide \( p \) and \( p \) can be used this to compute \( \left( \frac{p}{\Delta} \right) \).

Let us look at another example, the discriminant \( \Delta = -84 = -2^2 \cdot 3 \cdot 7 \) with three different prime factors. For this discriminant there are four different equivalence classes of forms: \( Q_1 = x^2 + 21y^2 \), \( Q_2 = 3x^2 + 7y^2 \), \( Q_3 = 2x^2 + 2xy + 11y^2 \), and \( Q_4 = 5x^2 + 4xy + 5y^2 \). The topographs of these forms were shown earlier in the chapter. To see which odd primes are represented in discriminant \( -84 \) we compute:

\[
\left( \frac{-84}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{3}{p} \right) \left( \frac{4}{p} \right) \left( \frac{7}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{3}{p} \right) \left( \frac{7}{p} \right)
\]

As in the example of \( \Delta = 40 \) we can make a table of the values of these Legendre symbols for the 24 numbers mod 84 that are not divisible by the prime divisors 2, 3, 7 of 84. Using the fact that the squares mod 3 are \( (\pm 1)^2 = 1 \) and the squares mod 7 are \( (\pm 1)^2 = 1 \), \( (\pm 2)^2 = 4 \), and \( (\pm 3)^2 = 2 \), we obtain the table below:

<table>
<thead>
<tr>
<th>( \left( \frac{1}{p} \right) )</th>
<th>( \left( \frac{3}{p} \right) )</th>
<th>( \left( \frac{4}{p} \right) )</th>
<th>( \left( \frac{7}{p} \right) )</th>
</tr>
</thead>
<tbody>
<tr>
<td>1</td>
<td>1</td>
<td>1</td>
<td>1</td>
</tr>
<tr>
<td>5</td>
<td>1</td>
<td>1</td>
<td>1</td>
</tr>
<tr>
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</tbody>
</table>

The twelve cases when the product \( \left( \frac{-1}{p} \right) \left( \frac{p}{3} \right) \left( \frac{p}{7} \right) \) is +1 give the congruence classes of primes not dividing \( \Delta \) that are represented by one of the four forms, and we can determine which form it is by looking at the values of \( \left( \frac{p}{3} \right) \) and \( \left( \frac{p}{7} \right) \) for each of the four forms. As noted earlier, these values can be computed directly from the coefficients of \( x^2 \) and \( y^2 \) that are not divisible by 3 for \( \left( \frac{p}{3} \right) \) or by 7 for \( \left( \frac{p}{7} \right) \). For example, for \( Q_2 = 3x^2 + 7y^2 \) the coefficient of \( y^2 \) tells us that \( \left( \frac{3}{p} \right) = \left( \frac{7}{p} \right) = +1 \) and the coefficient of \( x^2 \) tells us that \( \left( \frac{3}{p} \right) = \left( \frac{7}{p} \right) = -1 \). Thus we have \( \left( \frac{Q_2}{p} \right) = (+1, -1) \) for \( Q_2 \), and in a similar way we find that \( \left( \frac{Q_1}{p} \right) = (1, +1) \) for \( Q_1 = x^2 + 21y^2 \), \( (-1, +1) \) for \( Q_3 = 2x^2 + 2xy + 11y^2 \), and \( (-1, -1) \) for \( Q_4 = 5x^2 + 4xy + 5y^2 \). This allows us to determine which congruence classes of primes are represented by which forms, as indicated in the table, since the product \( \left( \frac{-1}{p} \right) \left( \frac{p}{3} \right) \left( \frac{p}{7} \right) \) must be +1.

Another case we looked at was \( \Delta = -56 \) where there were three inequivalent forms \( Q_1 = x^2 + 14y^2 \), \( Q_2 = 2x^2 + 7y^2 \), and \( Q_3 = 3x^2 + 2xy + 5y^2 \). Here we have \( \left( \frac{-56}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right) \left( \frac{7}{p} \right) = \left( \frac{2}{p} \right) \left( \frac{7}{p} \right) \). The table of values for these Legendre symbols for congruence classes of numbers mod 56 not divisible by 2 or 7 is:
The prime divisors of 56. Moreover, since each odd prime $p$ dividing $-\Delta$ we can deduce that primes $p \equiv 1, 9, 15, 23, 25, 39 \mod 56$ are represented by $Q_1$, primes $p \equiv 3, 5, 13, 19, 27, 45 \mod 56$ are represented by $Q_2$, while primes $p \equiv 1, 9, 15, 23, 25, 39 \mod 56$ are represented by $Q_3$. However the values of the Legendre symbols in the table do not allow us to distinguish between $Q_1$ and $Q_2$.

Each row in one of the tables above can be regarded as a function assigning a number $\pm 1$ to each congruence class of numbers $n$ coprime to the discriminant $\Delta$. Such a function is called a character and the table is called a character table. For each odd prime $p$ dividing $\Delta$ there is a character given by the Legendre symbol $\left( \frac{n}{p} \right)$. There is sometimes also a character associated to the prime 2 in a somewhat less transparent way. In the example $\Delta = -84$ this is the character defined by the first row of the table, which assigns the values $+1$ to numbers $n = 4k + 1$ and $-1$ to numbers $n = 4k + 3$. We will denote this character by $\chi_4$ to indicate that its values $\chi_4(n) = \pm 1$ depend only on the value of $n \mod 4$. Thus $\chi_4(p) = \left( \frac{-1}{p} \right)$ when $p$ is an odd prime, but $\chi_4(n)$ is defined for all odd numbers $n$, not just primes. One can check that an explicit formula for $\chi_4$ is $\chi_4(n) = (-1)^{(n-1)/2}$ although we will not be needing this formula.

In the example with $\Delta = -56$ the character corresponding to the prime 2 is given by the row labeled $\left( \frac{2}{p} \right)$. This character associates the value $+1$ to an odd number $n \equiv \pm 1 \mod 8$ and the value $-1$ when $n \equiv \pm 3 \mod 8$. We will denote it by $\chi_8$ since its values $\chi_8(n) = \pm 1$ depend only on $n \mod 8$. We have $\chi_8(p) = \left( \frac{2}{p} \right)$ for all odd primes $p$, but $\chi_8(n)$ is defined for all odd numbers $n$. There is again an explicit formula $\chi_8(n) = (-1)^{(n^2-1)/8}$ that we will not use.

By analogy we can also introduce the notation $\chi_p(n) = \left( \frac{n}{p} \right)$ for $p$ an odd prime not dividing $n$. From the table we see that $\left( \frac{2}{p} \right) \left( \frac{p}{7} \right)$ is $(+1)(+1)$ for $p \equiv 1, 9, 15, 23, 25, 39 \mod 56$ and $(-1)(-1)$ for $p \equiv 3, 5, 13, 19, 27, 45 \mod 56$. Thus the primes that are represented in discriminant $-56$ are the primes in these twelve congruence classes, along with 2 and 7, the prime divisors of 56. Moreover, since $\left( \frac{p}{7} \right)$ has the value $+1$ for numbers in the topographs of $Q_1$ and $Q_2$ not divisible by 7, and the value $-1$ for numbers in the topograph of $Q_3$ not divisible by 7, we can deduce that primes $p \equiv 1, 9, 15, 23, 25, 39 \mod 56$ are represented by $Q_1$ or $Q_2$ while primes $p \equiv 3, 5, 13, 19, 27, 45 \mod 56$ are represented by $Q_3$. However the values of the Legendre symbols in the table do not allow us to distinguish between $Q_1$ and $Q_2$.

<table>
<thead>
<tr>
<th>$(\frac{2}{p})$</th>
<th>$Q_3$</th>
<th>$Q_3$</th>
<th>$Q_3$</th>
<th>$Q_3$</th>
<th>$Q_3$</th>
</tr>
</thead>
<tbody>
<tr>
<td>$+1$</td>
<td>$-1$</td>
<td>$+1$</td>
<td>$-1$</td>
<td>$+1$</td>
<td>$-1$</td>
</tr>
<tr>
<td>$+1$</td>
<td>$-1$</td>
<td>$+1$</td>
<td>$-1$</td>
<td>$-1$</td>
<td>$+1$</td>
</tr>
<tr>
<td>$(\frac{Q_1}{Q_2})$</td>
<td>$29$</td>
<td>$31$</td>
<td>$33$</td>
<td>$37$</td>
<td>$39$</td>
</tr>
<tr>
<td>$(\frac{Q_1}{Q_2})$</td>
<td>$41$</td>
<td>$43$</td>
<td>$45$</td>
<td>$47$</td>
<td>$51$</td>
</tr>
<tr>
<td>$-1$</td>
<td>$+1$</td>
<td>$+1$</td>
<td>$-1$</td>
<td>$-1$</td>
<td>$+1$</td>
</tr>
<tr>
<td>$+1$</td>
<td>$-1$</td>
<td>$+1$</td>
<td>$+1$</td>
<td>$-1$</td>
<td>$-1$</td>
</tr>
<tr>
<td>$(\frac{Q_1}{Q_2})$</td>
<td>$23$</td>
<td>$25$</td>
<td>$27$</td>
<td>$31$</td>
<td>$33$</td>
</tr>
</tbody>
</table>

<table>
<thead>
<tr>
<th>$(\frac{2}{p})$</th>
<th>$Q_3$</th>
<th>$Q_3$</th>
<th>$Q_3$</th>
<th>$Q_3$</th>
</tr>
</thead>
<tbody>
<tr>
<td>$-1$</td>
<td>$+1$</td>
<td>$+1$</td>
<td>$-1$</td>
<td>$+1$</td>
</tr>
<tr>
<td>$+1$</td>
<td>$-1$</td>
<td>$+1$</td>
<td>$+1$</td>
<td>$-1$</td>
</tr>
<tr>
<td>$(\frac{Q_1}{Q_2})$</td>
<td>$29$</td>
<td>$31$</td>
<td>$33$</td>
<td>$37$</td>
</tr>
</tbody>
</table>

From the table we see that $\left( \frac{2}{p} \right) \left( \frac{p}{7} \right)$ is $(+1)(+1)$ for $p \equiv 1, 9, 15, 23, 25, 39 \mod 56$ and $(-1)(-1)$ for $p \equiv 3, 5, 13, 19, 27, 45 \mod 56$. Thus the primes that are represented in discriminant $-56$ are the primes in these twelve congruence classes, along with 2 and 7, the prime divisors of 56. Moreover, since $\left( \frac{p}{7} \right)$ has the value $+1$ for numbers in the topographs of $Q_1$ and $Q_2$ not divisible by 7, and the value $-1$ for numbers in the topograph of $Q_3$ not divisible by 7, we can deduce that primes $p \equiv 1, 9, 15, 23, 25, 39 \mod 56$ are represented by $Q_1$ or $Q_2$ while primes $p \equiv 3, 5, 13, 19, 27, 45 \mod 56$ are represented by $Q_3$. However the values of the Legendre symbols in the table do not allow us to distinguish between $Q_1$ and $Q_2$.
Chapter 6  Representations by Quadratic Forms  150

Let us consider now how characters can be associated to the prime 2 in general. Since characters arise from primes that divide the discriminant, this means we are interested in even discriminants, and the characters we seek should assign a value \( \pm 1 \) to each odd number. We would like the analog of Proposition 6.10 to hold, so characters for the prime 2 should take the same value on all odd numbers in the topograph of a form of the given discriminant. By Lemma 6.11 this just means that the characters should have the same value for odd numbers in adjacent regions of the topographs.

Even discriminants are always multiples of 4, so we can write an even discriminant as \( \Delta = 4\delta \). For adjacent regions in a topograph with labels \( n \) and \( k \) we have \( \Delta = h^2 - 4nk \) where \( h \) is the label on the edge between the two regions. Since \( \Delta \) is even, so is \( h\) and we can write \( h = 2l \). The discriminant equation then becomes \( 4\delta = 4l^2 - 4nk \) or just \( \delta = l^2 - nk \).

There will turn out to be six different cases. The first two are when \( \delta \) is odd, which means that \( \Delta \) is divisible by 4 but not 8. In these two cases we consider congruences mod 4, the highest power of 2 dividing \( \Delta \). Since \( \delta \) is odd and both \( n \) and \( k \) are odd, the equation \( \delta = l^2 - nk \) implies that \( l \) must be even, so \( l^2 \equiv 0 \pmod{4} \) and we have \( nk \equiv -\delta \pmod{4} \). Multiplying both sides of this congruence by \( k \), we get \( n \equiv -\delta k \pmod{4} \) since \( k^2 \equiv 1 \pmod{4} \), \( k \) being odd. Multiplying the congruence \( n \equiv -\delta k \) by \( k \) again gives the previous congruence \( nk \equiv -\delta \) so the two congruences are equivalent.

**Case 1:** \( \delta = 4m - 1 \). The congruence condition \( n \equiv -\delta k \pmod{4} \) is then \( n \equiv k \pmod{4} \). Thus Lemma 6.11 implies that the character \( \chi_4 \) assigning \( +1 \) to integers \( 4s + 1 \) and \( -1 \) to integers \( 4s - 1 \) has the same value for all odd numbers in the topograph of a form of discriminant \( \Delta = 4(4m - 1) \). We might try reversing the values of \( \chi_4 \), assigning the value \( +1 \) to integers \( 4s - 1 \) and \( -1 \) to integers \( 4s + 1 \), but this just gives the function \( -\chi_4 \) which doesn’t really give any new information that \( \chi_4 \) doesn’t give. In practice \( \chi_4 \) turns out to be more convenient to use than \( -\chi_4 \) would be.

An example for the case \( \delta = 4m - 1 \) is the discriminant \( \Delta = -84 \) considered earlier, where the first row of the character table gave the values for \( \chi_4 \).

**Case 2:** \( \delta = 4m + 1 \). The difference from the previous case is that the congruence condition is now \( n \equiv -k \pmod{4} \). This means the mod 4 value of odd numbers in the topograph is not constant, and so we do not get a character for the prime 2. As an example, consider the form \( x^2 + 3y^2 \) with \( \Delta = -12 \).
As one can see, there are odd numbers in the topograph of \( x^2 + 3y^2 \) congruent to both 1 and 3 mod 4. The situation is not improved by considering odd numbers mod 8 instead of mod 4 since the topograph contains numbers congruent to each of 1, 3, 5, 7 mod 8. Trying congruences modulo higher powers of 2 does not help either.

The absence of a character for the prime 2 when \( \delta = 4m + 1 \) could perhaps be predicted from the calculation of \( (\Delta/p) \). Since \( \delta \) is odd we have \( \Delta = 4\delta = 4p_1 \cdots p_r \) for odd primes \( p_i \) and so \( (\Delta/p) = (p_1/p) \cdots (p_r/p) \). This equals \( (p/p_1) \cdots (p/p_r) \) since the number of \( p_i \)'s congruent to 3 mod 4 is even when \( \delta = 4m + 1 \). Thus the value of \( (\Delta/p) \) depends only on the characters associated to the odd prime factors of \( \Delta \).

There remain the cases that \( \delta \) is even. The next two cases are when \( \Delta \) is divisible by 8 but not by 16. After that is the case that \( \Delta \) is divisible by 16 but not by 32, and finally the case that \( \Delta \) is divisible by 32. In all these cases we will consider congruences mod 8, so the equation \( \delta = l^2 - nk \) becomes \( \delta \equiv l^2 - nk \mod 8 \). Since \( \delta \) is now even while \( n \) and \( k \) are still odd, this congruence implies \( l \) is odd, and so \( l^2 \equiv 1 \mod 8 \) and the congruence can be written as \( nk \equiv 1 - \delta \mod 8 \). Since \( k^2 \equiv 1 \mod 8 \) when \( k \) is odd, we can multiply both sides of the congruence \( nk \equiv 1 - \delta \mod 8 \) by \( k \) to obtain the equivalent congruence \( n \equiv (1 - \delta)k \mod 8 \).

**Case 3:** \( \delta \equiv 2 \mod 8 \). The congruence is then \( n \equiv -k \mod 8 \). It follows that in the topograph of a form of discriminant \( \Delta = 4(8m + 2) \) either the odd numbers must all be congruent to \( \pm 1 \mod 8 \) or they must all be congruent to \( \pm 3 \mod 8 \). Thus the character \( \chi_8 \) which takes the value +1 on numbers \( 8s \pm 1 \) and −1 on numbers \( 8s \pm 3 \) has a constant value, either +1 or −1, for all odd numbers in the topograph.

An example for this case is \( \Delta = 40 \). Here the two rows of the character table computed earlier gave the values for \( \chi_8 \) and \( \chi_5 \).

**Case 4:** \( \delta \equiv 6 \mod 8 \). Now the congruence \( n \equiv (1 - \delta)k \mod 8 \) becomes \( n \equiv -5k \), or equivalently \( n \equiv 3k \mod 8 \). This implies that all odd numbers in the topograph of a form of discriminant \( \Delta = 4(8m + 6) \) must be congruent to 1 or 3 mod 8, or they must all be congruent to 5 or 7 mod 8. The character associated to the prime 2 in this case has the value +1 on numbers \( 8s + 1 \) and \( 8s + 3 \), and the value −1 on numbers \( 8s + 5 \) and \( 8s + 7 \). We have not encountered this character previously, so
let us give it the new name $\chi'_8$. However, it is not entirely new since it is actually just
the product $\chi_4 \chi_8$ as one can easily check by evaluating this product on $1, 3, 5, \text{and} 7$.

A simple example is $\Delta = -8$ with class number 1. Here we have $(\frac{\Delta}{p}) = (\frac{-8}{p}) = (\frac{-1}{p}) (\frac{2}{p})$ which equals $+1$ for $p \equiv 1, 3 \text{ mod } 8$ and $-1$ for $p \equiv 5, 7 \text{ mod } 8$ so this is just the character $\chi'_8$.

Another example is $\Delta = 24$ where there are the two forms $Q_1 = x^2 - 6y^2$ and $Q_2 = 6x^2 - y^2$. We have $(\frac{\Delta}{p}) = (\frac{24}{p}) = (\frac{2}{p}) (\frac{2}{p}) = (\frac{2}{p}) (\frac{-1}{p}) (\frac{3}{p})$. The character table is

<table>
<thead>
<tr>
<th></th>
<th>1</th>
<th>5</th>
<th>7</th>
<th>11</th>
<th>13</th>
<th>17</th>
<th>19</th>
<th>23</th>
<th>25</th>
<th>29</th>
<th>31</th>
<th>35</th>
<th>37</th>
<th>41</th>
<th>43</th>
<th>47</th>
</tr>
</thead>
<tbody>
<tr>
<td>$\chi'_8$</td>
<td>+1</td>
<td>-1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>-1</td>
</tr>
<tr>
<td>$\chi_3$</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
</tr>
</tbody>
</table>

Thus $Q_1$ represents primes $p \equiv 1, 19 \text{ mod } 24$ and $Q_2$ represents primes $p \equiv 5, 23 \text{ mod } 24$.

**Case 5:** $\delta \equiv 4 \text{ mod } 8$. Now we have the congruence $n \equiv -3k \text{ mod } 8$. Thus in
the topograph of a form of discriminant $\Delta = 4(8m + 4)$ all odd numbers must be congruent to 1 or 5 mod 8, or they must all be congruent to 3 or 7 mod 8. More simply, one can say that all odd numbers in the topograph must be congruent to 1 mod 4 or they must all be congruent to 3 mod 4. Thus we obtain the character $\chi_4$ again.

An example is $\Delta = -48$ where we have the two forms $Q_1 = x^2 + 12y^2$ and $Q_2 = 3x^2 + 4y^2$ as well as the nonprimitive forms $Q_3 = 2x^2 + 6y^2$ and $Q_4 = 4x^2 + 4xy + 4y^2$. We have $(\frac{\Delta}{p}) = (\frac{-3}{p}) = (\frac{-1}{p}) (\frac{3}{p}) = (\frac{9}{p})$. This is the character $\chi_3$. We also have the character $\chi_4$ that we just described. The character table is

<table>
<thead>
<tr>
<th></th>
<th>1</th>
<th>5</th>
<th>7</th>
<th>11</th>
<th>13</th>
<th>17</th>
<th>19</th>
<th>23</th>
<th>25</th>
<th>29</th>
<th>31</th>
<th>35</th>
<th>37</th>
<th>41</th>
<th>43</th>
<th>47</th>
</tr>
</thead>
<tbody>
<tr>
<td>$\chi_4$</td>
<td>+1</td>
<td>+1</td>
<td>-1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>-1</td>
</tr>
<tr>
<td>$\chi_3$</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
<td>+1</td>
<td>-1</td>
</tr>
</tbody>
</table>

The columns repeat every four columns since $(\frac{-1}{p})$ and $(\frac{9}{p})$ are determined by the value of $p \text{ mod } 12$. In contrast with earlier examples, the representability of a prime $p > 3$ in discriminant $-48$ is determined by one character, $\chi_3$, and the other character $\chi_4$ serves only to decide which of the forms $Q_1$ and $Q_2$ achieves the representation. Note that $\chi_4$ says nothing about the nonprimitive forms $Q_3$ and $Q_4$ whose values are all even. On the other hand, from $\chi_3$ we can deduce that all values of $Q_3$ not divisible by 3 must be congruent to 2 mod 3 while for $Q_4$ they must be congruent to 1 mod 3.

**Case 6:** $\delta \equiv 0 \text{ mod } 8$, so $\Delta$ is a multiple of 32. In this case the congruence $n \equiv (1-\delta)k \text{ mod } 8$ becomes simply $n \equiv k \text{ mod } 8$. Thus all odd numbers in the topograph of a form of discriminant $\Delta = 32m$ must lie in the same congruence class mod 8. The two characters $\chi_4$ and $\chi_8$ can now both occur independently, as shown in the following chart listing their values on the four classes 1, 3, 5, 7 mod 8:
As an example consider the discriminant $\Delta = -32$. Here there are two primitive forms $Q_1 = x^2 + 8y^2$ and $Q_2 = 3x^2 + 2xy + 3y^2$ along with one nonprimitive form $Q_3 = 2x^2 + 4y^2$. We have $\left(\frac{\Delta}{p}\right) = \left(\frac{-2}{p}\right) = \left(\frac{-1}{p}\right)\left(\frac{2}{p}\right)$ with the two factors being the two independent characters for the prime 2. The full character table is then just a four-fold repetition of the previous shorter table:

| 1 3 5 7 | \chi_4 | +1 -1 +1 -1 |
|         | \chi_8 | +1 -1 -1 +1 |

This finishes the analysis of the six cases for characters associated to the prime 2. The results are summarized in the following table, where we have interchanged the first two cases for convenience:

<table>
<thead>
<tr>
<th>$\Delta$</th>
<th>$4(4m + 1)$</th>
<th>$4(4m + 3)$</th>
<th>$8(4m + 1)$</th>
<th>$8(4m + 3)$</th>
<th>$16(2m + 1))$</th>
<th>$32m$</th>
</tr>
</thead>
<tbody>
<tr>
<td>$\chi$</td>
<td>$\chi_4$</td>
<td>$\chi_8$</td>
<td>$\chi_8'$</td>
<td>$\chi_4$</td>
<td>$\chi_4$, $\chi_8$</td>
<td></td>
</tr>
</tbody>
</table>

We have now defined a set of characters for each discriminant $\Delta$, with one character for each odd prime dividing $\Delta$ and either zero, one, or two characters for the prime 2 when $\Delta$ is even. The character table for discriminant $\Delta$ has one row for each of these characters.

If one restricts attention to fundamental discriminants then in the table above only the second, third, and fourth columns on the right can arise since in the other three cases there always exist nonprimitive forms of the given discriminant. Thus the characters that arise in the three cases of fundamental discriminants are exactly $\chi_4$, $\chi_8$, and $\chi_8'$.

Another useful observation is that the value of the Legendre symbol $\left(\frac{\Delta}{p}\right)$ is determined just by the characters for discriminant $\Delta$. The formulas for $\left(\frac{\Delta}{p}\right)$ are given by the following table from Proposition 6.9:

<table>
<thead>
<tr>
<th>$\Delta$</th>
<th>$\left(\frac{\Delta}{p}\right)$</th>
</tr>
</thead>
<tbody>
<tr>
<td>$2^{2l}(4m + 1)$</td>
<td>$\left(\frac{p}{p_1}\right) \cdots \left(\frac{p}{p_k}\right)$</td>
</tr>
<tr>
<td>$2^{2l}(4m + 3)$</td>
<td>$\left(\frac{-1}{p}\right) \left(\frac{p}{p_1}\right) \cdots \left(\frac{p}{p_k}\right)$</td>
</tr>
<tr>
<td>$2^{2l+1}(4m + 1)$</td>
<td>$\left(\frac{2}{p}\right) \left(\frac{p}{p_1}\right) \cdots \left(\frac{p}{p_k}\right)$</td>
</tr>
<tr>
<td>$2^{2l+1}(4m + 3)$</td>
<td>$\left(\frac{-1}{p}\right) \left(\frac{2}{p}\right) \left(\frac{p}{p_1}\right) \cdots \left(\frac{p}{p_k}\right)$</td>
</tr>
</tbody>
</table>

Characters obviously determine the terms $\left(\frac{p}{p_i}\right)$ in these formulas. For the prime 2 one has to compare this table with the previous one. The first four cases in the previous table fall into the four corresponding cases in this table, so characters determine $\left(\frac{\Delta}{p}\right)$.
in these cases. When \( \Delta = 16(2m + 1) \) in the previous table we are in one of the first two cases here, so we have \( \chi_4 \) available to determine the value of \( \left( \frac{\Delta}{p} \right) \). Finally, when \( \Delta = 32m \) both \( \chi_4 \) and \( \chi_8 \) are available so \( \left( \frac{\Delta}{p} \right) \) is again determined by the characters.

A nice property satisfied by characters is that they are multiplicative, so \( \chi(mn) = \chi(m)\chi(n) \) for all \( m \) and \( n \) for which \( \chi \) is defined. For the characters \( \chi_p \) associated to odd primes \( p \) this is just the basic property \( \left( \frac{mn}{p} \right) = \left( \frac{m}{p} \right)\left( \frac{n}{p} \right) \) of Legendre symbols. For the prime 2 the characters \( \chi_4 \) and \( \chi_8 \) are multiplicative as well. For \( \chi_4 \) this holds since \( \chi_4(1 \cdot 1) = +1 = \chi_4(1)\chi_4(1) \), \( \chi_4(1 \cdot 3) = -1 = \chi_4(1)\chi_4(3) \), and \( \chi_4(3 \cdot 3) = +1 = \chi_4(3)\chi_4(3) \). Similarly for \( \chi_8 \) we have \( \chi_8(\pm1 \cdot \pm1) = +1 = \chi_8(\pm1)\chi_8(\pm1) \), \( \chi_8(\pm1 \cdot \pm3) = -1 = \chi_8(\pm1)\chi_8(\pm3) \), and \( \chi_8(\pm3 \cdot \pm3) = +1 = \chi_8(\pm3)\chi_8(\pm3) \). The multiplicativity of \( \chi_8' \) follows since \( \chi_8' = \chi_4\chi_8 \).

In fact \( \chi_4, \chi_8, \) and \( \chi_8' \) are the only multiplicative functions from the odd integers mod 8 to \( \{ \pm1 \} \), apart from the trivial function assigning \(+1\) to all four of \( 1, 3, 5, 7 \). To see this, note first that each of \( 3, 5, 7 \) has square equal to \( 1 \) mod 8 and the product of any two of \( 3, 5, 7 \) is the third, mod 8. This means that a multiplicative function \( \chi \) from odd integers mod 8 to \( \{ \pm1 \} \) is completely determined by the two values \( \chi(3) \) and \( \chi(5) \) since \( \chi(1) = \chi(3)\chi(3) \) and \( \chi(7) = \chi(3)\chi(5) \). For \( \chi_4 \) the values on \( 3 \) and \( 5 \) are \(-1, +1\), for \( \chi_8 \) they are \(-1, -1\), and for \( \chi_8' = \chi_4\chi_8 \) they are \(+1, -1\). The only other possibility is \(+1, +1\) but this leads to the trivial character.

Having the full list of characters now, our next goal will be to verify that some of the more salient features of the character tables we have looked at hold in general.

**Proposition 6.12.** Character tables have the following properties:

1. The columns contain all possible combinations of \(+1\) and \(-1\).
2. Each such combination occurs in the same number of columns.
3. If the discriminant \( \Delta \) is not a square then half of the columns have \( \left( \frac{\Delta}{p} \right) = +1 \) and half have \( \left( \frac{\Delta}{p} \right) = -1 \) for primes \( p \) in the congruence class corresponding to the column.

For example, when \( \Delta \) is a fundamental discriminant the formula for \( \left( \frac{\Delta}{p} \right) \) shows that it is just the product of all the characters in the character table, so the combinations of \( \pm1 \)'s that give \( \left( \frac{\Delta}{p} \right) = +1 \) in these cases are just the combinations with an even number of \( -1 \)'s. This need not be true for nonfundamental discriminants as the earlier example \( \Delta = -48 \) shows.

From statement (3) in the Proposition we immediately deduce:

**Corollary 6.13.** For hyperbolic and elliptic forms, the primes not dividing the discriminant \( \Delta \) that are represented by forms of discriminant \( \Delta \) are the primes in exactly half of the congruence classes mod \( \Delta \) of numbers coprime to \( \Delta \).

The proof of Proposition 6.12 will use the following general fact about congruences often referred to as the **Chinese Remainder Theorem** since it was used in ancient
Chapter 6  
Representations by Quadratic Forms

Chinese manuscripts to solve mathematical puzzles of a certain type:

**Proposition 6.14.** A collection of congruence conditions

\[ x \equiv a_1 \mod m_1 \]
\[ x \equiv a_2 \mod m_2 \]
\[ \vdots \]
\[ x \equiv a_k \mod m_k \]

always has a simultaneous solution provided that no two \( m_i \)'s have a common divisor greater than 1. Moreover such a solution is unique modulo the product \( m_1 \cdots m_k \).

The hypothesis that the various moduli \( m_i \) are coprime is certainly necessary. For example, there is no common solution to the two congruences \( x \equiv 5 \mod 6 \) and \( x \equiv 7 \mod 15 \) since the first congruence implies \( x \equiv 2 \mod 3 \) while the second congruence implies \( x \equiv 1 \mod 3 \).

**Proof:** Let us first prove the existence of a common solution \( x \) when there are just two congruences \( x \equiv a_1 \mod m_1 \) and \( x \equiv a_2 \mod m_2 \). In this case the desired number \( x \) will have the form \( x = a_1 + x_1 m_1 = a_2 + x_2 m_2 \) for some pair of yet-to-be-determined numbers \( x_1 \) and \( x_2 \). We can rewrite the equation \( a_1 + x_1 m_1 = a_2 + x_2 m_2 \) as \( a_2 - a_1 = m_1 x_1 - m_2 x_2 \). From our study of linear Diophantine equations in Chapter 2 we know that this equation has a solution \( (x_1, x_2) \) with integers \( x_1 \) and \( x_2 \) whenever \( m_1 \) and \( m_2 \) are coprime. This gives a simultaneous solution for the original two congruences \( x \equiv a_1 \mod m_1 \) and \( x \equiv a_2 \mod m_2 \), namely \( x = a_1 + x_1 m_1 = a_2 + x_2 m_2 \).

For more than two congruences we may suppose by induction that we have a number \( x = a \) satisfying all but the last congruence \( x \equiv a_k \mod m_k \). From the preceding paragraph we know that a number \( x \) exists satisfying the two congruences \( x \equiv a \mod m_1 \cdots m_{k-1} \) and \( x \equiv a_k \mod m_k \). This gives the desired solution to all \( k \) congruences since \( a \equiv a_i \mod m_i \) for each \( i < k \) by the inductive hypothesis.

Now we show that a simultaneous solution \( x \) of the given set of congruences is unique modulo \( m_1 \cdots m_k \). Let \( y \) be another solution. Then the difference \( x - y \) is congruent to 0 mod each of the numbers \( m_1, \cdots, m_k \), which means that it is divisible by each \( m_i \) and hence by their product since they have no common factors. Thus \( x \equiv y \mod m_1 \cdots m_k \). \( \square \)

Another fact that will be needed in the proof of Proposition 6.12 is the following:

**Lemma 6.15.** For a power \( q^r \) of an odd prime \( q \) exactly half of the \( q^r - q^{r-1} \) congruence classes mod \( q^r \) of numbers \( a \) not divisible by \( q \) satisfy \( \left( \frac{a}{q} \right) = +1 \).

**Proof:** First we do the case \( r = 1 \). The \( q - 1 \) nonzero congruence classes mod \( q \) are \( \pm 1, \pm 2, \cdots, \pm (q - 1)/2 \). Their squares \( (\pm 1)^2, (\pm 2)^2, \cdots, (\pm (q - 1)/2)^2 \) are all distinct since if \( a^2 \equiv b^2 \mod q \) then \( q \) divides \( a^2 - b^2 = (a - b)(a + b) \), so since \( q \) is
prime it must divide either \( a - b \) or \( a + b \) which means that either \( a \equiv b \) or \( a \equiv -b \) mod \( q \). Thus there are exactly \((q - 1)/2\) squares mod \( q \), which is half of the \( q - 1 \) nonzero congruence classes.

When \( r > 1 \) the congruence classes mod \( q^r \) we are counting are obtained from the numbers \( a \) in the interval \([0, q]\) with \( \left( \frac{a}{q} \right) = +1 \) by translating these numbers into the intervals \([q, 2q]\), then \([2q, 3q]\), and so on. There are \( q^{r-1} \) of these intervals in \([0, q^r]\). Thus half of the \( q^{r-1}(q - 1) = q^r - q^{r-1} \) congruence classes mod \( q^r \) of numbers not divisible by \( q \) are squares mod \( q \).

\[
\text{Proof of Proposition 6.12: Let us write } \Delta = \varepsilon 2^s p_1^{r_1} \cdots p_k^{r_k} \text{ where } \varepsilon = \pm 1, \ s \geq 0, \text{ and the } p_i \text{’s are the distinct odd prime divisors of } \Delta. \text{ Thus the characters for this discriminant are } \chi_{p_1}, \ldots, \chi_{p_k} \text{ and either zero, one, or two more characters associated to the prime 2 when } s > 0.
\]

To prove statement (1) choose numbers \( a_1, \ldots, a_k \) realizing preassigned values \( \varepsilon_i = \pm 1 \) for each \( \chi_{p_i} \). When \( s > 0 \) we also choose a number 1, 3, 5, or 7 to realize any preassigned pair of values for \( \chi_4 \) and \( \chi_8 \), hence for any preassigned values for the characters associated to the prime 2. By the Chinese Remainder Theorem there is a number \( a \) congruent to each \( a_i \) mod \( p_i^{r_i} \) and to the chosen number 1,3,5,7 mod 8. The number \( a \) is coprime to \( \Delta \) since it is nonzero mod \( p_i^{r_i} \) for each \( i \) and is odd when \( s > 0 \). Thus the column in the character table corresponding to \( a \) realizes the chosen values for all the characters, proving (1).

To prove (2) we will count the number of columns in the character table realizing a given combination of values \( \pm 1 \) and see that this number does not depend on which combination is chosen. By the preceding Lemma the number of choices for \( a_i \) is \( p_i^{r_i-1}(p_i - 1)/2 \), so the Chinese Remainder Theorem implies that when \( s = 0 \) the number of congruence classes mod \( \Delta \) realizing a given combination of values \( \pm 1 \) is the product of these numbers \( p_i^{r_i-1}(p_i - 1)/2 \). When \( s > 0 \) but there is no character for the prime 2 the product of the numbers \( p_i^{r_i-1}(p_i - 1)/2 \) is multiplied by \( 2^{s-1} \) since this is the number of odd congruence classes mod \( 2^s \). If there is one character for the prime 2 the number \( 2^{s-1} \) is cut in half, and if there are two characters for the prime 2 it is cut in half again. Thus in all cases the number of columns realizing a given combination of \( \pm 1 \)’s is independent of the combination.

For (3) a formula for \( \left( \frac{\Delta}{p} \right) \) is given in Proposition 6.9 with four different cases depending on the prime factorization of \( \Delta \), and with the notational difference that the primes \( p_i \) there are not required to be distinct. If \( \Delta \) is a square then the applicable formula is the first of the four formulas since an odd square is 1 mod 4, and in fact the formula degenerates to just the constant +1 since its terms all cancel out, as each prime factor of \( \Delta \) occurs to an even power. However when \( \Delta \) is not a square then the terms in the first of the four formulas do not all cancel out, and in the other three formulas as well there is always at least one term remaining after cancellations, either
\[\left(\frac{-1}{p}\right), \left(\frac{2}{p}\right),\text{ or the product } \left(\frac{-1}{p}\right)\left(\frac{2}{p}\right)\text{ in the fourth formula.}\]

In view of property (2), to prove (3) it will suffice to show that when \(\Delta\) is not a square, the set of combinations of values \(\pm 1\) in the character table that give \(\left(\frac{\Delta}{p}\right) = +1\) has the same number of elements as the set of combinations that give \(\left(\frac{\Delta}{p}\right) = -1\). But this is obvious since we can interchange these two sets by choosing one term in the formula for \(\left(\frac{\Delta}{p}\right)\) (after cancellation) and switching the sign of the value \(\pm 1\) for this term, keeping the values for the other characters unchanged.

\[\square\]

To conclude this section we will say a few more things about the Chinese Remainder Theorem. First we describe a geometric picture that makes the content of the Chinese Remainder Theorem a little more visual. Consider the case of two simultaneous congruences \(x \equiv a \mod m\) and \(x \equiv b \mod n\) where \(m\) and \(n\) are coprime. We can then label the \(mn\) unit squares in an \(m \times n\) rectangle by the numbers \(1, 2, 3, \ldots\) starting in the lower left corner and continuing upward to the right at a 45 degree angle as shown in the following figure for the case of a \(4 \times 9\) rectangle:

\[
\begin{array}{cccccccc}
28 & 20 & 12 & 4 & 32 & 24 & 16 & 8 & 36 \\
19 & 11 & 3 & 31 & 23 & 15 & 7 & 35 & 27 \\
10 & 2 & 30 & 22 & 14 & 6 & 34 & 26 & 18 \\
1 & 29 & 21 & 13 & 5 & 33 & 25 & 17 & 9 \\
\end{array}
\]

Whenever we run over the top edge we jump back to the bottom in order to continue, and when we reach the right edge we jump back to the left edge. This amounts to taking congruence classes mod \(m\) horizontally and mod \(n\) vertically. What the Chinese Remainder Theorem says is that when \(m\) and \(n\) are coprime, each unit square in the \(m \times n\) rectangle is labeled exactly once by a number from 1 to \(mn\). (Without the coprimeness some squares would have no labels while others would have multiple labels.) The figure thus illustrates that specifying a congruence class mod \(mn\) is equivalent to specifying a pair of congruence classes mod \(m\) and mod \(n\) via the projections onto the two axes.

For the case of three simultaneous congruences there is an analogous picture with a three-dimensional rectangular box partitioned into unit cubes. More generally, for \(k\) congruences one would be dealing with a \(k\)-dimensional box.

A common situation for applying the Chinese Remainder Theorem is to start with a number \(n\) factored into primes as \(n = p_1^{r_1} \cdots p_k^{r_k}\) for distinct primes \(p_1, \ldots, p_k\), so that a congruence \(x \equiv a \mod n\) is equivalent to a set of \(k\) congruences \(x \equiv a_i \mod p_i^{r_i}\). If we add the condition that each \(a_i\) is not divisible by the corresponding
prime $p_i$ then a simultaneous solution $x = a$ for all $k$ congruences must be coprime to $n$ since $a \equiv a_i \mod p_i^{r_i}$ implies $a \equiv a_i \mod p_i$ and we assume $a_i$ is nonzero mod $p_i$ so $a$ is also nonzero mod $p_i$. Conversely, if $a$ is coprime to $n$ and satisfies a set of congruences $a \equiv a_i \mod p_i^{r_i}$ and hence $a \equiv a_i \mod p_i$, then $a_i$ must be nonzero mod $p_i$ since $a$ is. Thus congruence classes mod $n$ of numbers $a$ coprime to $n$ are equivalent to congruence classes mod $p_i^{r_i}$ of numbers $a_i$ coprime to $p_i$, one for each $i$.

In the geometric picture for the case $k = 2$ with a rectangular array of unit squares, when we require $a_1$ to be coprime to $p_1$ we are omitting the numbers in certain vertical columns of squares, the columns whose horizontal coordinate is a multiple of $p_1$. Similarly, when we require $a_2$ to be coprime to $p_2$ we omit the numbers in the horizontal rows whose vertical coordinate is a multiple of $p_2$. The numbers in the boxes that are not omitted are then the numbers coprime to $n = p_1^{r_1} p_2^{r_2}$. Here is the picture for the case $n = 2^2 \cdot 3^2$ that we showed earlier:

<table>
<thead>
<tr>
<th></th>
<th>28</th>
<th>20</th>
<th>12</th>
<th>4</th>
<th>32</th>
<th>24</th>
<th>16</th>
<th>8</th>
<th>36</th>
</tr>
</thead>
<tbody>
<tr>
<td>1</td>
<td>19</td>
<td>11</td>
<td>3</td>
<td>31</td>
<td>23</td>
<td>15</td>
<td>7</td>
<td>35</td>
<td>27</td>
</tr>
<tr>
<td>10</td>
<td>2</td>
<td>30</td>
<td>22</td>
<td>14</td>
<td>6</td>
<td>34</td>
<td>26</td>
<td>18</td>
<td></td>
</tr>
<tr>
<td>1</td>
<td>29</td>
<td>21</td>
<td>13</td>
<td>5</td>
<td>33</td>
<td>25</td>
<td>17</td>
<td>9</td>
<td></td>
</tr>
</tbody>
</table>

For $k = 3$ we would be omitting the cubes in certain slices parallel to the three coordinate planes, and similarly for $k > 3$.

The function which assigns to each positive integer $n$ the number of congruence classes mod $n$ of numbers coprime to $n$ is called the Euler phi function $\varphi(n)$. The arguments above show that $\varphi(n) = \varphi(p_1^{r_1}) \cdots \varphi(p_k^{r_k})$ when $n = p_1^{r_1} \cdots p_k^{r_k}$ for distinct primes $p_i$. For a prime $p$ we have $\varphi(p^r) = p^r - p^{r-1} = p^{r-1}(p - 1)$ since we are counting how many numbers remain from $1, 2, \ldots, p^r$ when we delete $p, 2p, \ldots, (p^{r-1})p = p^r$. Thus there is an explicit formula for $\varphi(n)$ in terms of the prime factorization of $n$. For example in the figure above we see that $\varphi(36) = \varphi(4)\varphi(9) = 2 \cdot 6 = 12$.

The function $\varphi$ can be used to count the number of columns in a character table. For discriminant $\Delta$ these columns correspond to congruence classes mod $\Delta$ of number coprime to $\Delta$, so the number of columns is $\varphi(|\Delta|)$. For example for $\Delta = -84$ we have $\varphi(84) = \varphi(3)\varphi(4)\varphi(7) = 2 \cdot 2 \cdot 6 = 24$, as the character table earlier in the chapter showed.
Finally let us apply the Chinese Remainder Theorem to give an example of a Diophantine equation that has a solution mod $n$ for each positive integer $n$ but does not have an actual integer solution. The example is the equation $2x^2 + 7y^2 = 1$. This obviously has no integer solutions, although it does have rational solutions such as $(x, y) = (1/3, 1/3)$ and $(3/5, 1/5)$. (These could be found by looking for integer solutions of $2x^2 + 7y^2 = z^2$, or in other words, squares $z^2$ in the topograph of $2x^2 + 7y^2$.) The rational solution $(1/3, 1/3)$ will give an integer solution mod $n$ provided that 3 has a multiplicative inverse “1/3” mod $n$, which happens whenever 3 does not divide $n$. For example for $n = 22$ a multiplicative inverse for 3 is 15 since $3 \cdot 15 \equiv 1 \mod 22$, and this leads to the solution $2 \cdot 15^2 + 7 \cdot 15^2 = 2025 \equiv 1 \mod 22$. Thus we see that $2x^2 + 7y^2 = 1$ has a solution mod $n$ whenever 3 does not divide $n$. Using the other rational solution $(3/5, 1/5)$ in a similar fashion we can solve $2x^2 + 7y^2 = 1$ mod $n$ whenever 5 does not divide $n$ by finding a multiplicative inverse for 5 mod $n$. Now if we factor an arbitrary $n$ as $n_1n_2$ where $n_1$ and $n_2$ are coprime and 3 does not divide $n_1$ and 5 does not divide $n_2$, then solutions $(x_1, y_1)$ mod $n_1$ and $(x_2, y_2)$ mod $n_2$ can be combined by applying the Chinese Remainder Theorem twice to give a pair $(x, y)$ with $(x, y) \equiv (x_1, y_1)$ mod $n_1$ and $(x, y) \equiv (x_2, y_2)$ mod $n_2$, so we have $2x^2 + 7y^2 \equiv 1 \mod n_1$ and mod $n_2$, hence also mod $n$.

3. Genus

Recall the term “genus” that was introduced earlier: If two forms of the same discriminant cannot be distinguished by looking only at their values modulo the discriminant, then the two forms are said to be of the same genus. Here it is best to restrict attention just to primitive forms. We can now give this notion a more precise meaning by saying that two primitive forms of discriminant $\Delta$ have the same genus if each character for discriminant $\Delta$ takes the same value on the two forms, where the value of a character on a form means its value on all numbers in the topograph of the form not divisible by the prime associated to the character. In terms of character tables this is saying that two forms have the same genus if they realize the same combinations of +1’s and −1’s in the columns of the table.

The number of genera in discriminant $\Delta$ is thus at most $2^\kappa$ where $\kappa$ is the number of characters in discriminant $\Delta$. In all the character tables we have looked at, only half of the $2^\kappa$ symbols were actually realized by forms, and in fact this is true generally:

**Theorem 6.16.** *If $\Delta$ is not a square then the number of genera of primitive forms of discriminant $\Delta$ is $2^{\kappa-1}$ where $\kappa$ is the number of characters in discriminant $\Delta$.*

This turns out to be fairly hard to prove. The original proof by Gauss required a somewhat lengthy digression into the theory of quadratic forms in three variables. An exposition of this proof can be found in the book by Flath listed in the Bibliography. We will give a different proof that deduces the result rather quickly from things we have already done using Dirichlet’s Theorem about primes in arithmetic progressions.
discussed at the end of Section 6.1, which we will not prove. We will not need the full strength of Dirichlet’s theorem, and in fact all we will actually need is that each congruence class of numbers \( x \equiv b \mod a \) contains at least one odd prime if \( a \) and \( b \) are coprime. One might think this would be easier to prove than that there are infinitely many primes in the congruence class, but this seems not to be the case.

\textit{Proof of Theorem 6.16 using Dirichlet’s Theorem:} In the examples of character tables we have given, the columns corresponding to forms are the ones where \( \left( \frac{\Delta}{p} \right) = +1 \).

Part (3) of Proposition 6.12 says that \( \left( \frac{\Delta}{p} \right) \) is +1 for half of the columns of the character table and −1 for the other half. We proved this by showing that exactly half of the combinations of ±1’s in the columns of the table have \( \left( \frac{\Delta}{p} \right) = +1 \) for odd primes \( p \) in the congruence class corresponding to the column. Dirichlet’s Theorem implies that each of these congruence classes does actually contain at least one odd prime \( p \). We know that if \( \left( \frac{\Delta}{p} \right) = +1 \) then there exists a form of discriminant \( \Delta \) representing \( p \). Thus the combinations of ±1’s in the columns with \( \left( \frac{\Delta}{p} \right) = +1 \) correspond to actual genera, containing at least one form. On the other hand if \( \left( \frac{\Delta}{p} \right) = -1 \) for a column of the table then there can be no form realizing the combination of ±1’s in this column, so there is no genus of forms realizing this combination of ±1’s.

\( \Box \)

From this theorem we can deduce two very strong corollaries.

\textbf{Corollary 6.17.} \textit{The number of genera in discriminant \( \Delta \) is equal to the number of equivalence classes of primitive forms of discriminant \( \Delta \) that have mirror symmetry.}

This may seem a little surprising since there is no apparent connection between genera and mirror symmetry. A possible explanation might be that each genus contained exactly one equivalence class of primitive forms with mirror symmetry, but this is not always true. For example when \( \Delta = -56 \) we saw earlier in the chapter that there are two genera and two equivalence classes of mirror symmetric forms, but both these forms belong to the same genus. The true explanation will come in Chapter 7 when we study the class group.

\textit{Proof:} The number of equivalence classes of primitive forms with mirror symmetry was computed in Theorem 5.8 to be \( 2^{k-1} \) in most cases, where \( k \) is the number of distinct prime divisors of \( \Delta \). The exceptions are discriminants \( \Delta = 4(4m + 1) \) when \( 2^{k-1} \) is replaced by \( 2^{k-2} \), and \( \Delta = 32m \) when \( 2^{k-1} \) is replaced by \( 2^k \). In the nonexceptional cases we have \( k = \kappa \), the number of characters in discriminant \( \Delta \) since there is one character for each prime dividing \( \Delta \). When \( \Delta = 4(4m + 1) \) there is no character for the prime 2 so \( \kappa = k - 1 \), and when \( \Delta = 32m \) there are two characters for the prime 2 so \( \kappa = k + 1 \). The result follows.

\( \Box \)

\textbf{Corollary 6.18.} \textit{For a fixed discriminant \( \Delta \), each genus of primitive forms consists of a single equivalence class of forms if and only if all the topographs of primitive forms of discriminant \( \Delta \) have mirror symmetry.}
Proof: Let \( E(\Delta) \) be the set of equivalence classes of primitive forms of discriminant \( \Delta \) and let \( G(\Delta) \) be the set of genera of primitive forms of discriminant \( \Delta \). There is a natural function \( \Phi : E(\Delta) \to G(\Delta) \) assigning to each equivalence class of forms the genus of these forms. The function \( \Phi \) is onto since there is at least one form in each genus, by the definition of genus. If all primitive forms of discriminant \( \Delta \) have mirror symmetry then the previous corollary says that the sets \( E(\Delta) \) and \( G(\Delta) \) have the same number of elements. Then since \( \Phi \) is onto it must also be one-to-one. This means that each genus consists of a single equivalence class of forms.

Conversely, if each genus consists of a single equivalence class then \( \Phi \) is one-to-one. Since \( \Phi \) is also onto, this means it is a one-to-one correspondence so \( E(\Delta) \) and \( G(\Delta) \) have the same number of elements. By the preceding corollary this means that the equivalence classes of primitive forms with mirror symmetry account for all the elements of \( E(\Delta) \), and the proof is complete.

\( \square \)

4. Representing Non-primes

We have been focusing on determining which primes are represented in a given discriminant, and now we turn to the corresponding problem for non-primes. Recall that when we say that a form \( \mathit{Q} \) represents a number \( n \) we mean that \( n \) appears in the topograph of \( \mathit{Q} \), so \( \mathit{Q}(x, y) = n \) for some primitive pair \( (x, y) \). The primitivity of \( (x, y) \) is automatic when \( n \) is prime, or more generally a product of distinct primes, but otherwise it is a genuine restriction.

Our general criterion is that a number \( n \) is represented by at least one form of discriminant \( \Delta \) if and only if \( n \) factors as a product \( n = p_1^{e_1} p_2^{e_2} \cdots p_k^{e_k} \) of powers of distinct primes \( p_i \) each of which is represented by some form of discriminant \( \Delta \), where the exponents \( e_i \) of the primes \( p_i \) not dividing \( \Delta \) are unrestricted, while for primes \( p_i \) dividing \( \Delta \) but not the conductor we have \( e_i \leq 1 \). In particular this is all \( p_i \) dividing \( \Delta \) if \( \Delta \) is a fundamental discriminant.

The main result in this section will be an answer to this question. This will be in terms of the conductor of \( \Delta \), the positive number \( d \) such that \( \Delta = d^2 \Delta' \) for \( \Delta' \) a fundamental discriminant.

**Theorem 6.19.** A number \( n > 1 \) is represented by at least one form of discriminant \( \Delta \) exactly when \( n \) factors as a product \( n = p_1^{e_1} p_2^{e_2} \cdots p_k^{e_k} \) of powers of distinct primes \( p_i \) each of which is represented by some form of discriminant \( \Delta \), where the exponents \( e_i \) of the primes \( p_i \) not dividing \( \Delta \) are unrestricted, while for primes \( p_i \) dividing \( \Delta \) but not the conductor we have \( e_i \leq 1 \). In particular this is all \( p_i \) dividing \( \Delta \) if \( \Delta \) is a fundamental discriminant. The condition on the exponents \( e_i \) of primes \( p_i \) dividing the conductor is more complicated: First write \( \Delta = p_i^s q \) with \( p_i^s \) the highest power of \( p_i \) dividing \( \Delta \). Then if \( p_i \) is odd the condition on \( e_i \) is that either (a) \( e_i \leq s \) or (b) \( e_i \geq s + 1 \) and \( s \) is even and \( \left( \frac{\Delta}{p_i} \right) = +1 \). If \( p_i = 2 \) then the condition on \( e_i \) is that
either (a) \( e_i \leq s - 2 \) or (b) \( s \) is even and one of the following three possibilities holds:
(1) \( e_i = s - 1 \), (2) \( e_i = s \) and \( q = 4l + 1 \), or (3) \( e_i \geq s + 1 \) and \( q = 8l + 1 \).

This finishes the story for the representation problem in the cases of discriminants for which all forms are equivalent, such as those at the first level of complexity that we considered at the beginning of the chapter.

Some partial information about the representation problem for arbitrary discriminants \( \Delta \) can be obtained from the fact that all primitive forms representing a given number \( n \) coprime to \( \Delta \) belong to the same genus, since fixing \( n \) fixes the values of all characters. In Proposition 7.24 we will see that this holds more generally for numbers \( n \) coprime to the conductor of \( \Delta \), and in particular for all \( n \) if \( \Delta \) is a fundamental discriminant. An example where it fails to hold for a nonfundamental discriminant is \( \Delta = -99 \) with conductor 3. Here the two forms \( Q_1 = x^2 + xy + 25y^2 \) and \( Q_2 = 5x^2 + xy + 5y^2 \) both represent 27 and 81, as we saw in the discussion following the statement of Proposition 6.2, but \( Q_1 \) and \( Q_2 \) have different genera since the character \( \chi_3 \) has the value +1 on numbers in the topograph of \( Q_1 \) not divisible by 3, such as the number 1, while \( \chi_3 \) has the value -1 on numbers in the topograph of \( Q_2 \) not divisible by 3, such as 5.

Before proving the theorem let us look at some examples to illustrate the more complicated case of nonfundamental discriminants.

**Example:** \( \Delta = -12 \) with conductor 2. The two forms here are \( Q_1 = x^2 + 3y^2 \) and the nonprimitive form \( Q_2 = 2x^2 + xy + 2y^2 \). The primes represented in discriminant \(-12\) are 2, 3, and primes \( p \) with \( \left( \frac{-12}{p} \right) = \left( \frac{-3}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{3}{p} \right) = \left( \frac{p}{3} \right) = +1 \), so these are the primes \( p \equiv 1 \mod 3 \). The theorem says that the numbers represented in discriminant \(-12\) are the numbers \( n = 2^a3^bp_1 \cdots p_k \) with \( a \leq 2 \), \( b \leq 1 \), and each \( p_i \) a prime congruent to 1 mod 3. (When we apply the theorem for \( p_i = 2 \) we have \( s = 2 \) and \( q = -3 \).) We can in fact determine which of \( Q_1 \) and \( Q_2 \) is giving these representations. The form \( Q_2 \) is twice \( x^2 + xy + y^2 \) and we have already determined which numbers the latter form represents, namely the products \( 3^b p_1 \cdots p_k \) with \( b \leq 1 \) and each prime \( p_i \equiv 1 \mod 3 \). Thus, of the numbers represented by \( Q_1 \) or \( Q_2 \), the numbers represented by \( Q_2 \) are those with \( a = 1 \). None of these numbers with \( a = 1 \) are represented by \( Q_1 \) since \( x^2 + 3y^2 \) is never 2 mod 4, as \( x^2 \) and \( y^2 \) must be 0 or 1 mod 4.

**Example:** \( \Delta = -28 \) with conductor 2 again. Here the only two forms up to equivalence are \( Q_1 = x^2 + 7y^2 \) and \( Q_2 = 2x^2 + 2xy + 4y^2 \) which is not primitive.
The primes represented in discriminant \(-28\) are 2, 7, and odd primes \(p\) with \((\frac{-28}{p}) = +1\) so \(p \equiv 1, 2, 4 \mod 7\). According to the theorem the numbers represented by \(Q_1\) or \(Q_2\) are the numbers \(n = 2^a \tau_b p_1 \cdots p_k\) with \(b \leq 1\) and odd primes \(p_i \equiv 1, 2, 4 \mod 7\). There is no restriction on \(a\) since when we apply the theorem with \(p_i = 2\) we have \(s = 2\) and \(q = -7 = 81 + 1\). We can say exactly which numbers are represented by \(Q_2\) since it is twice the form \(x^2 + xy + 2y^2\) of discriminant \(-7\), which is a fundamental discriminant of class number 1 so the theorem tells us which numbers this form represents, namely the numbers \(2^a \tau_b p_1 \cdots p_k\) with \(a \geq 1\), \(b \leq 1\) and odd primes \(p_i \equiv 1, 2, 4 \mod 7\). Hence \(Q_1\) must represent at least the numbers \(2^a \tau_b p_1 \cdots p_k\) with \(a = 0\), \(b \leq 1\), and odd primes \(p_i \equiv 1, 2, 4 \mod 7\). These numbers are all odd since \(a = 0\), but \(Q_1\) also represents some even numbers since \(x^2 + 7y^2\) is even whenever both \(x\) and \(y\) are odd. From the topograph we might conjecture that \(Q_1\) represents exactly the numbers \(2^a \tau_b p_1 \cdots p_k\) with \(a = 0\) or \(a \geq 3\), and with \(b \leq 1\) and the same odd primes \(p_i \equiv 1, 2, 4 \mod 7\) as before. It is not difficult to exclude \(a = 1\) and \(a = 2\) by considering the values of \(x^2 + 7y^2\) mod 4 and mod 8. In Chapter 8 we will show that all the conjectured numbers with \(a \geq 3\) are in fact represented by \(Q_1\).

Now we turn to proving the theorem, using the familiar criterion that a number \(n\) is represented by at least one form of discriminant \(\Delta\) if and only if \(\Delta\) is a square mod \(4n\). As a first step we have:

**Lemma 6.20.** If a number \(x\) is a square mod \(m_1\) and is also a square mod \(m_2\) where \(m_1\) and \(m_2\) are coprime, then \(x\) is a square mod \(m_1 m_2\).

For example, the number 2 is a square mod 7 (since \(3^2 \equiv 2 \mod 7\)) and also mod 17 (since \(6^2 \equiv 2 \mod 17\)) so 2 must also be a square mod \(7 \cdot 17 = 119\). And in fact \(2 \equiv 11^2 \mod 119\).

**Proof:** This will follow from the Chinese Remainder Theorem. Suppose that a number \(x\) is a square mod \(m_1\) and also a square mod \(m_2\) where \(m_1\) and \(m_2\) are coprime. This means there are numbers \(a_1\) and \(a_2\) such that \(x \equiv a_1^2 \mod m_1\) and \(x \equiv a_2^2 \mod m_2\). By the Chinese Remainder Theorem there exists a number \(a\) that is congruent to \(a_1\)
mod $m_1$ and to $a_2$ mod $m_2$. Then $x \equiv a_1^2 \equiv a^2 \mod m_1$ and $x \equiv a_2^2 \equiv a^2 \mod m_2$. This implies $x \equiv a^2 \mod m_1 \cdot m_2$ since the difference $x - a^2$ is divisible by both $m_1$ and $m_2$ and hence by their product $m_1 \cdot m_2$ since we assume $m_1$ and $m_2$ are coprime. This shows that $x$ is a square mod $m_1 \cdot m_2$.

As a simple application of the preceding lemma we can deduce that if a discriminant $\Delta$ is a square modulo an odd number $n$ then it is a square mod $4n$ since discriminants are always squares mod 4, being either 0 or 1 mod 4. Earlier we proved this fact using a special argument, but now we see that it follows from a more general principle.

The preceding lemma reduces the problem of representation in a fixed discriminant to the case of representing prime powers. It remains to reduce further from prime powers to just primes themselves. For most primes this will be possible using the following result:

**Lemma 6.21.** If a number $x$ is a square mod $p$ for an odd prime $p$ not dividing $x$, then $x$ is also a square mod $p^r$ for each $r > 1$. The corresponding statement for the prime $p = 2$ is that if an odd number $x$ is a square mod 8 then $x$ is also a square mod $2^r$ for each $r > 3$.

For example, 2 is a square mod 7 since $2 \equiv 3^2 \mod 7$, so 2 is also a square mod $7^2$, namely $2 \equiv 10^2 \mod 49$. It is also a square mod $7^3 = 343$ since $2 \equiv 108^2 \mod 343$. Likewise it must be a square mod $7^4$, mod $7^5$, etc. The proof of the lemma will give a method for refining the initial congruence $2 \equiv 3^2 \mod 7$ to each subsequent congruence $2 \equiv 10^2 \mod 49$, $2 \equiv 108^2 \mod 343$, etc.

For the prime $p = 2$ we have to begin with squares mod 8 since 3 is a square mod 2 but not mod 4, while 5 is a square mod 4 but not mod 8.

**Proof of the Lemma:** We will show that if $x$ is a square mod $p^r$ then it is also a square mod $p^{r+1}$, assuming $r \geq 1$ in the case that $p$ is odd and $r \geq 3$ in the case $p = 2$. By induction this will prove the lemma.

We begin by assuming that $x$ is a square mod $p^r$, so there is a number $y$ such that $x \equiv y^2 \mod p^r$ or in other words $p^r$ divides $x - y^2$, say $x - y^2 = p^rl$ for some integer $l$. We seek a number $z$ such that $x \equiv z^2 \mod p^{r+1}$, so it is reasonable to look for a $z$ with $z \equiv y \mod p^r$, or in other words $z = y + kp^r$ for some $k$. Thus we want to choose $k$ so that $x \equiv (y + kp^r)^2 \mod p^{r+1}$. This means we want $p^{r+1}$ to divide the number

$$x - (y + kp^r)^2 = x - (y^2 + 2kp^r y + k^2 p^{2r})$$
$$= (x - y^2) - 2kp^r y - k^2 p^{2r}$$
$$= p^r l - 2kp^r y - k^2 p^{2r}$$
$$= p^r (l - 2ky - k^2 p^r)$$
For this to be divisible by \( p^{r+1} \) means that \( p \) should divide \( l - 2ky - k^2p^r \). Since we assume \( r \geq 1 \) this is equivalent to \( p \) dividing \( l - 2ky \), or in other words, \( l - 2ky = pq \) for some integer \( q \). Rewriting this as \( l = 2yk + pq \) we see that this linear Diophantine equation with unknowns \( k \) and \( q \) always has a solution when \( p \) is odd since \( 2y \) and \( p \) are coprime if \( p \) is odd, in view of the fact that \( p \) does not divide \( y \) since \( x \equiv y^2 \mod p^r \) and we assume \( x \) is not divisible by \( p \). This finishes the induction step in the case that \( p \) is odd.

When \( p = 2 \) this argument breaks down at the last step since the equation \( l = 2yk + pq \) becomes \( l = 2yk + 2q \) and this will not have a solution when \( l \) is odd. To modify the proof so that it works for \( p = 2 \) we would like to get rid of the factor 2 in the equation \( l = 2yk + pq \) which arose when we squared \( y + kp^r \). To do this, suppose that instead of trying \( z = y + k \cdot 2^r \) we try \( z = y + k \cdot 2^{r-1} \). Then we would want \( 2^{r+1} \) to divide

\[
x - (y + k \cdot 2^{r-1})^2 = (x - y^2) - k \cdot 2^r y - k^22^{2r-2}
\]

\[
= 2^r l - k \cdot 2^r y - k^22^{2r-2}
\]

\[
= 2^r (l - ky - k^22^{r-2})
\]

Assuming \( r \geq 3 \), this means 2 should divide \( l - ky \), or in other words \( l = yk + 2q \) for some integer \( q \). The number \( y \) is odd since \( y^2 \equiv x \mod 2^r \) and \( x \) is odd by assumption. This implies the equation \( l = yk + 2q \) has a solution \( (k, q) \). \( \square \)

**Corollary 6.22.** If a prime \( p \) not dividing the discriminant \( \Delta \) is represented by a form of discriminant \( \Delta \) then every power of \( p \) is also represented by a form of discriminant \( \Delta \).

**Proof:** First we consider odd primes \( p \). Since we assume \( p \) is represented in discriminant \( \Delta \) we know that \( \Delta \) is a square mod \( p \). The preceding lemma then says that \( \Delta \) is a square mod each power of \( p \), so all powers of \( p \) are represented in discriminant \( \Delta \). For \( p = 2 \) the argument is almost the same. In this case the representability of 2 implies that \( \Delta \) is a square mod \( 4p = 8 \) so the lemma implies that \( \Delta \) is also a square mod all higher powers of 2, so all powers of 2 are represented. \( \square \)

Now let \( p \) be a prime that does divide the discriminant \( \Delta \). In this case we know there is a form of discriminant \( \Delta \) representing \( p \), and it remains to determine whether powers of \( p \) are also represented. If \( p \) does not divide the conductor of \( \Delta \) then Proposition 6.1 says that no power of \( p \) is represented in discriminant \( \Delta \). This finishes the proof of the first part of the theorem, and in particular the case of fundamental discriminants. For primes dividing the conductor we will use the following result:

**Lemma 6.23.** For a given prime \( p \) suppose that a number \( x \) divisible by \( p \) factors as \( p^s q \) where \( p \) does not divide \( q \), so \( p^s \) is the largest power of \( p \) dividing \( x \). Then:

(a) \( x \) is a square mod \( p^r \) for each \( r \leq s \).
(b) If \( r > s \) and \( s \) is odd then \( x \) is not a square mod \( p^r \).

(c) If \( r > s \) and \( s \) is even then \( x \) is a square mod \( p^r \) if and only if \( q \) is a square mod \( p^{r-s} \).

**Proof:** Part (a) is easy since \( x \) is 0 mod \( p^s \) hence also mod \( p^r \) if \( r \leq s \), and 0 is always a square mod anything.

For (b) we assume \( r > s \) and \( s \) is odd. Suppose \( p^s q \) is a square mod \( p^r \), so \( p^s q = y^2 + lp^r \) for some integers \( y \) and \( l \). Then \( p^s \) divides \( y^2 + lp^r \) and it divides \( lp^r \) (since \( r > s \)) so \( p^s \) divides \( y^2 \). Since \( s \) is assumed to be odd and the exponent of \( p \) in \( y^2 \) must be even, this implies \( p^{s+1} \) divides \( y^2 \). It also divides \( lp^r \) since \( s + 1 \leq r \), so from the equation \( p^s q = y^2 + lp^r \) we conclude that \( p \) divides \( q \), contrary to definition of \( q \). This contradiction shows that \( p^s q \) is not a square mod \( p^r \) when \( r > s \) and \( s \) is odd, so statement (b) is proved.

For (c) we assume \( r > s \) and \( s \) is even. As in part (b), if \( p^s q \) is a square mod \( p^r \) we have an equation \( p^s q = y^2 + lp^r \) and this implies that \( p^s \) divides \( y^2 \). Since \( s \) is now even, this means \( y^2 = p^s z^2 \) for some number \( z \). Canceling \( p^s \) from \( p^s q = y^2 + lp^r \) yields an equation \( q = z^2 + lp^{r-s} \), which says that \( q \) is a square mod \( p^{r-s} \). Conversely, if \( q \) is a square mod \( p^{r-s} \) we have an equation \( q = z^2 + lp^{r-s} \) and hence \( p^s q = p^s z^2 + lp^r \). Since \( s \) is even, this says that \( p^s q \) is a square mod \( p^r \). \( \Box \)

Now we can finish the proof of the last part of Theorem 6.19, which for convenience we restate as the following corollary to the preceding lemma.

**Corollary 6.24.** Let \( p \) be a prime dividing the discriminant \( \Delta \), and let us write \( \Delta = p^s q \) with \( p^s \) the highest power of \( p \) dividing \( \Delta \). If \( p \) is odd then \( p^r \) is represented in discriminant \( \Delta \) if and only if either (a) \( r \leq s \) or (b) \( s \) is even and \( \left( \frac{q}{p} \right) = +1 \). If \( p = 2 \) then \( 2^r \) is represented in discriminant \( \Delta \) if and only if either (a) \( r \leq s - 2 \) or (b) \( s \) is even and one of the following three possibilities holds: \( r = s - 1 \); \( r = s \) and \( q = 4k + 1 \); or \( r \geq s + 1 \) and \( q = 8k + 1 \).

**Proof:** When \( p \) is odd this follows immediately from the lemma by taking \( x = \Delta \), using the fact that \( q \) is a square mod powers of \( p \) exactly when it is a square mod \( p \).

When \( p = 2 \) we need to determine when \( \Delta \) is a square mod \( 4 \cdot 2^r = 2^{r+2} \). By the lemma this happens only when \( r \leq s - 2 \) or when \( s \) is even and \( q \) (which is odd) is a square mod \( 2^{r+2-s} \). When \( r + 2 - s = 1 \), so \( r = s - 1 \), every \( q \) is a square mod \( 2^{r+2-s} = 2 \). When \( r + 2 - s = 2 \), so \( r = s \), \( q \) is a square mod \( 2^{r+2-s} = 4 \) only when \( q = 4k + 1 \). And when \( r + 2 - s \geq 3 \), so \( r \geq s + 1 \), \( q \) is a square mod \( 2^{r+2-s} \) only when \( q = 8k + 1 \). \( \Box \)
5. Proof of Quadratic Reciprocity

First let us show that quadratic reciprocity can be expressed more concisely as a single formula

\[
\left( \frac{p}{q} \right) \left( \frac{q}{p} \right) = (-1)^{\frac{p-1}{2} \cdot \frac{q-1}{2}}
\]

(\ast)

Here \( p \) and \( q \) are distinct odd primes. Since they are odd, the fractions \( \frac{p-1}{2} \) and \( \frac{q-1}{2} \) are integers. The only way the exponent \( \frac{p-1}{2} \cdot \frac{q-1}{2} \) can be odd is for both factors to be odd, so \( \frac{p-1}{2} = 2k + 1 \) and \( \frac{q-1}{2} = 2l + 1 \), which is equivalent to saying \( p = 4k + 3 \) and \( q = 4l + 3 \). Thus the only time that the right side of the formula (\ast\) can be \(-1\) is when \( p \) and \( q \) are both congruent to 3 mod 4, and quadratic reciprocity is the assertion that the left side of (\ast\) has exactly this property.

There will be three main steps in the proof of quadratic reciprocity. The first is to derive an explicit algebraic formula for \( \left( \frac{a}{p} \right) \) due originally to Euler. The second step is to use this formula to give a somewhat more geometric interpretation of \( \left( \frac{a}{p} \right) \) in terms of the number of dots in a certain triangular pattern. Then the third step is the actual proof of quadratic reciprocity using symmetry properties of the patterns of dots. This proof is due to Eisenstein, first published in 1844, simplifying an earlier proof by Gauss who was the first to give a full proof of quadratic reciprocity.

**Step 1.** In what follows we will always use \( p \) to denote an odd prime, and the symbol \( a \) will always denote an arbitrary nonzero integer not divisible by \( p \). When we write a congruence such as \( a \equiv b \) this will always mean congruence mod \( p \), even if we do not explicitly say mod \( p \).

Euler's formula is:

\[
\left( \frac{a}{p} \right) \equiv a^{\frac{p-1}{2}} \mod p
\]

For example, for \( p = 11 \) Euler's formula says \( \left( \frac{2}{11} \right) = 2^5 = 32 \equiv -1 \mod 11 \) and \( \left( \frac{3}{11} \right) = 3^5 = 243 \equiv +1 \mod 11 \). These are the correct values since the squares mod 11 are \((\pm 1)^2 = 1, (\pm 2)^2 = 4, (\pm 3)^2 = 9, (\pm 4)^2 \equiv 5, \) and \((\pm 5)^2 \equiv 3\).

Note that Euler's formula determines the value of \( \left( \frac{a}{p} \right) \) uniquely since \(+1\) and \(-1\) are not congruent mod \( p \) since \( p > 2 \). It is not immediately obvious that the number \( a^{\frac{p-1}{2}} \) should always be congruent to either \(+1\) or \(-1\) mod \( p \), but when we prove Euler's formula we will see that this has to be true.

As a special case, taking \( a = -1 \) in Euler's formula gives the calculation

\[
\left( \frac{-1}{p} \right) = (-1)^{\frac{p-1}{2}} = \begin{cases} 
+1 & \text{if } p = 4k + 1 \\
-1 & \text{if } p = 4k + 3
\end{cases}
\]

Before proving Euler's formula we will need to derive a few preliminary general facts about congruences modulo a prime \( p \). The first fact is that each of the numbers \( a = 1, 2, \ldots, p-1 \) has a multiplicative inverse mod \( p \). To see why this is true, notice that each such \( a \) is coprime to \( p \), so we know from Chapter 2 that the equation
$ax + py = 1$ has an integer solution $(x, y)$. This equation can be rewritten as the congruence $ax \equiv 1 \text{ mod } p$, which says that $x$ is an inverse for $a \text{ mod } p$. Note that any two choices for $x$ here are congruent mod $p$ since if $ax \equiv 1$ and $ax' \equiv 1$ then multiplying both sides of $ax' \equiv 1$ by $x$ gives $xax' \equiv x$, and $xa \equiv 1$ so we conclude that $x \equiv x'$.

Which numbers equal their own inverse mod $p$? If $a \cdot a \equiv 1$, then we can rewrite this as $a^2 - 1 \equiv 0$, or in other words $(a + 1)(a - 1) \equiv 0$. This is certainly a valid congruence if $a \equiv \pm 1$, so suppose that $a \not\equiv \pm 1$. The factor $a + 1$ is then not congruent to 0 mod $p$ so it has a multiplicative inverse mod $p$, and if we multiply the congruence $(a + 1)(a - 1) \equiv 0$ by this inverse, we get $a - 1 \equiv 0$ so $a \equiv 1$, contradicting the assumption that $a \not\equiv \pm 1$. This argument shows that the only numbers among $1, 2, \cdots, p - 1$ that are congruent to their inverses mod $p$ are 1 and $p - 1$.

An application of this fact is a result known as Wilson’s Theorem:

\[(p - 1)! \equiv -1 \text{ modulo } p \text{ whenever } p \text{ is prime.}\]

To see why this is true, observe that in the product $(p - 1)! = (1)(2) \cdots (p - 1)$ each factor other than 1 and $p - 1$ can be paired up with its multiplicative inverse mod $p$ and these two terms multiply together to give 1 mod $p$, so the whole product is congruent to just $(1)(p - 1)$ mod $p$. Since $p - 1 \equiv -1$ mod $p$ this gives Wilson’s Theorem.

Now let us prove the following congruence known as Fermat’s Little Theorem:

\[a^{p-1} \equiv 1 \text{ mod } p \text{ whenever } p \text{ is an odd prime not dividing } a.\]

To see this, note first that the numbers $a, 2a, 3a, \cdots, (p - 1)a$ are all distinct mod $p$ since we know that $a$ has a multiplicative inverse mod $p$, so in a congruence $ma \equiv na$ we can multiply both sides by the inverse of $a$ to deduce that $m \equiv n$. Let us call this property that $ma \equiv na$ implies $m \equiv n$ the cancellation property for congruences mod $p$.

Thus the set \{a, 2a, 3a, \cdots, (p - 1)a\} is the same mod $p$ as \{1, 2, 3, \cdots, p - 1\} since both sets have $p - 1$ elements and neither set contains numbers that are 0 mod $p$. If we take the product of all the numbers in each of these two sets we obtain the congruence

\[(a)(2a)(3a) \cdots (p - 1)a \equiv (1)(2)(3) \cdots (p - 1) \text{ mod } p\]

We can cancel the factors $2, 3, \cdots, p - 1$ from both sides by repeated applications of the cancellation property. The result is the congruence $a^{p-1} \equiv 1$ claimed by Fermat’s Little Theorem.

Now we can prove Euler’s formula for $\left(\frac{a}{p}\right)$. The first case is that $\left(\frac{a}{p}\right) = 1$, so $a$ is a square mod $p$ and $a \equiv x^2$ for some $x \neq 0$. In this case we have $a^{p-1} \equiv x^{p-1} \equiv 1$ by Fermat’s Little Theorem. So in this case Euler’s formula $\left(\frac{a}{p}\right) \equiv a^{\frac{p-1}{2}}$ is valid, both sides being $+1$.  


The other case is that \( \left( \frac{a}{p} \right) = -1 \) so \( a \) is not a square mod \( p \). Observe first that the congruence \( xy \equiv a \) has a solution \( y \) mod \( p \) for each \( x \neq 0 \) since \( x \) has an inverse \( x^{-1} \) mod \( p \) so we can take \( y = x^{-1}a \). Moreover the solution \( y \) is unique mod \( p \) since \( xy_1 \equiv xy_2 \) implies \( y_1 \equiv y_2 \) by the cancellation property. Since we are in the case that \( a \) is not a square mod \( p \) the solution \( y \) of \( xy \equiv a \) satisfies \( y \neq x \). Thus the numbers \( 1, 2, 3, \ldots, p - 1 \) are divided up into \( \frac{p-1}{2} \) pairs \( \{x_1, y_1\}, \{x_2, y_2\}, \ldots, \{x_{\frac{p-1}{2}}, y_{\frac{p-1}{2}}\} \) with \( x_iy_i \equiv a \) for each \( i \). Multiplying all these \( \frac{p-1}{2} \) pairs together, we get

\[
a^{\frac{p-1}{2}} \equiv x_1y_1x_2y_2 \cdots x_{\frac{p-1}{2}}y_{\frac{p-1}{2}}
\]

The product on the right is just a rearrangement of \( (1)(2)(3) \cdots (p-1) \), and Wilson's Theorem says that this product is congruent to \( -1 \) mod \( p \). Thus we see that Euler's formula \( \left( \frac{a}{p} \right) \equiv a^{\frac{p-1}{2}} \) holds also when \( \left( \frac{a}{p} \right) = -1 \), completing the proof in both cases.

A consequence of Euler's formula is the multiplicative property of Legendre symbols that we stated and used earlier in the chapter:

\[
\left( \frac{ab}{p} \right) = \left( \frac{a}{p} \right) \left( \frac{b}{p} \right)
\]

This holds since \( (ab)^{\frac{p-1}{2}} = a^{\frac{p-1}{2}}b^{\frac{p-1}{2}} \).

**Step 2.** Here our aim is to express the Legendre symbol \( \left( \frac{a}{p} \right) \) in more geometric terms. To begin, consider a rectangle in the first quadrant of the \( xy \)-plane that is \( p \) units wide and \( a \) units high, with one corner at the origin and the opposite corner at the point \( (p,a) \). For example for \( p = 7 \) and \( a = 5 \) we have the picture

![Diagram of a rectangle with lattice points](image)

We will be interested in points that lie strictly in the interior of the rectangle and whose coordinates are integers. Points satisfying the latter condition are called *lattice points*. The number of lattice points in the interior is then \( (p-1)(a-1) \) since their \( x \)-coordinates can range from 1 to \( p - 1 \) and their \( y \)-coordinates from 1 to \( a - 1 \), independently.

The diagonal of the rectangle from \( (0,0) \) to \( (p,a) \) does not pass through any of these interior lattice points since we assume that the prime \( p \) does not divide \( a \), so
the fraction $a/p$, which is the slope of the diagonal, is in lowest terms. (If there were an interior lattice point on the diagonal, the slope of the diagonal would be a fraction with numerator and denominator smaller than $a$ and $p$.) Since there are no interior lattice points on the diagonal, exactly half of the lattice points inside the rectangle lie on each side of the diagonal, so the number of lattice points below the diagonal is $\frac{1}{2}(p - 1)(a - 1)$. This is an integer since $p$ is odd, which makes $p - 1$ even.

A more refined question one can ask is how many lattice points below the diagonal have even $x$-coordinate and how many have odd $x$-coordinate. Here there is no guarantee that these two numbers must be equal, and indeed if they were equal then both numbers would have to be $\frac{1}{4}(p - 1)(a - 1)$ but this fraction need not be an integer, for example when $p = 7$ and $a = 4$.

We denote the number of lattice points that are below the diagonal and have even $x$-coordinate by the letter $e$. Here is a figure showing the values of $e$ when $p = 7$ and $a$ ranges from 1 to 6:

<table>
<thead>
<tr>
<th>$a$</th>
<th>$e$</th>
<th>$\left(\frac{a}{\sqrt{7}}\right)$</th>
</tr>
</thead>
<tbody>
<tr>
<td>1</td>
<td>0</td>
<td>1</td>
</tr>
<tr>
<td>2</td>
<td>2</td>
<td>1</td>
</tr>
<tr>
<td>3</td>
<td>3</td>
<td>-1</td>
</tr>
<tr>
<td>4</td>
<td>6</td>
<td>+1</td>
</tr>
<tr>
<td>5</td>
<td>7</td>
<td>-1</td>
</tr>
<tr>
<td>6</td>
<td>9</td>
<td>-1</td>
</tr>
</tbody>
</table>
A slightly more complicated example is when $p = 13$ and $a$ goes from 1 to 12:

$$\begin{array}{|c|c|c|}
\hline
a & e & \left( \frac{a}{13} \right) \\
\hline
12 & 36 & +1 \\
11 & 33 & -1 \\
10 & 30 & +1 \\
9 & 26 & +1 \\
8 & 23 & -1 \\
7 & 21 & -1 \\
6 & 15 & -1 \\
5 & 13 & -1 \\
4 & 10 & +1 \\
3 & 6 & +1 \\
2 & 3 & -1 \\
1 & 0 & +1 \\
\hline
\end{array}$$

The way that $e$ varies with $a$ seems somewhat unpredictable. What we will show is that just knowing the parity of $e$ is already enough to determine the value of the Legendre symbol via the formula

$$\left( \frac{a}{p} \right) = (-1)^e$$

To prove this we first derive a formula for $e$. The segment of the vertical line $x = u$ going from the $x$-axis up to the diagonal has length $ua/p$ since the slope of the diagonal is $a/p$. If $u$ is a positive integer the number of lattice points on this line segment is $\left\lfloor \frac{ua}{p} \right\rfloor$, the greatest integer $n \leq \frac{ua}{p}$. Now if we add up these numbers of lattice points for $u$ running through the set of even numbers $E = \{2, 4, \cdots, p - 1\}$ we get

$$e = \sum_{E} \left\lfloor \frac{ua}{p} \right\rfloor$$

The way to compute $\left\lfloor \frac{ua}{p} \right\rfloor$ is to apply the division algorithm for integers, dividing $p$ into $ua$ to obtain $\left\lfloor \frac{ua}{p} \right\rfloor$ as the quotient with a remainder that we denote $r(u)$. Thus we have the formula

$$ua = p\left\lfloor \frac{ua}{p} \right\rfloor + r(u) \quad (1)$$

This formula implies that the number $\left\lfloor \frac{ua}{p} \right\rfloor$ has the same parity as $r(u)$ since $u$ is even and $p$ is odd. This relation between parities implies that the number $(-1)^e$ that we are interested in can also be computed as

$$(-1)^e = (-1)\sum_{E} \left\lfloor \frac{ua}{p} \right\rfloor = (-1)\sum_{E} r(u) \quad (2)$$
With this last expression in mind we will focus our attention on the remainders \( r(u) \).

The number \( r(u) \) lies strictly between 0 and \( p \) and can be either even or odd, but in both cases we can say that \((-1)^{r(u)} r(u)\) is congruent to an even number in the interval \((0, p)\) since if \( r(u) \) is odd, so is \((-1)^{r(u)} r(u)\) and then adding \( p \) to this gives an even number between 0 and \( p \). Thus there is always an even number \( s(u) \) between 1 and \( p \) that is congruent to \((-1)^{r(u)} r(u) \mod p\). Obviously \( s(u) \) is unique since no two numbers in the interval \((0, p)\) are congruent \( \mod p \).

A key fact about these even numbers \( s(u) \) is that they are all distinct as \( u \) varies over the set \( E \). For suppose we have \( s(u) = s(v) \) for another even number \( v \) in \( E \). Thus \( r(u) \equiv \pm r(v) \mod p \), which implies \( au \equiv \pm av \mod p \) in view of the equation (1) above. We can cancel the \( a \) from both sides of this congruence to get \( u \equiv \pm v \). However we cannot have \( u \equiv -v \) because the number between 0 and \( p \) that is congruent to \( -v \) is \( p - v \), so we would have \( u = p - v \) which is impossible since \( u \) and \( v \) are even while \( p \) is odd. Thus we must have \( u \equiv +v \), hence \( u = v \) since these are numbers strictly between 0 and \( p \). This shows that the numbers \( s(u) \) are all distinct.

Now consider the product of all the numbers \((-1)^{r(u)} r(u)\) as \( u \) ranges over the set \( E \). Written out, this is
\[
\left[ (-1)^{r(2)} r(2) \right] \left[ (-1)^{r(4)} r(4) \right] \cdots \left[ (-1)^{r(p-1)} r(p-1) \right]
\]
By equation (1) we have \( r(u) \equiv u a \mod p \), so this product is congruent \( \mod p \) to
\[
\left[ (-1)^{r(2)} 2a \right] \left[ (-1)^{r(4)} 4a \right] \cdots \left[ (-1)^{r(p-1)} (p-1)a \right]
\]
On the other hand, by the definition of the numbers \( s(u) \) the product (3) is congruent \( \mod p \) to
\[
[s(2)] [s(4)] \cdots [s(p-1)]]
\]
There are \( \frac{p-1}{2} \) factors here and they are all distinct even numbers in the interval \((0, p)\) as we showed in the previous paragraph, so they are just a rearrangement of the numbers 2, 4, \( \cdots \), \( p-1 \). Thus we have the congruence
\[
\left[ (-1)^{r(2)} 2a \right] \left[ (-1)^{r(4)} 4a \right] \cdots \left[ (-1)^{r(p-1)} (p-1)a \right] \equiv (2)(4) \cdots (p-1) \mod p
\]
We can cancel the factors 2, 4, \( \cdots \), \( p-1 \) from both sides of this congruence to obtain
\[
(-1)^{\sum_{E} r(u)} a^{\frac{p-1}{2}} \equiv 1 \mod p
\]
Both the factors \((-1)^{\sum_{E} r(u)}\) and \( a^{\frac{p-1}{2}} \) are \( \pm 1 \mod p \) and their product is 1 so they must be equal \( \mod p \) (using the fact that 1 and \(-1\) are not congruent modulo an odd prime). By Euler's formula we have \( a^{\frac{p-1}{2}} \equiv \left( \frac{a}{p} \right) \mod p \), so from the earlier formula (2) we conclude that \( \left( \frac{a}{p} \right) = (-1)^{e} \). This finishes Step 2 in the proof of quadratic reciprocity.

**Step 3.** Now we specialize the value of \( a \) to be an odd prime \( q \) distinct from \( p \). As in Step 2 we consider a \( p \times q \) rectangle.
We know that \( \left( \frac{a}{p} \right) = (-1)^e \) where \( e \) is the number of lattice points with even \( x \)-coordinate inside the rectangle and below the diagonal. Suppose that we divide the rectangle into two equal halves separated by the vertical line \( x = \frac{p}{2} \). This line does not pass through any lattice points since \( p \) is odd. This vertical line cuts off two smaller triangles from the two large triangles above and below the diagonal of the rectangle. Call the lower small triangle \( L \) and the upper one \( U \), and let \( l \) and \( u \) denote the number of lattice points with even \( x \)-coordinate in \( L \) and \( U \) respectively. We note that \( u \) has the same parity as the number of lattice points with even \( x \)-coordinate in the quadrilateral below \( U \) in the right half of the rectangle since each column of lattice points in the rectangle has \( q-1 \) points, an even number. Thus \( e \) has the same parity as \( l + u \), hence \( (-1)^e = (-1)^{l+u} \).

The next thing to notice is that rotating the triangle \( U \) by 180 degrees about the center of the rectangle carries it onto the triangle \( L \). This rotation takes the lattice points in \( U \) with even \( x \)-coordinate onto the lattice points in \( L \) with odd \( x \)-coordinate. Thus we obtain the formula \( \left( \frac{a}{p} \right) = (-1)^t \) where \( t \) is the total number of lattice points in the triangle \( L \).

Reversing the roles of \( p \) and \( q \), we can also say that \( \left( \frac{p}{q} \right) = (-1)^{t'} \) where \( t' \) is the number of lattice points in the triangle \( L' \) above the diagonal and below the horizontal line \( y = \frac{q}{2} \) bisecting the rectangle. Then \( t + t' \) is the number of lattice points in the small rectangle formed by \( L \) and \( L' \) together. This number is just \( \frac{p-1}{2} \cdot \frac{q-1}{2} \). Thus we have

\[
\left( \frac{a}{p} \right) \left( \frac{p}{q} \right) = (-1)^t (-1)^{t'} = (-1)^{t+t'} = (-1)^{\frac{p-1}{2} \cdot \frac{q-1}{2}}
\]

which finally finishes the proof of quadratic reciprocity.

We can also use the geometric interpretation of \( \left( \frac{a}{p} \right) \) to prove the formula for \( \left( \frac{2}{p} \right) \).
that was stated earlier in this chapter, namely

\[
\left( \frac{2}{p} \right) = \begin{cases} 
+1 & \text{if } p = 8k \pm 1 \\
-1 & \text{if } p = 8k \pm 3
\end{cases}
\]

We have shown that \( \left( \frac{2}{p} \right) = (-1)^e \) where \( e \) is the number of lattice points inside a \( p \times 2 \) rectangle lying below the diagonal and having even \( x \) coordinate, as indicated in the following figure which shows the diagonals for \( p = 3, 5, 7, \ldots, 17 \):

Another way to describe \( e \) is to say that it is equal to the number of even integers in the interval from \( p/2 \) to \( p \). We do not need to assume that \( p \) is prime in order to count these points below the diagonals, just that \( p \) is odd. One can see what the pattern is just by looking at the figure: Each time \( p \) increases by 2 there is one more even number at the right end of the interval \( (p/2, p) \), and there may or may not be one fewer even number at the left end of the interval, depending on whether \( p \) is increasing from \( 4k - 1 \) to \( 4k + 1 \) or from \( 4k + 1 \) to \( 4k + 3 \). It follows that the parity of \( e \) depends only on the value of \( p \mod 8 \) as in the table for \( p \leq 17 \), so \( e \) is even for \( p \equiv \pm 1 \mod 8 \) and \( e \) is odd for \( p \equiv \pm 3 \mod 8 \).
Exercises

1. For the form $Q(x, y) = x^2 + xy - y^2$ do the following things:
   (a) Draw enough of the topograph to show all the values less than 100 that occur in the topograph. This form is hyperbolic and it takes the same negative values as positive values, so you need not draw all the negative values.
   (b) Make a list of the primes less than 100 that occur in the topograph, and a list of the primes less than 100 that do not occur.
   (c) Characterize the primes in the two lists in part (b) in terms of congruence classes modulo $|\Delta|$ where $\Delta$ is the discriminant of $Q$.
   (d) Characterize the nonprime values in the topograph in terms of their factorizations into primes in the lists in part (b).
   (e) Summarize the previous parts by giving a simple criterion for determining the numbers $n$ such that $Q(x, y) = n$ has an integer solution $(x, y)$, primitive or not. The criterion should say something like $Q(x, y) = n$ is solvable if and only if $n = m^2 p_1 \cdots p_k$ where each $p_i$ is a prime such that ...
   (e) Check that all forms having the same discriminant as $Q$ are equivalent to $Q$.

2. Do the same things for the form $x^2 + xy + 2y^2$, except that this time you only need to consider values less than 50 instead of 100.

3. For discriminant $\Delta = -24$ do the following:
   (a) Verify that the class number is 2 and find two quadratic forms $Q_1$ and $Q_2$ of discriminant $-24$ that are not equivalent.
   (b) Draw topographs for $Q_1$ and $Q_2$ showing all values less than 100. (You don’t have to repeat parts of the topographs that are symmetric.)
   (c) Divide the primes less than 100 into three lists: those represented by $Q_1$, those represented by $Q_2$, and those represented by neither $Q_1$ nor $Q_2$. (No primes are represented by both $Q_1$ and $Q_2$.)
   (d) Characterize the primes in the three lists in part (c) in terms of congruence classes modulo $|\Delta| = 24$.
   (e) Characterize the nonprime values in the topograph of $Q_1$ in terms of their factorizations into primes in the lists in part (c), and then do the same thing for $Q_2$. Your answers should be in terms of whether there are an even or an odd number of prime factors from certain of the lists.
   (f) Summarize the previous parts by giving a criterion for which numbers $n$ the equation $Q_1(x, y) = n$ has an integer solution and likewise for the equation $Q_2(x, y) = n$. 

4. This problem will show how things can be more complicated than in the previous problems.
(a) Show that the number of equivalence classes of forms of discriminant $-23$ is 2 while the number of proper equivalence classes is 3, and find reduced forms $Q_1$ and $Q_2$ of discriminant $-23$ that are not equivalent.
(b) Draw the topographs of $Q_1$ and $Q_2$ up to the value 70. (Again you don’t have to repeat symmetric parts.)
(c) Find a number $n$ that occurs in both topographs, and find the $x$ and $y$ values that give $Q_1(x_1, y_1) = n = Q_2(x_2, y_2)$. (This sort of thing never happens in the previous problems.)
(d) Find a prime $p_1$ in the topograph of $Q_1$ and a different prime $p_2$ in the topograph of $Q_2$ such that $p_1$ and $p_2$ are congruent modulo $|\Delta| = 23$. (This sort of thing also never happens in the previous problems.)

5. As a sort of converse to Wilson’s theorem, show that if $n$ is not a prime then $(n - 1)!$ is not congruent to $-1$ mod $n$. More precisely, when $n > 4$ and $n$ is not prime, show that $n$ divides $(n - 1)!$, so $(n - 1)! \equiv 0$ mod $n$. What happens when $n = 4$?

6. Determine the values of $\Delta$ for which there exists a quadratic form of discriminant $\Delta$ that represents 5, and also determine the discriminants $\Delta$ for which there does not exist a form representing 5.

7. Verify that the statement of quadratic reciprocity is true for the following pairs of primes $(p, q)$: (3, 5), (3, 7), (3, 13), (5, 13), (7, 11), and (13, 17).

8. (a) There is an example near the end of this chapter that works out which primes are represented by some form of discriminant 13, using quadratic reciprocity for the key step. Do the same thing for discriminant 17.
(b) Show that all forms of discriminant 17 are equivalent to the principal form $x^2 + xy - 4y^2$.
(c) Draw enough of the topograph of $x^2 + xy - 4y^2$ to show all values between $-70$ and 70, and verify that the primes that occur are precisely the ones predicted by your answer in part (a).

9. Using quadratic reciprocity as in part (a) of the previous problem, figure out which primes are represented by at least one form of discriminant $\Delta$ for the following values of $\Delta$: $-3, 8, -20, 21$.

10. Consider a discriminant $\Delta = q^2$, $q > 0$, corresponding to 0-hyperbolic forms. Using the description of the topographs of such forms obtained in the previous chapter, show:
(a) Every number is represented by at least one form of discriminant $\Delta$, so in particular all primes are represented.
(b) The primes represented by a given form of discriminant $\Delta$ are exactly the primes
in certain congruence classes mod $q$ (and hence also mod $\Delta$).

(c) For each of the values $q = 1$, $2$, $7$, and $15$ determine the class number for discriminant $\Delta = q^2$ and find which primes are represented by the forms in each equivalence class.

11. Show that the calculation of the Legendre symbol $\left( \frac{-1}{p} \right)$ can also be obtained using the method in the proof of quadratic reciprocity involving counting certain lattice points in a $(p - 1) \times p$ rectangle.

12. Consider forms $Q = ax^2 + bxy + cy^2$ of discriminant $\Delta$. Show that the following three conditions are equivalent:
   
   (1) The coefficients $a$, $b$, and $c$ of $Q$ are all odd.
   
   (2) $Q$ represents only odd numbers.
   
   (3) $\Delta \equiv 5 \mod 8$.

13. For which fundamental discriminants $\Delta$ is there a form of discriminant $\Delta$ representing $|\Delta|$? What about nonfundamental discriminants?

14. Show there are infinitely many primes of the form $6m - 1$ by an argument similar to the one used for $4m - 1$.

15. In terms of their prime factorizations, which numbers are sums of two nonzero squares? Which squares are sums two nonzero squares?
Chapter 7. The Class Group for Quadratic Forms

In the previous chapter we obtained an answer to the question of which numbers \( n \) are represented by at least one form of a given determinant, where by “represent” we mean “appear in the topograph”, so we consider only the values \( Q(x, y) \) for primitive pairs \((x, y)\). The answer was in terms of certain congruence conditions on the prime divisors of \( n \). We could also pin down the genus of the forms representing \( n \) via congruence conditions. What one would really like to do is refine these results to determine which equivalence classes of forms represent \( n \), and for this it is natural to consider only primitive forms. Determining which primes each primitive form represents is quite a difficult and subtle problem about which much is known, but it requires considerably deeper mathematics than we can cover in this book so will say nothing more about the case of primes. What we will do in the present chapter is study the question for nonprimes, assuming one already knows the answer for primes. For fundamental discriminants we will obtain a fairly complete picture, while for nonfundamental discriminants there will remain certain ambiguities, with examples showing the extra complication in these cases.

The main tool will be a method for multiplying forms of a given discriminant that corresponds to multiplying the numbers represented by these forms. This multiplication of forms gives rise to a commutative group structure on the set of proper equivalence classes of primitive forms of a given discriminant, called the class group. This group structure has other uses too. For example we will use it to give a good explanation for why the number of genera in a given discriminant is equal to the number of equivalence classes of primitive forms in that discriminant whose topographs have mirror symmetry.

1. Multiplication of Forms

Since we will often be dealing with several different forms at a time it will be very helpful to shorten the notation by writing a form \( ax^2 + bxy + cy^2 \) simply as \([a, b, c]\), retaining only the essential information of the coefficients.

Recall that a number \( a \) is represented by a form \( Q \) if and only if \( a \) appears in the topograph of \( Q \), and this in turn is equivalent to \( a \) appearing as the leading coefficient of a form \([a, b, c]\) equivalent to \( Q \). A simple observation is that if \( a \) factors as \( a = a_1a_2 \) then the forms \([a_1a_2, b, c]\), \([a_1, b, a_2c]\), and \([a_2, b, a_1c]\) all have the same discriminant. This gives an explicit demonstration of the fact that if a number \( a \) is represented in discriminant \( \Delta \) then so is any divisor of \( a \), which we proved in a different way in the previous chapter.

One might wonder about the reverse process of combining or “multiplying” the two forms \([a_1, b, a_2c]\) and \([a_2, b, a_1c]\) to obtain the form \([a_1a_2, b, c]\). More generally one might ask whether it is possible to multiply two forms \([a_1, b_1, c_1]\) and \([a_2, b_2, c_2]\) of discriminant \( \Delta \) to obtain another form \([a_1a_2, b, c]\) of discriminant \( \Delta \), for certain
coefficients $b$ and $c$. However, there have to be some restrictions on $a_1$ and $a_2$, otherwise the product of any two numbers represented in discriminant $\Delta$ would also be represented in discriminant $\Delta$, which is not always the case.

Pursuing this idea, note first that since the discriminant $\Delta = b^2 - 4ac$ is fixed, the coefficient $c$ in the hypothetical product form $[a_1 a_2, b, c]$ is determined by $a = a_1 a_2$ and $b$. (We will always assume $\Delta$ is not a square so $a$ and $c$ will always be nonzero.) So the question becomes how to choose $b$. As a special case, suppose that $b_1 = b_2$ and we choose $b = b_1 = b_2$. Then the discriminant equations $b_1^2 - 4a_1 c_1 = b_2^2 - 4a_2 c_2 = b^2 - 4a_1 a_2 c$ simplify to $a_1 c_1 = a_2 c_2 = a_1 a_2 c$. These will be satisfied if $c_1 = a_2 c$ and $c_2 = a_1 c$. Thus we arrive at the earlier situation of multiplying the forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ to obtain $[a_1 a_2, b, c]$. For example the product of $[2,0,15]$ and $[3,0,10]$ would be $[6,0,5]$.

A pair of forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ is said to be concordant. Note that two forms $[a_1, b, c_1]$ and $[a_2, b, c_2]$ of the same discriminant are concordant if $a_1$ divides $c_2$ since this implies $c_2 = a_1 c$ for some integer $c$, hence $a_1 c_1 = a_2 c_2 = a_2 a_1 c$ so we can cancel the factor $a_1$ to get $c_1 = a_2 c$ since we are only dealing with forms whose outer coefficients are nonzero.

Our goal is to show that this scheme of defining the product of two concordant forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ to be $[a_1 a_2, b, c]$ actually works to give a group structure in $\text{CG}(\Delta)$, the set of proper equivalence classes of primitive forms of discriminant $\Delta$.

Since we wish to consider only primitive forms the following result will be useful:

**Lemma 7.1.** If the concordant forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ are primitive then so is their product $[a_1 a_2, b, c]$. The converse is true if $a_1$ and $a_2$ are coprime.

Some extra condition is needed in the converse since for example the primitive form $[4,0,1]$ factors as the product of the nonprimitive concordant forms $[2,0,2]$ and $[2,0,2]$.

**Proof:** If the coefficients of $[a_1 a_2, b, c]$ have a common divisor then they have a common prime divisor, which will divide either $a_1$ or $a_2$, as well as $b$ and $c$, so one of the forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ will not be primitive. This gives the first statement. For the second, if one of $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ is not primitive, say $[a_1, b, a_2 c]$, then its coefficients will be divisible by some prime $p$. If $a_1$ and $a_2$ are coprime, then $p$ dividing $a_1$ and $a_2 c$ implies that $p$ divides $c$. Thus $p$ divides all three coefficients of $[a_1 a_2, b, c]$, making it nonprimitive. □

**Proposition 7.2.** For each pair of primitive forms $Q_1$ and $Q_2$ of discriminant $\Delta$ there exist forms $Q'_1$ and $Q'_2$ properly equivalent to $Q_1$ and $Q_2$ respectively, such that $Q'_1$ and $Q'_2$ are concordant.

There will be two main steps in the proof. We separate these off as lemmas since they will be used later as well.
Lemma 7.3. Let $Q_1 = [a_1, b_1, c_1]$ and $Q_2 = [a_2, b_2, c_2]$ be forms of the same discriminant with $a_1$ and $a_2$ coprime. Then there exist forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ that are properly equivalent to $Q_1$ and $Q_2$ respectively.

Proof: The first thing to consider is how to find a pair of forms properly equivalent to $Q_1$ and $Q_2$ having the same middle coefficient $b$. Recall that the edges in the topograph of a form have labels which we denoted $h$, with the sign of $h$ changing when the orientation of the edge is reversed. For a region in the topograph of $Q_1$ labeled $a_1$ we orient the edges bordering this region all in the same direction so that the region lies to the left as we move along the edges in the direction specified by their orientation. The edge labels then form an arithmetic progression with increment $2a_1$. One of these edges is labeled $b_1$, so the other edge labels are $b_1 + 2a_1 m$ for integers $m$. Similarly, in the topograph of $Q_2$ we have a region labeled $a_2$ whose bordering edges have labels $b_2 + 2a_2 n$ for integers $n$.

We would like one of the edge labels $b_1 + 2a_1 m$ to equal one of the edge labels $b_2 + 2a_2 n$. This means we would like to find integers $m$ and $n$ such that $a_1 m - a_2 n = (b_2 - b_1)/2$. Note that the right side of this equation is an integer since the edge labels in a topograph always have the same parity as the discriminant, which is the same for both forms by assumption. We know the equation $a_1 m - a_2 n = (b_2 - b_1)/2$ always has an integer solution $(m, n)$ if $a_1$ and $a_2$ are coprime. Thus we can find edges bordering the $a_1$ and $a_2$ regions with the same label $h = b$. The two given forms are therefore equivalent to forms $[a_1, b, c_1]$ and $[a_2, b, c_2]$, and in fact properly equivalent because of our rule for how to orient the edges. Equating the discriminants of these two forms leads to the equation $a_1 c_1 = a_2 c_2$ and since $a_1$ and $a_2$ are coprime this implies that $a_1$ divides $c_2$, so $c_2 = a_1 c$ for some integer $c$. As noted earlier, this implies that $c_1 = a_2 c$. Thus we have two concordant forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ properly equivalent to the original forms $Q_1$ and $Q_2$. 

Lemma 7.4. Given a primitive form $Q$ and a positive integer $n$, then $Q$ represents some number coprime to $n$.

Proof: We can assume $n > 1$ since every number is coprime to 1. Let $Q = [a, b, c]$ and let the prime factorization of $n$ be $n = p_1^{e_1} \cdots p_k^{e_k}$ for distinct primes $p_i$. We separate the primes $p_i$ into three disjoint classes: (1) primes that do not divide $a$, (2) primes that divide $a$ but not $c$, and (3) primes that divide both $a$ and $c$. Let $s$ denote the product of all the primes $p_i$ of type (1) and let $r$ be the product of all the primes $p_i$ of type (2). We have $Q(r, s) = ar^2 + brs + cs^2$ and we claim that this number is coprime to $n$. If not, then some $p_i$ divides $Q(r, s)$. If $p_i$ is of type (1) then $p_i$ divides $s$ but not $a$ so since it divides $Q(r, s)$ it must divide $r$, a contradiction since $r$ and $s$ are coprime. Similarly, if $p_i$ is of type (2) it divides $r$ but not $c$ so it must divide $s$, again a contradiction. And if $p_i$ is of type (3) it divides $a$ and $c$ but not $r$ or $s$ so it must divide $b$, which is impossible since the form $Q$ is primitive. Thus
$Q(r, s)$ and $n$ are coprime. Since $r$ and $s$ are coprime we then have a representation $Q(r, s) = ar^2 + brs + cs^2$ of a number coprime to $n$. \hfill \Box$

**Proof of Proposition 7.2:** Choose a number $a_1$ in the topograph of $Q_1$. By Lemma 7.4 the topograph of $Q_2$ contains some number $a_2$ coprime to $a_1$. Thus $Q_1$ and $Q_2$ are properly equivalent to forms $[a_1, b_1, c_1]$ and $[a_2, b_2, c_2]$. Lemma 7.3 then finishes the proof. \hfill \Box

To illustrate how to multiply forms let us look at a few examples in the case $\Delta = -104$. Here there are four equivalence classes of forms, with topographs shown below.

\begin{align*}
Q_1(x, y) &= x^2 + 26y^2 \\
Q_2(x, y) &= 2x^2 + 13y^2 \\
Q_3(x, y) &= 3x^2 + 2xy + 9y^2 \\
Q_4(x, y) &= 5x^2 + 4xy + 6y^2
\end{align*}

Since only the first two forms have mirror symmetry, the class number is 6.

Let us compute the product of $Q_2$ and $Q_3$ using the method in the proof of Lemma 7.3. To begin we need regions in the topographs of $Q_2$ and $Q_3$ with coprime labels, so the simplest thing is to use the region labeled 2 in the topograph of $Q_2$ and the region labeled 3 in the topograph of $Q_3$. For the $Q_2$ topograph the edge between the 2 and 13 regions is labeled 0 so the next edges bordering the 2 region are labeled 4, 8, 12, \ldots. For the 3 region in the topograph of $Q_3$ the bordering edges are labeled 2, 8, 14, \ldots starting with the edge adjacent to the 9 region. The number 8 is in both these arithmetic progressions so we choose this for $b$. In the $Q_2$ topograph this edge labeled 8 is between the regions labeled 2 and 21 so the form we want is $[2, 8, 21]$. For $Q_3$ the edge labeled 8 is between the 3 and 14 regions so the form corresponding to this edge is $[3, 8, 14]$. The product of these two concordant forms
is then \([6, 8, 7]\). The values of this form at \((x, y) = (0, 1), (1, 0), \text{ and } (1, 1)\) are 6, 7, and 21 so from the topograph of \(Q_4\) we see that this form is properly equivalent to \(Q_4\). Thus we have \(Q_2Q_3 = Q_4\) in \(CG(\Delta)\).

The product \(Q_4Q_4\), or in other words \(Q_4^2\), can be computed in the same way using the regions in the topograph of \(Q_4\) with the coprime labels 5 and 6. For the edges bordering the 5 region the labels starting with the edge between the 5 and 6 regions are 4, 14, 24, \(\cdots\). For the edges bordering the 6 region we can start with the same edge but now this edge must be oriented in the opposite direction in order to have the 6 region on our left as we move forward. The edge labels are then \(-4, 8, 20, \cdots\). Continuing these arithmetic progressions a little farther we find the common label 44 on the edge between the 5 and 102 regions, and on the edge between the 6 and 85 regions. Thus we have the concordant forms \([5, 44, 102]\) and \([6, 44, 85]\), with product \([30, 44, 17]\). The coefficients 30 and 17 appear in adjacent regions in the topograph of \(Q_3\) so \(Q_4^2\) is properly equivalent to either \(Q_3\) or the mirror image form. We can determine which by evaluating \([30, 44, 17]\) at \((x, y) = (-1, 1)\), giving the value 3. Thus in the topograph of \([30, 44, 17]\) the values 30, 17, 3 appear in clockwise order around a vertex, while in the topograph of \(Q_3\) they are in counterclockwise order, so these two topographs are mirror images and hence \(Q_4^2\) is properly equivalent to the mirror image form of \(Q_3\).

Alternatively, we could continue the arithmetic progressions 4, 14, 24, \(\cdots\) and \(-4, 8, 20, \cdots\) backwards and find the common edge label \(-16\) between the 5 and 18 regions and between the 6 and 15 regions. This gives the concordant forms \([5, -16, 18]\) and \([6, -16, 15]\). The topograph of their product \([30, -16, 3]\) has labels 30, 3, 17 at \((x, y) = (1, 0), (0, 1), (1, 1)\) in counterclockwise order. The numbers 30, 3, 17 appear in the topograph of \(Q_3\) in clockwise order around a vertex, so we reach the same conclusion that \(Q_4^2\) is properly equivalent to the mirror image form of \(Q_3\).

Returning to the general theory, we next prove the crucial fact that multiplication of proper equivalence classes of primitive forms by choosing a concordant pair of forms in these classes does not depend on which concordant pair we choose.

**Proposition 7.5.** For a fixed discriminant let \(Q_1, Q_2\) be a pair of concordant primitive forms and let \(Q_1', Q_2'\) be another such pair properly equivalent to \(Q_1\) and \(Q_2\) respectively. Then the products \(Q_1Q_2\) and \(Q_1'Q_2'\) are properly equivalent.

The proof will involve a certain amount of calculation, and to ease the burden it will be convenient to express quadratic forms in terms of matrices. This is based on the simple observation that a form \(ax^2 + bxy + cy^2\), regarded as a \(1 \times 1\) matrix \((ax^2 + bxy + cy^2)\), can be obtained as a product of a \(1 \times 2\) matrix, a \(2 \times 2\) matrix,
and a $2 \times 1$ matrix:

$$
\begin{pmatrix}
  x \\
y
\end{pmatrix}
\begin{pmatrix}
a & b/2 \\
 b/2 & c
\end{pmatrix}
\begin{pmatrix}
x \\
y
\end{pmatrix}
= \begin{pmatrix}
  ax + by/2 & bx/2 + cy
\end{pmatrix}
\begin{pmatrix}
x \\
y
\end{pmatrix}
= \begin{pmatrix}
  ax^2 + bxy + cy^2
\end{pmatrix}
$$

Thus we are expressing the form $ax^2 + bxy + cy^2$ as a matrix $\begin{pmatrix}
a & b/2 \\
 b/2 & c
\end{pmatrix}$. The entries $b/2$ might not be integers, but this will not matter for our purposes. For notational simplicity let us denote $b/2$ as $b$ so the matrix for $ax^2 + bxy + cy^2$ becomes $\begin{pmatrix}
a & b \\
b & c
\end{pmatrix}$.

When we do a change of variables by means of a matrix $\begin{pmatrix}
p & q \\
q & r
\end{pmatrix}$ with determinant $ps - qr = 1$, replacing $\begin{pmatrix}
x \\
y
\end{pmatrix}$ by $\begin{pmatrix}
p x + qy \\
r x + sy
\end{pmatrix}$, then the product $\begin{pmatrix}
x & y
\end{pmatrix} \begin{pmatrix}
a & b \\
b & c
\end{pmatrix} \begin{pmatrix}
x \\
y
\end{pmatrix}$ becomes $\begin{pmatrix}
x & y
\end{pmatrix} \begin{pmatrix}
a & b \\
b & c
\end{pmatrix} \begin{pmatrix}
p & q \\
q & r
\end{pmatrix} \begin{pmatrix}
  a' & b' \\
b' & c'
\end{pmatrix} \begin{pmatrix}
x \\
y
\end{pmatrix}$, with the second matrix being the transpose of the fourth matrix. Thus the matrix $\begin{pmatrix}
a & b \\
b & c
\end{pmatrix}$ for the form $ax^2 + bxy + cy^2$ is replaced by the matrix $\begin{pmatrix}
a' & b' \\
b' & c'
\end{pmatrix} = \begin{pmatrix}
p & q \\
q & r
\end{pmatrix} \begin{pmatrix}
a & b \\
b & c
\end{pmatrix} \begin{pmatrix}
p & q \\
q & r
\end{pmatrix}^{-1}$ for the new form $a'x^2 + b'xy + c'y^2$. We can write this last equation as

$$
\begin{pmatrix}
p & q \\
q & r
\end{pmatrix}
\begin{pmatrix}
a & b \\
b & c
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
= \begin{pmatrix}
a' & b' \\
b' & c'
\end{pmatrix}
\begin{pmatrix}
s & -q \\
r & p
\end{pmatrix}
$$

where this last matrix is the inverse of $\begin{pmatrix}
p & q \\
q & r
\end{pmatrix}$ since $ps - qr = 1$.

Proof of the Proposition: To abbreviate notation we will use the symbol $\approx$ for proper equivalence of forms.

Let $Q_1 = [a_1, b, a_2c]$ and $Q_2 = [a_2, b, a_1c]$, with $Q_1' = [a_1', b', a_2'c']$ and $Q_2' = [a_2', b', a_1'c']$. To begin the proof we look at the special case that $Q_1 = Q_1'$, so $a_1 = a_1'$, $b = b'$, and $a_2c = a_2'c'$. We assume $Q_2 \approx Q_2'$ so by the remarks preceding the proof there is an integer matrix $\begin{pmatrix}
p & q \\
q & r
\end{pmatrix}$ of determinant 1 such that

$$
\begin{pmatrix}
p & q \\
q & r
\end{pmatrix}
\begin{pmatrix}
a_2 & b \\
b & a_1c
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
= \begin{pmatrix}
a' & b' \\
b' & c'
\end{pmatrix}
\begin{pmatrix}
s & -q \\
r & p
\end{pmatrix}
$$

Multiplied out, this becomes

$$
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
= \begin{pmatrix}
a' & b' \\
b' & c'
\end{pmatrix}
\begin{pmatrix}
s & -q \\
r & p
\end{pmatrix}
$$

To show $Q_1Q_2 \approx Q_1Q_2'$ we would like to find an integer matrix $\begin{pmatrix}
p & q \\
q & r
\end{pmatrix}$ of determinant 1 such that

$$
\begin{pmatrix}
p & q \\
q & r
\end{pmatrix}
\begin{pmatrix}
a_1a_2 & b \\
b & c
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
= \begin{pmatrix}
a' & b' \\
b' & c'
\end{pmatrix}
\begin{pmatrix}
s & -q \\
r & p
\end{pmatrix}
$$

or in other words

$$
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
\begin{pmatrix}
ap + br & bp + a_1cr \\
a_2q + bs & bq + a_1cs
\end{pmatrix}
= \begin{pmatrix}
a' & b' \\
b' & c'
\end{pmatrix}
\begin{pmatrix}
s & -q \\
r & p
\end{pmatrix}
$$

We can convert the upper left entries in the two matrices in (*) to the corresponding entries in (**) by multiplying by $a_1$ provided that we choose $p = p$, $s = s$, and
The equality of the upper left entries in \((*)\) will imply equality of the corresponding entries in \((***)\). All four of the upper right entries in \((*)\) and \((***)\) will then be equal if we choose \(q = q/a_1\). All four lower left entries will also be equal, and the lower right entries in \((*)\) will be \(a_1\) times those in \((***)\). Note that the matrices \(\left(\begin{array}{cc}
p & q \\
r & s \\
\end{array}\right)\) and \(\left(\begin{array}{cc}
\bar{p} & \bar{q} \\
\bar{r} & \bar{s} \\
\end{array}\right)\) have the same determinant. The only problem is that \(\bar{q} = q/a_1\) will only be an integer if \(a_1\) divides \(q\). To guarantee that it does, observe that the equality of the upper right entries in \((*)\) implies that \(a_1cr = -a_2^2q\), so if \(a_1\) is coprime to \(a_2\) then \(a_1\) will divide \(q\). Thus we have proved the proposition in the special case \(Q_1 = Q_1'\) provided that \(a_1\) and \(a_2\) are coprime.

In the case just considered we assumed \(Q_1 = Q_1'\) which implied that \(b = b'\). Now let us assume merely that \(b = b'\) along with the previous hypothesis that \(a_1\) and \(a_2\) are coprime. Under these conditions the desired equivalence \(Q_1Q_2 \approx Q_1'Q_2'\) will be obtained as the combination of two equivalences \(Q_1Q_2 \approx Q_1Q_2' \approx Q_1'Q_2'\), but first we have to check that \(Q_1\) and \(Q_2'\) are concordant so that \(Q_1Q_2'\) is defined. Since \(b = b'\) and the determinants of \(Q_1\) and \(Q_1'\) are equal we have \(a_1a_2c = a_1'a_2'c'\). Since \(a_1\) and \(a_2\) are coprime it follows that \(a_1\) divides \(a_1'c'\). As we saw earlier, this suffices to imply that \(Q_1 = [a_1, b, a_2c]\) and \(Q_2' = [a_1', b, a_2'c']\) are concordant.

Assuming that \(a_1\) and \(a_2\) are coprime, the previous case \(Q_1 = Q_2\) now gives an equivalence \(Q_1Q_2 \approx Q_1Q_2\). Switching the roles of \(Q_1\) and \(Q_2\) as well as \(Q_1'\) and \(Q_2'\), the previous case also gives \(Q_1Q_2' \approx Q_1'Q_2'\) using the same assumption that \(a_2\) and \(a_1\) are coprime. We conclude that \(Q_1Q_2 \approx Q_1'Q_2'\).

Next we consider how to arrange that \(b = b'\). The hypothesis that will allow this is that \(a_1a_2\) is coprime to \(a_1'a_2'\), which is equivalent to saying that each of \(a_1\) and \(a_2\) is coprime to each of \(a_1'\) and \(a_2'\). If \(a_1a_2\) and \(a_1'a_2'\) are coprime we know by an earlier argument that the arithmetic progressions \(b + a_1a_2m\) and \(b' + a_1'a_2'n\) have a common value \(B\). This will also be a value in each of the arithmetic progressions \(b + a_1m, b + a_2n, b' + a_1'm,\) and \(b' + a_2'n\). Thus we have forms \(\bar{Q}_i = [a_i, B, \bar{c}_i] \approx Q_i\) for \(i = 1, 2\), and similarly \(\bar{Q}_i' = [a_i', B, \bar{c}_i'] \approx Q_i'\).

Let us check that \(\bar{Q}_1\) and \(\bar{Q}_2\) are concordant. This will be true if the first coefficient of one form divides the third coefficient of the other, say \(a_2\) divides \(\bar{c}_1\). The forms \(Q_1\) and \(\bar{Q}_1\) have the same discriminant so \(b^2 - 4a_1a_2c = B^2 - 4a_1\bar{c}_1\). Substituting \(B = b + 2a_1a_2m\) and simplifying, we get \(a_1a_2c = a_1a_2bm + a_1^2a_2^2m^2 - a_1\bar{c}_1\). After canceling a factor of \(a_1\) from both sides this becomes \(a_2c = a_2bm + a_1a_2^2m^2 - \bar{c}_1\) which implies that \(a_2\) divides \(\bar{c}_1\). Thus \(\bar{Q}_1\) and \(\bar{Q}_2\) are concordant, and by the same reasoning \(\bar{Q}_1'\) and \(\bar{Q}_2'\) are concordant, so we can form the products \(\bar{Q}_1\bar{Q}_2\) and \(\bar{Q}_1'\bar{Q}_2'\).

We have \(Q_1Q_2 \approx \bar{Q}_1\bar{Q}_2\) since the label \(B\) occurs on an edge bordering the region labeled \(a_1a_2\) in the topographs of both of these product forms. Similarly \(Q_1'Q_2' \approx \bar{Q}_1'\bar{Q}_2'\). We can now apply the previous case \(b = b'\) to the four forms \(\bar{Q}_1, \bar{Q}_2, \bar{Q}_1', \bar{Q}_2'\) since the leading coefficients \(a_1\) and \(a_2\) of the first and fourth forms are coprime. Thus we have \(\bar{Q}_1\bar{Q}_2 \approx \bar{Q}_1'\bar{Q}_2'\) and hence \(Q_1Q_2 \approx \bar{Q}_1\bar{Q}_2 \approx \bar{Q}_1'\bar{Q}_2' \approx Q_1'Q_2'\). This proves
the proposition under the assumption that \( a_1a_2 \) is coprime to \( a_1'a_2' \).

Now we can finish the proof by reducing to the case just considered. Choose a number \( A_1 \) represented by \( Q_1 \) coprime to \( a_1a_2a_1'a_2' \), and then choose a number \( A_2 \) represented by \( Q_2 \) and coprime to \( A_1a_1a_2'a_2' \). Since \( A_1 \) and \( A_2 \) are coprime there exist concordant forms \( \hat{Q}_1 = [A_1,B,A_2] \) and \( \hat{Q}_2 = [A_2,B,A_1] \) with \( \hat{Q}_1 \sim Q_1 \) and \( \hat{Q}_2 \sim Q_2 \). Since \( A_1A_2 \) is coprime to \( a_1a_2 \) the previous case implies that \( Q_1Q_2 \sim \hat{Q}_1\hat{Q}_2 \). The previous case also implies that \( \hat{Q}_1\hat{Q}_2 \approx Q'_1Q'_2 \) since \( A_1A_2 \) is coprime to \( a_1'a_2' \) and \( \hat{Q}_1 \approx Q_1' \) and \( \hat{Q}_2 \approx Q_2' \). Thus \( Q_1Q_2 \sim \hat{Q}_1\hat{Q}_2 \approx Q'_1Q'_2 \) and we are done.

\[ \square \]

For proper equivalence classes of primitive forms of a fixed discriminant we have seen that if two classes represent coprime numbers, then the product class represents the product of the two numbers. The next proposition says that we can drop the coprimeness condition on the two numbers if we allow “representations” \( Q(x,y) = n \) with nonprimitive pairs \((x,y)\).

**Proposition 7.6.** If \( Q_1 \) and \( Q_2 \) are concordant forms with product \( Q_1Q_2 \) then each product \( Q_1(x_1,y_1)Q_2(x_2,y_2) \) can be expressed as \( Q_1Q_2(X,Y) \) where \( X \) and \( Y \) are certain explicit functions of \((x_1,y_1)\) and \((x_2,y_2)\) given in terms of the coefficients of \( Q_1 \) and \( Q_2 \).

**Proof:** Let \( Q_1(x,y) = a_1x^2 + bx_1y + a_2cy^2 \) and \( Q_2(x,y) = a_2x^2 + bxy + a_1cy^2 \). It will suffice to express a product \( Q_1(x_1,y_1)Q_2(x_2,y_2) \) as \( a_1a_2X^2 + bXY + cY^2 \) where \( X \) and \( Y \) are given in terms of the coefficients \( a_1,a_2,c \) and the variables \( x_1,y_1,x_2,y_2 \).

First we compute \( Q_1(x_1,y_1)Q_2(x_2,y_2) \):

\[
(a_1x_1^2 + bx_1y_1 + a_2cy_1^2)(a_2x_2^2 + bx_2y_2 + a_1cy_2^2) \\
= a_1a_2x_1^2x_2^2 + a_1bx_1x_2y_2 + a_2^2cx_1^2y_2^2 + a_2bx_1x_2y_1 + b^2x_1x_2y_1y_2 \\
+ a_1bcx_1y_1y_2^2 + a_2^2cx_2^2y_1^2 + a_2bcx_2y_2^2y_1 + a_1a_2c^2y_1^2y_2^2 
\]

There are nine terms here and we label them (1)–(9). We want to choose \( X \) and \( Y \) so that the sum of these nine terms is equal to \( a_1a_2X^2 + bXY + cY^2 \). Only the terms (1) and (9) contain the factor \( a_1a_2 \) appearing in \( a_1a_2X^2 \) so to get (1) it is reasonable to start with \( X = x_1x_2 \). Then to get (9) we expand this to

\[
X = x_1x_2 \pm cy_1y_2 
\]

where we allow a sign \( \pm \) for flexibility later in the calculation. Now we have

\[
a_1a_2x^2 = a_1a_2x_1^2x_2^2 \pm 2a_1a_2cx_1x_2y_1y_2 + a_1a_2c^2y_1^2y_2^2 
\]

Here the outer terms are (1) and (9) but the middle term does not appear among (1)–(9) so we will have to have something that cancels it out later.

Next, to get the term (2) we start with \( Y = a_1x_1y_2 \) so that \( bXY \) starts with \( a_1bx_1^2x_2y_2 \) which is (2). For symmetry let us expand \( Y \) to \( Y = a_1x_1y_2 + a_2x_2y_1 \).
This leads to

\[ bXY = a_1bx_1^2x_2y_2 + a_2bx_1x_2^2y_1 \pm a_1bcx_1y_1y_2^2 + a_2bcx_2y_1^2y_2 \]

and

\[ cY^2 = a_1^2cx_1^2y_2^2 + 2a_1a_2cx_1x_2y_1y_2 + a_2^2cx_2^2y_1^2 \]

The first two terms of \( bXY \) give (2) and (4) while the outer terms of \( cY^2 \) give (3) and (7). Also, if we choose the sign \( \pm \) in \( X \) to be minus then the middle term of \( cY^2 \) cancels the middle term of \( a_1a_2X^2 \). The third and fourth terms of \( bXY \) would give (6) and (8) except they now have the wrong sign. We have also not yet accounted for the term (5). To do this let us add another term to \( Y \) so that \( X \) and \( Y \) are now

\[ X = x_1x_2 - cy_1y_2 \]
\[ Y = a_1x_1y_2 + a_2x_2y_1 + by_1y_2 \]

Then we have

\[ bXY = a_1bx_1^2x_2y_2 + a_2bx_1x_2^2y_1 - a_1bcx_1y_1y_2^2 - a_2bcx_2y_1^2y_2 \]
\[ + b^2x_1x_2y_1y_2 - b^2cy_1y_2^2 \]
\[ cY^2 = a_1^2cx_1^2y_2^2 + 2a_1a_2cx_1x_2y_1y_2 + a_2^2cx_2^2y_1^2 \]
\[ + b^2cy_1^2y_2^2 + 2a_1bcx_1y_1y_2^2 + 2a_2bcx_1y_1^2y_2 \]

Now the fifth term of \( bXY \) gives (5). Also the third and fourth terms of \( bXY \) combine with the last two terms of \( cY^2 \) to give (6) and (8) with the correct sign. Finally the last term of \( bXY \) cancels the fourth term of \( cY^2 \) and we are done. \( \square \)

[Explain general product formulas.]
2. The Class Group for Forms

In the previous section we defined a method for multiplying any two elements of the set $CG(\Delta)$ of proper equivalence classes of primitive forms of discriminant $\Delta$, namely choose a pair of concordant forms $Q_1 = a_1x^2 + bxy + a_2cy^2$ and $Q_2 = a_2x^2 + bxy + a_1cy^2$ in the two proper equivalence classes and then the product of these two classes is the class containing the form $Q_1Q_2 = a_1a_2x^2 + bxy + cy^2$. Note that the form $Q_1Q_2$ is the same as $Q_2Q_1$ so this multiplication operation in $CG(\Delta)$ is commutative.

**Proposition 7.7.** The set $CG(\Delta)$ with the multiplication just described is a group, that is, (1) the multiplication is associative, (2) there is an identity element whose product with any element is that element, and (3) each element has an inverse, so that the product of an element and its inverse is the identity element.

**Proof:** First we check that the class of the principal form serves as an identity element. Given a form $[a,b,c]$, this is concordant to the form $[1,b,ac]$ since the middle coefficient is the same for both forms and the first coefficient of each form divides the third coefficient of the other form. Thus we can form the product of the two forms and it equals the first form. The second form represents 1 so it is equivalent to the principal form, hence also properly equivalent to it since the principal form has mirror symmetry. Thus the class of the principal form is an identity element for the multiplication in $CG(\Delta)$.

Each form $[a,b,c]$ is concordant to its mirror image form $[c,b,a]$, and their product is $[ac,b,1]$ which represents 1 hence is properly equivalent the principal form. Thus inverses in $CG(\Delta)$ are obtained by taking mirror image forms.

Proving associativity will take a little more work. Here we start with three forms $Q_1, Q_2, Q_3$ giving three classes in $CG(\Delta)$. Choose a number $a_1$ in the topograph of $Q_1$, then a number $a_2$ in the topograph of $Q_2$ coprime to $a_1$, then a number $a_3$ in the topograph of $Q_3$ coprime to $a_1a_2$. Each $Q_i$ is then properly equivalent to a form $[a_i,b_i,c_i]$. Since each $a_i$ is coprime to the other two, the Chinese Remainder Theorem guarantees that there is a number $b$ congruent to $b_i$ mod $a_i$ for each $i$. We would like these congruences to be mod $2a_i$ instead of just mod $a_i$. To arrange this we go back and first choose $a_1$ coprime to 2, then $a_2$ coprime to $2a_1$, then $a_3$ coprime to $2a_1a_2$, so each $a_i$ is odd. Next, when we apply the Chinese Remainder Theorem we find $b$ congruent to each $b_i$ mod $a_i$ and also congruent to $\Delta$ mod 2, hence also congruent to each $b_i$ mod 2. Then $b$ will be congruent to each $b_i$ mod $2a_i$ since 2 and $a_i$ are coprime.

Having chosen $b$ in this way, each form $[a_i,b_i,c_i]$ is properly equivalent to a form $[a_i,b,c'_i]$. Equating discriminants of the first two of these new forms, we see that $a_1c'_1 = a_2c'_2$ so $a_2$ divides $a_1c'_1$ and hence it divides $c'_1$ since $a_1$ and $a_2$ are coprime. Similarly $a_3$ divides $c'_1$. Since $a_2$ and $a_3$ are coprime this means that $a_2a_3$...
divides $c'_1$ and we can write $c'_1 = a_2 a_3 c$ for some integer $c$. Equating discriminants then gives $c'_2 = a_1 a_3 c$ and $c'_3 = a_1 a_2 c$. Thus we have the three forms $[a_1, b, a_2 a_3 c]$, $[a_2, b, a_1 a_3 c]$, and $[a_3, b, a_1 a_2 c]$. Each pair of these forms is concordant. If we multiply the first two forms we get $[a_1 a_2, b, a_3 c]$, and then multiplying this by the third form gives $[a_1 a_2 a_3, b, c]$. We get the same result if we first multiply the second and third forms and then multiply their product by the first form. This proves associativity.

The identity element in a group is always unique since if two elements $g$ and $h$ both act as the identity then $gh = h$ since $g$ is an identity, but we also have $gh = g$ since $h$ is an identity, so $g = h$. Another general fact is that each element $g$ in a group has a unique inverse since if $h$ and $h'$ are two possibly different inverses for $g$, so both $gh$ and $gh'$ are the identity, then we have $gh = gh'$ so after multiplying both sides of this equation by any inverse $g^{-1}$ we get $h = h'$.

Forms whose topographs have mirror symmetry give elements of $CG(\Delta)$ that are equal to their inverses. The converse is also true since if a topograph is properly equivalent to its mirror image, this says it has an orientation-reversing symmetry and all such symmetries are mirror reflections by Proposition 5.7.

We can now re-examine some of the examples in the first section of Chapter 6 to verify that the conjectured group structures on $CG(\Delta)$ are in fact correct. We will be using the fact that if two primitive forms $Q_1$ and $Q_2$ of the discriminant $\Delta$ represent coprime numbers $a_1$ and $a_2$ then their product $Q_1 Q_2$ (in the group $CG(\Delta)$) represents $a_1 a_2$. This is a consequence of results in the previous section, particularly Lemma 7.3.

First consider the case $\Delta = 40$. Here there were two equivalence classes of forms, given by $Q_1 = x^2 - 10 y^2$ and $Q_2 = 2 x^2 - 5 y^2$. Both topographs have mirror symmetry so proper equivalence is the same as equivalence. Thus the group $CG(\Delta)$ has two elements, and we will use the same symbols $Q_1$ and $Q_2$ for these elements of $CG(\Delta)$. The identity element of $CG(\Delta)$ is $Q_1$ since this is the principal form. Since $Q_2 = Q_2^{-1}$ by the mirror symmetry of its topograph, we have $Q_2 Q_2 = Q_1$, the identity element of $CG(\Delta)$. This determines the group structure in $CG(\Delta)$ completely, and it agrees with what we observed from the topographs in Chapter 6.

Next consider the case $\Delta = -84$ where there were four equivalence classes of forms $Q_1$, $Q_2$, $Q_3$, and $Q_4$. All four topographs have mirror symmetry so $CG(\Delta)$ has four elements. The principal form $Q_1$ gives the identity element, and $Q_i Q_i = Q_1$ for each $i$ by the mirror symmetry. It remains to determine the products $Q_2 Q_3$, $Q_2 Q_4$, and $Q_3 Q_4$. In this case we can do this just from the topographs. For example, $Q_2$ represents 3 and $Q_3$ represents 2 so $Q_2 Q_3$ must represent 6. Thus we can determine $Q_2 Q_3$ by seeing which topograph contains 6, and we find it is $Q_4$, so $Q_2 Q_3 = Q_4$. Similarly one sees that $Q_2 Q_4 = Q_3$ using the numbers 3 and 5, and $Q_3 Q_4 = Q_2$ using
2 and 5. We could also deduce the last two formulas from \( Q_2Q_3 = Q_4 \) by multiplying both sides by \( Q_2 \) or \( Q_3 \).

For the third example we take \( \Delta = -56 \) where there were three equivalence classes of forms \( Q_1, Q_2, \) and \( Q_3 \). For \( Q_1 \) and \( Q_2 \) the topographs have mirror symmetry but not for \( Q_3 \) so there is another form \( Q_4 \) whose topograph is the mirror image of the one for \( Q_3 \), with \( Q_4 = Q_3^{-1} \) in \( CG(\Delta) \). Again we have \( Q_1 \) the identity in \( CG(\Delta) \) and we

have \( Q_2Q_2 = Q_1 \) by mirror symmetry. However it is not so easy to determine \( Q_3Q_3 \). The topograph of \( Q_3 \) contains 3 and 5 so the topograph of \( Q_3Q_3 \) must contain 15, but 15 is in the topographs of both \( Q_1 \) and \( Q_2 \) so this is inconclusive. The same thing happens for other pairs of primes in the topograph of \( Q_3 \) such as 3, 13 or 5, 19. However, since the topograph of \( Q_3 \) does not have mirror symmetry, we know that \( Q_3 \) is not \( Q_3^{-1} \) hence \( Q_3Q_3 \) is not \( Q_1 \) so it must be \( Q_2 \). Thus all four elements of \( CG(\Delta) \) are powers of \( Q_3 \), namely \( Q_3, Q_3^2 = Q_2, Q_3^3 = Q_2^2 = Q_1, \) and \( Q_3^4 = Q_4 \) since \( Q_3^4 = Q_1 \) implies \( Q_3^3 = Q_3^{-1} \) which is \( Q_4 \). This determines the structure of \( CG(\Delta) \) completely. For example \( Q_2Q_4 = Q_3^2Q_3^3 = Q_3^5 = Q_3 \) since \( Q_3^4 = Q_1 \).

In the preceding examples the group \( CG(\Delta) \) was small enough that its structure could be determined just from the topographs. This is not always the case in more complicated examples. The difficulty is that a form \( Q \) and its inverse \( Q^{-1} \) have mirror image topographs containing exactly the same numbers, so from the topographs one may be able to compute a product \( Q_iQ_j = Q_k^{-1} \) but one cannot always tell which exponent +1 or −1 is correct.

We illustrate this problem with an example, discriminant \( \Delta = -104 \) where we showed the topographs of the four equivalence classes of forms in the previous section. Since the first two forms \( Q_1 \) and \( Q_2 \) have mirror symmetry while the second two \( Q_3 \) and \( Q_4 \) do not, the group \( CG(\Delta) \) has six elements, with the principal form \( Q_1 \) the identity and \( Q_2^2 = Q_1 \). From the product \( 3 \cdot 17 = 51 \) we see that \( Q_3^2 \) is \( Q_1 \) or \( Q_3^{±1} \), but \( Q_1 \) is ruled out since the topograph of \( Q_3 \) does not have mirror symmetry, and \( Q_3^{±1} \) is ruled out since \( Q_3^2 = Q_3 \) would imply \( Q_3 = Q_1 \). Thus \( Q_3^2 = Q_3^{-1} \), or equivalently, \( Q_3^3 = Q_1 \). Similarly, we can try to compute \( Q_4^2 \) from the product \( 5 \cdot 7 = 35 \) which appears in the topographs of \( Q_1 \) and \( Q_3 \). The possibility that \( Q_4^2 \) is \( Q_1 \) is ruled out since \( Q_4 \) does not have mirror symmetry. Thus \( Q_4^2 = Q_3^{-1} \), but we cannot tell which exponent is correct from the topographs and the argument we used to compute \( Q_3^2 \) does not work here. In fact we computed \( Q_4^2 \) in the previous section by finding a pair of concordant forms properly equivalent to \( Q_4 \), and \( Q_4^2 \) turned out to be \( Q_3^{-1} \), the mirror image of \( Q_3 \).

Let us see what the higher powers of \( Q_4 \) are. Note first that \( Q_4^6 = (Q_4^2)^3 = (Q_3^3)^3 = (Q_3^{-1})^3 = Q_1 \) since \( Q_3^3 = Q_1 \). This implies that \( Q_4^3 = Q_4^{-1} \) and \( Q_4^4 = Q_4^{-2} = (Q_4^2)^{-1} = Q_3 \). For \( Q_4^3 \), this equals \( Q_4^2 \cdot Q_4 = Q_3^{-1}Q_4 \), and from the product \( 3 \cdot 5 = 15 \) we see from the topographs that this equals \( Q_2 \) or \( Q_4^{±1} \). But \( Q_4^3 = Q_4 \) implies \( Q_4^2 = Q_1 \) and \( Q_4^3 = Q_4^{-1} \) implies \( Q_4^4 = Q_1 \), contradicting previous calculations in both cases, so
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we deduce that \( Q_1^3 = Q_2 \). Thus all six elements of \( CG(\Delta) \) are powers of \( Q_4 \), namely \( Q_4, Q_4^2 = Q_3^{-1}, Q_4^3 = Q_2, Q_4^4 = Q_3, Q_4^5 = Q_4^{-1} \), and \( Q_4^6 = Q_1 \). This determines the multiplication in \( CG(\Delta) \) completely.

We would like to use the class group to determine which primitive forms of a given discriminant \( \Delta \) represent a given number \( n \). It will be convenient to assume that \( n \) is positive. This is no restriction when \( \Delta < 0 \) since there is no need to consider elliptic forms with negative values. When \( \Delta > 0 \), if we know which forms represent positive \( n \) then the negatives of these forms will be the ones representing \(-n\).

Here is our main result, where for convenience we continue to use the same symbol for a primitive form and for the element of \( CG(\Delta) \) that it determines:

**Theorem 7.8.** (1) Let a number \( n > 1 \) be factored as \( n = p_1^{e_1} \cdots p_k^{e_k} \) for distinct primes \( p_i \) with \( e_i > 0 \) for each \( i \). Then the primitive forms of discriminant \( \Delta \) that represent \( n \) are the products \( Q_1 \cdots Q_k \) where \( Q_i \) is a primitive form representing \( p_i^{e_i} \).

(2) The forms of discriminant \( \Delta \) representing a power \( p^e \) of a prime \( p \) not dividing the conductor of \( \Delta \) are primitive and are exactly the forms \( Q^{\pm e} \) where \( Q \) is a form representing \( p \). If \( p \) divides \( \Delta \) but not the conductor then \( e \) must be 1, but \( e \) is unrestricted if \( p \) does not divide \( \Delta \).

To put the two statements together, suppose we order the primes \( p_1, \ldots, p_k \) occurring in the factorization \( n = p_1^{e_1} \cdots p_k^{e_k} \) so that:

(a) The first \( j \) primes \( p_1, \ldots, p_j \) do not divide \( \Delta \).

(b) The next \( l \) primes \( p_{j+1}, \ldots, p_{j+l} \) divide \( \Delta \) but not the conductor of \( \Delta \).

(c) The remaining primes \( p_{j+l+1}, \ldots, p_k \) divide the conductor.

Then the elements in \( CG(\Delta) \) that represent \( n \) are the products

\[
Q_1^{e_1} \cdots Q_j^{e_j} Q_{j+1} \cdots Q_{j+l} Q_{j+l+1} \cdots Q_k
\]

where:

(i) \( Q_1, \ldots, Q_j \) are fixed choices of forms representing the primes \( p_1, \ldots, p_j \), with the signs \( \pm \) on the exponents \( e_1, \ldots, e_j \) chosen arbitrarily and independently.

(ii) The exponents \( e_{j+1}, \ldots, e_{j+l} \) are 1 and \( Q_{j+1}, \ldots, Q_l \) are the unique forms representing \( p_{j+1}, \ldots, p_{j+l} \).

(iii) For each \( i > j + l \) the forms \( Q_i \) range over all primitive forms representing \( p_i^{e_i} \).

In (ii) the forms \( Q_i \) have mirror symmetry since they can be defined by the formulas \( px^2 - ky^2 \) or \( px^2 + px y - ky^2 \) as in Proposition 6.4 so \( Q_1 = Q_{1}^{-1} \) and there is no need to write \( Q_1^{\pm 1} \). In (iii) the forms \( Q_i \) and exponents \( e_i \) that can occur are harder to determine. If \( e_i = 1 \) the only forms representing \( p_i \) are nonprimitive by Lemma 7.9 below so we must have \( e_i > 1 \). Some exponents \( e_i \) are excluded by Theorem 6.19. It
seems hard to predict which exponents $e_i$ are allowable and how many choices there are for $Q_i$ for the allowable exponents. We will look at some examples to illustrate this following the proof of the theorem.

In particular, if $\Delta$ is a fundamental discriminant then the conductor is 1 so case (iii) does not arise and the theorem gives a full reduction of the representation problem for nonprimes to the corresponding problem for primes.

Another consequence of the theorem is that any two primitive forms of discriminant $\Delta$ representing the same power of a prime not dividing the conductor are equivalent. For odd primes this was proved in Proposition 6.1, and now we see how it fits into the general picture.

Before proving the theorem we have a simple lemma:

**Lemma 7.9.** A form of discriminant $\Delta$ representing a power $p^e$ of a prime $p$ not dividing the conductor of $\Delta$ is primitive. If $p$ does divide the conductor then a form representing $p$ must be nonprimitive, but forms representing powers of $p$ can be either primitive or nonprimitive.

**Proof:** A form $Q$ representing a prime power $p^e$ is equivalent to a form $[p^e, b, c]$, so we may assume that $Q = [p^e, b, c]$. If this form is nonprimitive then $p$ must divide $b$ and $c$, so $Q = pQ'$ for some form $Q'$ of discriminant $\Delta/p^2$, which implies that $p$ divides the conductor of $\Delta$. Thus if $p$ does not divide the conductor then $Q$ must be primitive.

If $p$ divides the conductor of $\Delta$ then $\Delta/p^2$ is a discriminant. Letting $Q'$ be the principal form of discriminant $\Delta/p^2$, then $pQ'$ is a nonprimitive form of discriminant $\Delta$ representing $p$ since $Q'$ represents 1. Any other form of discriminant $\Delta$ representing $p$ is equivalent to $pQ'$ and so must be nonprimitive as well. An example of a primitive form representing a power of a prime dividing the conductor is $x^2 + 4y^2$ of discriminant $\Delta = -16$ with conductor 2, since this form represents 4 and 8 when $(x, y) = (0, 1)$ and $(2, 1)$. The nonprimitive form $2x^2 + 2y^2$ of the same discriminant represents 2 and 4.

**Proof of Theorem 7.8:** If $n$ is represented by a form $Q$ then $Q$ is properly equivalent to a form $[n, b, c]$. If $n$ factors as $n = a_1a_2\cdots a_k$ then $[n, b, c]$ factors as $[n, b, c] = [a_1, b, nc/a_1][n/a_1, b, a_1 c]$ with the latter two forms being concordant. If $k = 2$ this gives $[a_1a_2, b, c] = [a_1, b, a_2 c][a_2, b, a_1 c]$. If $k > 2$ we can factor $[n/a_1, b, a_1 c]$ further as $[a_2, b, nc/a_2][n/a_1a_2, b, a_1a_2 c]$. Continuing in this way we eventually get

$$[n, b, c] = [a_1, b, nc/a_1][a_2, b, nc/a_2]\cdots [a_k, b, nc/a_k]$$

with any two forms in the product on the right being concordant.

In particular for the prime factorization $n = p_1^{e_1}\cdots p_k^{e_k}$ we have $[n, b, c] = Q_1\cdots Q_k$ for $Q_i = [p_i^{e_i}, b, nc/p_i^{e_i}]$, a form representing $p_i^{e_i}$. By Lemma 7.1 the
form $[n, b, c]$ is primitive if and only if each $Q_i$ is primitive. This proves half of statement (1), that each primitive form representing $n$ can be expressed as a product $Q_1 \cdots Q_k$ with $Q_i$ a primitive form representing $p_i^{e_i}$. The other half is the statement that every such product $Q_1 \cdots Q_k$ is primitive and represents $n$. This follows by applying Lemma 7.3 repeatedly, first to forms $[p_1^{e_1}, b_1, c_1]$ and $[p_2^{e_2}, b_2, c_2]$ properly equivalent to $Q_1$ and $Q_2$, then to the product of the two resulting forms and a form $[p_3^{e_3}, b_3, c_3]$ properly equivalent to $Q_3$, and so on.

For part (2) of the theorem, a form representing $p^e$ is properly equivalent to a form $[p^e, b, c]$. As above, this factors as $[p^e, b, c] = [p, b, p^{e-1}c]$. If $p$ does not divide the conductor then the forms $Q = [p, b, p^{e-1}c]$ representing $p$ and $Q^e = [p^e, b, c]$ representing $p^e$ are primitive by the preceding lemma. Since forms representing primes are unique up to equivalence, any form representing $p$ must be properly equivalent to $Q$ or $Q^{-1}$. Hence the form we started with that represents $p^e$ is properly equivalent to the $e^{th}$ power of $Q$ or $Q^{-1}$, that is, to $Q^e$ or $Q^{-e}$.

The condition that $e = 1$ if $p$ divides $\Delta$ but not the conductor comes from the same requirement in Theorem 6.19. If $p$ does not divide $\Delta$ then Theorem 6.19 says that each power $p^e$ is represented by some form of discriminant $\Delta$, assuming $p$ itself is represented.

Let us look at a few examples. For $\Delta = -56$ we have already determined the group structure of $CG(\Delta)$ which has four elements, but we can use the preceding theorem to quickly rederive the group structure from the topographs. Observe that the prime 3, which does not divide the discriminant, is represented by the form $Q_3 = [3, 2, 5]$ while its powers $3^2$, $3^3$, and $3^4$ are represented by $Q_2 = [2, 0, 7]$, $Q_3$ again, and $Q_1 = [1, 0, 14]$. Since $3^2$ is represented by $Q_2$ this means $Q_3^2 = Q_2^{-1}$ but $Q_2 = Q_2^{-1}$ since the topograph of $Q_2$ has mirror symmetry, so $Q_3^2 = Q_2$. Next, $3^4$ is represented by $Q_1$, the identity in $CG(\Delta)$, so we have $Q_3^4 = Q_1$ and hence $Q_3^3 = Q_3^{-1}$, consistent with $3^3$ being represented by $Q_3$.

From this we can determine which forms represent a number $n = p_1^{e_1} \cdots p_k^{e_k}$, with $e_i \leq 1$ for $p_i = 2, 7$. Changing notation for convenience, let $Q$ be the form $[3, 2, 5]$ previously called $Q_3$. According to the previous theorem the forms representing $n$ are the products $(Q^{d_1})^{\pm e_1} \cdots (Q^{d_k})^{\pm e_k}$ where $Q^{d_i}$ is the power of $Q$ representing $p_i$. The resulting product is a power $Q^e$ where only the value of $e$ mod 4 matters. Primes $p_i$ represented by $Q^4 = 1$ can be ignored. The exponent sum can then be written as

$$e = \sum_{p_i \in Q} \pm e_i + \sum_{p_i \in Q^2} \pm 2e_i$$

where $p_i \in Q^j$ means primes $p_i$ represented by $Q^j$. There is no need for a sum over primes $p_i \in Q^3 = Q^{-1}$ since these are the same as the primes $p_i \in Q$ so these terms are already accounted for by the $\pm$ in the first sum. The sign $\pm$ in the second sum
does not matter since $Q^2 = Q^{-2}$. As we saw in Chapter 6 the forms $Q^0$ and $Q^2$ make up one genus while $Q$ and the equivalent form $Q^3 = Q^{-1}$ make up the other genus. The parity of $e$ thus determines the genus of the forms representing $n$. (Recall that forms representing a given number all belong to the same genus.) Notice that $n$ is represented by both $Q^0$ and $Q^2$ exactly when $e$ is even and at least one $e_i$ in the first sum is odd since this is the only time when the choice of the signs $\pm$ matters.

When $\Delta = -104$ we computed $CG(\Delta)$ to have six elements, the first through sixth powers of the form $Q = Q_4 = [5, 4, 6]$ with $Q^6 = 1$, the identity in $CG(\Delta)$. We can obtain most of this structure a little more efficiently now using the preceding theorem. Looking at the topographs we see that $5$, $5^2$, and $5^3$ are represented by $Q_4 = Q$, $Q_3$, and $Q_2$ so $Q^2 = Q_3^\pm 1$ and $Q^3 = Q_2^\pm 1 = Q_2$ (the topograph of $Q_2$ has mirror symmetry). Since $Q_2^2 = 1$ it follows that $Q^6 = Q_2^2 = 1$ so $Q^5 = Q^{-1}$ and $Q^4 = Q^{-2} = Q_3^{\pm 1}$. We cannot determine which sign in $Q^2 = Q_3^{\pm 1}$ is correct just from the topographs, but we computed that $Q^2 = Q_3^{\pm 1}$ earlier.

The story for representing a number $n = p_1^{e_1} \cdots p_k^{e_k}$ when $\Delta = -104$ is similar to what happened in the preceding example with $\Delta = -56$. For $\Delta = -104$ the exceptional primes $p_i$ with $e_i \leq 1$ are 2 and 13. The forms representing $n$ are the products $(Q^{q_i})^{\pm e_i} \cdots (Q^{q_k})^{\pm e_k}$ where $Q^{q_i}$ is the power of $Q$ representing $p_i$. Writing this product as $Q^e$ where only the value of $e$ mod 6 matters, the formula for $e$ now has another term:

$$e = \sum_{p_i \in Q} \pm e_i + \sum_{p_i \in Q^2} \pm 2e_i + \sum_{p_i \in Q^3} \pm 3e_i$$

The parity of $e$ again determines the genus, with one genus consisting of $Q^0 = 1$ and $Q^2$ (which is equivalent to $Q^4$) and the other genus consisting of $Q$ and $Q^3$ (with $Q^5 = Q^{-1}$ equivalent to $Q$). From the formula for $e$ one could work out when a number is represented by both forms within a genus and when it is represented by only one form.

As we have said, determining the primitive forms representing a power of a prime dividing the conductor can be difficult, like determining which form represents a given prime, but with the added complications that more than one form can represent the same prime power, and there can be different powers represented by the same form, sometimes even infinitely many different powers. One can get some idea of the complications from the table below which lists the powers $p^k$ of a prime $p$ dividing the conductor $d$ represented by the various forms of discriminant $\Delta$, both primitive and nonprimitive, for the nonfundamental discriminants in the range $-100 < \Delta < 0$. 

---

**Table:**

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<th>$\Delta$</th>
<th>$Q^0$</th>
<th>$Q^2$</th>
<th>$Q$</th>
<th>$Q^3$</th>
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<td>$Q_2$</td>
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<td>1</td>
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</tr>
</tbody>
</table>
Chapter 7 The Class Group for Quadratic Forms

The discriminant have exponents $k$ ranging over an interval from 0 to either a finite number or infinity, but the theorem gives no information about whether the representing forms are primitive or not. By Lemma 7.9 the prime $p$ itself is represented only by a nonprimitive form. For the discriminants for which the exponents range over a finite interval it seems that the largest exponent occurs only for primitive forms, but apart from this the nonprimitive forms represent all possible exponents.
When the interval of exponents is infinite it seems that all exponents are realized by both primitive and nonprimitive forms, except that primitive forms may omit a few small exponents. It is interesting that sometimes the exponents are split into disjoint sets represented by different forms, and sometimes they are not split up in this way. For the primitive forms representing infinitely many powers $p^k$ we have not proved that this happens, but it seems likely based on explicit calculations for the powers actually listed. In the first case of this sort, for $\Delta = -28$, we will prove in the next chapter that all the powers $2^k$ with $k \geq 3$ are in fact represented by the primitive form $[1,0,7]$.

From the preceding table we can also see examples where two primitive forms of the same discriminant but different genus can represent the same number, a power of a prime dividing the conductor. For discriminant $\Delta = -32$ the number 8 is represented by both $[1,0,8]$ and $[3,2,3]$ but these forms have different genus since the character $\chi_4$ is defined for $\Delta = -32$ but takes the value $+1$ at the number 1 in the topograph of $[1,0,8]$ and the value $-1$ at 3 in the topograph of $[3,2,3]$. Similarly for $\Delta = -63$ where the conductor is 3 both $[2,1,8]$ and $[4,1,4]$ represent 9 but these forms have different genus since the character $\chi_3$ has different values on the numbers 2 and 4 represented by these two forms.

This sort of behavior does not happen for representations of numbers coprime to the conductor. For numbers coprime to the discriminant this follows from the definition of a genus as the set of primitive forms of a fixed discriminant for which all characters take the same values, since these values are determined by any number in the topograph coprime to the discriminant. The generalization to numbers coprime to the conductor will be shown in Proposition 7.25.

3. Finite Abelian Groups

A group whose multiplication operation is commutative is usually referred to as an abelian group rather than a commutative group, although the latter term is sometimes used as well. The term “abelian” refers to the mathematician Niels Henrik Abel (1802–1829). The aim of this section is to explain the structure of abelian groups with finitely many elements. This structure is far simpler than for finite nonabelian groups which can be extremely complicated, with no hope of being completely classified.

The number of elements in a group $G$ is called the order of $G$. This can be finite or infinite, but for the class group $CG(\Delta)$ it is always finite since it is just the class number for discriminant $\Delta$.

For an element $g$ in a group $G$ the smallest positive integer $n$ such that $g^n$ is the identity is called the order of $g$ if such an $n$ exists, and otherwise the order of $g$ is said to be infinite. Each element $g$ in a finite group $G$ has finite order since the powers $g, g^2, g^3, \ldots$ cannot all be distinct elements of $G$, so we must have $g^m = g^n$ for some $m \neq n$, say $m < n$, and then if we multiply both $g^m$ and $g^n$ by $g^{-m}$, the
inverse of $g^m$, we see that $g^{n-m}$ is the identity. Thus some positive power of $g$ is the identity, and the smallest such power is the order of $g$. The identity element of a group always has order 1 and is obviously the only element of order 1.

It is a basic fact that the order of each element of a finite group $G$ divides the order of $G$, and we will show this later for finite abelian groups. As an illustration of this fact consider the group $CG(\Delta)$ for $\Delta = -104$ that we calculated in an example above. Here there was an element $Q$ such that $CG(\Delta)$ consists just of the powers $Q, Q^2, Q^3, Q^4, Q^5$ and $Q^6$, with $Q^6$ the identity element of $CG(\Delta)$, so the orders of $Q$ and $CG(\Delta)$ are both 6. The powers $Q^2$ and $Q^4$ have order 3 while $Q^3$ has order 2, as one can see by writing down the various powers of these elements. The element $Q^5 = Q^{-1}$ also has order 6, just like $Q$.

If an element $g$ of a group $G$ has order $n$ then all the powers $g, g^2, g^3, \ldots, g^n$ must be distinct elements of $G$, otherwise if two of these powers $g^i$ and $g^j$ were equal with $i < j$ we would have $g^{j-i}$ equal to the identity, with $j - i < n$, contradicting the assumption that $g$ has order $n$. If $g$ has order $n$ then the higher powers $g^{n+1}, g^{n+2}, \ldots$ just cycle through the powers $g, g^2, \ldots, g^n$ repeatedly. In particular the only powers of $g$ that are the identity element of $G$ are the powers $g^{kn}$ for integers $k$. The negative powers of $g$ are just the inverses of the positive powers, and these cycle through the same sequence $g, g^2, \ldots, g^n$ in reverse order since $g^{-1} = g^{n-1}$, $g^{-2} = g^{n-2}$, and so on.

A finite group $G$ is called cyclic if there is an element $g \in G$ such that every element of $G$ is a power of $g$, so the elements $g, g^2, \ldots$ cycle through all elements of $G$. The element $g$ is called a generator of $G$ in this case. If a generator $g$ has order $n$, so does $G$ since all the powers $g, g^2, g^3, \ldots, g^n$ must be distinct, as noted above. Thus a group of order $n$ is cyclic exactly when it contains an element of order $n$. Cyclic groups are automatically abelian since $g^k g^l$ and $g^l g^k$ both equal $g^{k+l}$.

Among the groups $CG(\Delta)$ that we computed above, $CG(\Delta)$ is cyclic of order 4 for $\Delta = -56$ and cyclic of order 6 for $\Delta = -104$, but for $\Delta = -84$ the group is not cyclic since it has order 4 but each element other than the identity has order 2.

Cyclic groups are easy to understand, and our next goal is to see that all finite abelian groups are built from cyclic groups by a fairly simple procedure. Given two groups $G_1$ and $G_2$, the product group $G_1 \times G_2$ is defined to be the set of all pairs $(g_1, g_2)$ with $g_1 \in G_1$ and $g_2 \in G_2$. The multiplication operation in $G_1 \times G_2$ is defined by $(g_1, g_2) \cdot (g'_1, g'_2) = (g_1 g'_1, g_2 g'_2)$, so the coordinates are multiplied separately. The identity element of $G_1 \times G_2$ is the pair $(g_1, g_2)$ with $g_1$ the identity in $G_1$ and $g_2$ the identity in $G_2$. The inverse of an element $(g_1, g_2)$ is $(g_1^{-1}, g_2^{-1})$. More generally one can define products $G_1 \times \cdots \times G_k$ of any collection of groups $G_1, \ldots, G_k$, with the elements of this product group being $k$-tuples $(g_1, \ldots, g_k)$ with $g_i \in G_i$ for each $i$.

If $G_1$ and $G_2$ are finite groups of orders $n_1$ and $n_2$, then $G_1 \times G_2$ has order $n_1 n_2$ since the two coordinates $g_1$ and $g_2$ of pairs $(g_1, g_2)$ in $G_1 \times G_2$ vary independently.
over $G_1$ and $G_2$. For an element $(g_1, g_2)$ in $G_1 \times G_2$, if $g_1$ has order $n_1$ and $g_2$ has order $n_2$ then the order of $(g_1, g_2)$ is the least common multiple of $n_1$ and $n_2$ since a power $(g_1, g_2)^n = (g_1^n, g_2^n)$ is the identity exactly when $n$ is a multiple of both $n_1$ and $n_2$, so the order of $g_1 g_2$ is the smallest such multiple. In particular, if $n_1$ and $n_2$ are coprime then $(g_1, g_2)$ has order $n_1 n_2$. This implies the following interesting fact:

**Proposition 7.10.** If $G_1$ and $G_2$ are cyclic of coprime orders $n_1$ and $n_2$, then $G_1 \times G_2$ is cyclic of order $n_1 n_2$.

**Proof:** Let $G_1$ consist of the powers $g_1, g_1^2, \cdots, g_1^{n_1}$ and let $G_2$ consist of the powers $g_2, g_2^2, \cdots, g_2^{n_2}$. If $n_1$ and $n_2$ are coprime then $(g_1, g_2)$ has order $n_1 n_2$ so the powers $(g_1, g_2), (g_1, g_2)^2, \cdots, (g_1, g_2)^{n_1 n_2}$ are all distinct and hence form the whole group $G_1 \times G_2$, which means that $G_1 \times G_2$ is cyclic of order $n_1 n_2$. \qed

**Theorem 7.11.** Every finite abelian group is a product of cyclic groups.

**Proof:** Let $G$ be a finite abelian group. We will use the notation $o(g)$ for the order of an element $g \in G$. The identity element of $G$ will be written simply as 1. Here is one preliminary fact we will need:

(1) If two elements $g_1$ and $g_2$ of $G$ have coprime orders then $o(g_1 g_2) = o(g_1) o(g_2)$.

This need not be true if $o(g_1)$ and $o(g_2)$ are not coprime. As an extreme example take $g_2$ to be $g_1^{-1}$. Another example would be to take $g_1$ to be an element of maximal order in $G$ and $g_2$ any element with $o(g_2) > 1$.

To prove statement (1) let $n_1 = o(g_1)$ and $n_2 = o(g_2)$. Certainly $(g_1 g_2)^{n_1 n_2} = 1$, so it will suffice to show that if $(g_1 g_2)^n = 1$ then $n$ is a multiple of $n_1 n_2$. If $(g_1 g_2)^n = 1$ then $g_1^n = g_2^{-n}$. Let $g$ be this element $g_1^n = g_2^{-n}$ of $G$. Then $g^{n_1} = g_1^{n_1} g_2^{-n_1} = 1$ so $o(g)$ divides $n_1$, and similarly $g^{n_2} = g_2^{-n_2} = 1$ so $o(g)$ divides $n_2$. Since $n_1$ and $n_2$ are assumed to be coprime, this means $o(g) = 1$. Thus $g = 1$, so $g_1^n = 1$ and $g_2^{-n} = 1$ which implies $g_2^n = 1$. Thus $n$ is divisible by $n_1$ and by $n_2$. Since $n_1$ and $n_2$ are coprime this implies that $n$ is divisible by $n_1 n_2$, which is what we wanted to show.

We will use (1) to prove the following:

(2) Let $m$ be the maximal order of elements of $G$. Then the order $o(g)$ of each element $g \in G$ divides $m$.

Before verifying this we make a small observation. If an element $g$ has order $n$ and $d$ is a divisor of $n$, say $n = dk$, then $g^k$ has order $d$.

Now suppose that statement (2) is false, so there is an element $g$ such that $o(g)$ does not divide the maximal order $m$. This means there is some prime power $p^k$ dividing $o(g)$ such that the highest power $p^l$ dividing $m$ has $l < k$. Let $g_1$ be a power of $g$ having order $p^k$ and let $g_2$ be an element of order $m/p^l$, for example $h^{m}$ where $h$ is an element of order $m$. Then by statement (1) the product $g_1 g_2$ has
order $p^k (m/p^l)$ which is greater than $m$ since $k > l$. This contradicts the maximality of $m$, so we conclude that $o(g)$ divides $m$ for all $g \in G$.

Having established (2) we can now begin the proof of the theorem. Let $g_1$ be an element of $G$ of maximal order $n_1$. If every element of $G$ is a power of $g_1$ then $G$ is cyclic and there is nothing else to prove. If there are elements of $G$ that are not powers of $g_1$ then we proceed by induction to find further elements $g_2, \ldots, g_q$ satisfying the following properties:

1. The elements $g_1, g_2, \ldots, g_q$ have orders $n_1, n_2, \ldots, n_q$ where each $n_i$ in this sequence is divisible by the next one $n_{i+1}$ and $n_i > 1$ for each $i$.
2. If $g_1^{k_1} \cdots g_a^{k_q} = g_1^{k_1} \cdots g_a^{k_q}$ then $k_i = k'_i$ mod $n_i$ for each $i$. Since each $g_i$ has order $n_i$ an equivalent statement is that if $g_1^{k_1} \cdots g_a^{k_q} = g_1^{k'_1} \cdots g_a^{k'_q}$ with $0 \leq k_i < n_i$ and $0 \leq k'_i < n_i$ for each $i$, then $k_i = k'_i$ for each $i$.

If we have elements $g_1, \ldots, g_q$ satisfying (3) and (4) such that their products $g_1^{k_1} \cdots g_a^{k_q}$ give all the elements of $G$, then by rewriting a product $g_1^{k_1} \cdots g_a^{k_q}$ as a $q$-tuple $(g_1^{k_1}, \ldots, g_a^{k_q})$ we see that $G$ is a product of cyclic groups of orders $n_1, \ldots, n_q$ and the proof will be complete.

If the products $g_1^{k_1} \cdots g_a^{k_q}$ do not account for all elements of $G$ then we will show how to find another element $g_{q+1}$ of order $n_{q+1}$ so that the conditions (3) and (4) are satisfied. This process can be iterated until all elements of $G$ are exhausted since at each step the number of products $g_1^{k_1} \cdots g_a^{k_q}$ at least doubles and $G$ has only finitely many elements.

To show how to find $g_{q+1}$ we consider congruence classes of elements of $G$ mod $g_1, \ldots, g_q$, which means that we consider each element $g$ as congruent to all the products $gg_1^{k_1} \cdots g_a^{k_q}$ for arbitrary exponents $k_i$. Let $[g]_q$ denote the congruence class of $g$, the set of all the elements $gg_1^{k_1} \cdots g_a^{k_q}$, including $g$ itself which occurs when each $k_i$ is 0. It is not hard to see that these congruence classes $[g]_q$ form an abelian group with the product defined by $[g]_q [g']_q = [gg']_q$. Let this group of congruence classes $[g]_q$ be denoted $G_q$.

If there are elements of $G$ that are not products $g_1^{k_1} \cdots g_a^{k_q}$ then $G_q$ has more than one element. Let $[g_{q+1}]_q$ be an element of $G_q$ of maximal order $n_{q+1}$ in $G_q$. First we check that $n_{q+1}$ divides $n_q$. Since $[g_{q+1}]_q^{n_{q+1}} = [1]_q$ we have $g_{q+1}^{n_{q+1}} = g_1^{k_1} \cdots g_a^{k_q}$ for some exponents $k_i$. Then in $G_{q-1}$ we have $[g_{q+1}]_{q-1}^{n_q} = [1]_{q-1}$ since all elements of $G_{q-1}$ have order dividing the maximal order, which is $n_q$ by the definition of $n_{q-1}$. The equation $[g_{q+1}]_{q-1}^{n_q} = [1]_{q-1}$ means that $g_{q+1}^{n_q}$ is a product of powers of $g_1, \ldots, g_{q-1}$, so it is certainly a product of powers of $g_1, \ldots, g_q$ which means $[g_{q+1}]_q^{n_q} = [1]_q$. Thus $n_q$ is a multiple of $n_{q+1}$, the order of $[g_{q+1}]_q$ in $G_q$, as we wanted to show. Since (3) holds it follows that $n_{q+1}$ divides each $n_i$ with $i < q$.

It is also true that $n_{q+1}$ divides each $k_i$ in the formula $g_{q+1}^{n_{q+1}} = g_1^{k_1} \cdots g_a^{k_q}$. To see this, consider the power $g_{q+1}^{n_{q+1}}$. We can write this as $g_{q+1}^{n_{q+1}} = (g_{q+1}^{n_{q+1}})^{n_i/n_{q+1}} = (g_{q+1}^{n_{i+1}})^{n_i/n_{q+1}} = g_1^{k_1} \cdots g_a^{k_q}$. This contradicts the maximality of $n_q$.
The proof gives a procedure for rechoosing $g_i$ than the maximal order of elements of $G_i$ just to ignoring the first coordinates of pairs $(h_{1i}, h_{2i})$ since we are free to change this coordinate arbitrarily by multiplying $(h_{1i}^{k_1}, h_{2i}^{k_2})$ by any element $(h_{1i}^{l_1}, 1)$. Next we choose an element $g_2$ of maximal order in $G_1$. For this we can choose $g_2 = (h_{1i}, h_{2i})$ for any $k_1$. If we choose $k_1$ to be 1 or 3 then $g_2$ will have order 4, which is larger than the maximal order of elements of $G_1$ which is 2. The next-to-last paragraph of the proof gives a procedure for rechoosing $g_2$ to have order equal to 2 rather than

$(g_1^{k_1} \cdots g_q^{k_q})^{n_i/n_q+1}$ with $n_i/n_q+1$ an integer since $n_q+1$ divides $n_i$. We can also write $g_i^{n_i+1}$ as a product $g_1^{l_1} \cdots g_q^{l_q}$ since $[g_i^{n_i+1}]_{i-1} = [g_i^{n_i+1}]_{i-1} = [1]_{i-1}$ as a consequence of the definition of $n_i$ as the maximal order of elements of $G_{i-1}$, so all elements of $G_{i-1}$ have order dividing $n_i$. Since the two expressions $(g_1^{k_1} \cdots g_q^{k_q})^{n_i/n_q+1}$ and $g_1^{l_1} \cdots g_q^{l_q}$ for $g_i^{n_i+1}$ are equal with $g_i$ not appearing in the second expression, the property $(4_q)$ implies that the exponent $k_i n_i/n_q+1$ on $g_i$ in the first expression must be a multiple of $n_i$, the order of $g_i$ by $(3_i)$. Thus we have $k_i n_i/n_q+1 = mn_i$ for some integer $m$. Canceling $n_i$ from this equation, we get $k_i/n_q+1 = m$ so $n_q+1$ divides $k_i$.

Next we would like to find an element $g_{q+1} g_1^{x_1} \cdots g_q^{x_q}$ congruent to $g_{q+1}$ mod $g_i, \cdots, g_q$ and having order $n_q+1$. The order of $g_{q+1} g_1^{x_1} \cdots g_q^{x_q}$ cannot be less than $n_q+1$ since it determines the same element of $G_q$ as $g_{q+1}$ and $[g_{q+1}]_{q}$ has order $n_q+1$ in $G_q$. Thus we just need to have $(g_{q+1} g_1^{x_1} \cdots g_q^{x_q})^n_q+1 = 1$. We have

$$(g_{q+1} g_1^{x_1} \cdots g_q^{x_q})^{n_q+1} = g_{q+1}^{n_q+1} g_1^{x_1 n_q+1} \cdots g_q^{x_q n_q+1} = g_1^{k_1} g_1^{x_1 n_q+1} \cdots g_q^{k_q}$$

and this will be 1 if $k_i + x_i n_q+1 = 0$ for each $i$. Solving $k_i + x_i n_q+1 = 0$ for $x_i$ gives $x_i = -k_i/n_q+1$ so $x_i$ is an integer since we have shown that $n_q+1$ divides $k_i$.

Having found an element $g_{q+1} g_1^{x_1} \cdots g_q^{x_q}$ of order $n_q+1$, we replace $g_{q+1}$ by this element, so the new $g_{q+1}$ has order $n_q+1$. It remains to check condition $(4_{q+1})$. If $g_1^{k_1} \cdots g_q^{k_q} g_{q+1}^{k_q+1} = g_1^{k_1'} \cdots g_q^{k_q'}$, then in $G_q$ we have $[g_{q+1}]_{q}^{k_q+1} = [g_{q+1}]_{q}^{k_q'}$. Since the order of $[g_{q+1}]_{q}$ in $G_q$ is $n_q+1$ this implies that $k_q+1 = k_q'$ mod $n_q+1$, hence $g_{q+1} = g_{q+1}$ in $G$ since $g_{q+1}$ has order $n_q+1$. We can then cancel $g_{q+1}^{k_q+1}$ and $g_{q+1}^{k_q'}$ from the equation $g_1^{k_1} \cdots g_q^{k_q} g_{q+1}^{k_q+1} = g_1^{k_1'} \cdots g_q^{k_q'}$ to get $g_1^{k_1} \cdots g_q^{k_q} = g_1^{k_1'} \cdots g_q^{k_q'}$. Since condition $(4_q)$ holds by induction we have $k_i = k_i'$ mod $n_i$ for each $i \leq q$. Thus $(4_{q+1})$ holds and we are done.

To illustrate how the preceding proof works suppose we start with the group $G = H_1 \times H_2$ where $H_1$ is cyclic of order 4 generated by an element $h_1$ and $H_2$ is cyclic of order 2 generated by an element $h_2$ of order 2. In this case we already know that $G$ is a product of cyclic groups, but suppose we forget this and just follow the proof through. At the first step we choose an element $g_1$ in $G$ of maximal order, so let us choose $g_1 = (h_{1i}, 1)$ which has order 4 in $G$. There are then two congruence classes of elements of $G$ mod $g_1$, namely the class consisting of the elements $(h_{1i}^{k_1}, h_{2i}^{k_2})$ with $k_2 = 0$ and the class with $k_2 = 1$, so the group $G_1$ of congruence classes mod $g_1$ has order 2. Intuitively, taking congruence classes mod $g_1$ amounts just to ignoring the first coordinates of pairs $(h_{1i}^{k_1}, h_{2i}^{k_2})$ since we are free to change this coordinate arbitrarily by multiplying $(h_{1i}^{k_1}, h_{2i}^{k_2})$ by any element $(h_{1i}^{l_1}, 1)$. Next we choose an element $g_2$ of maximal order in $G_1$. For this we can choose $g_2 = (h_{1i}, h_{2i})$ for any $k_1$. If we choose $k_1$ to be 1 or 3 then $g_2$ will have order 4, which is larger than the maximal order of elements of $G_1$ which is 2. The next-to-last paragraph of the proof gives a procedure for rechoosing $g_2$ to have order equal to 2 rather than
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4, so in the present example this would amount to choosing $k_1$ to be 0 or 2 rather than 1 or 3. Either choice $k_1 = 0$ or $k_1 = 2$ will work, but if we choose $k_1 = 0$ then the element $g_2$ becomes simply $(1, h_2)$ and a general product $g_1^{l_1}g_2^{l_2}$ becomes the general element $(h_1^{l_1}, h_2^{l_2})$ of $H_1 \times H_2$.

From the preceding theorem we can deduce a general fact mentioned earlier:

**Corollary 7.12.** Each element of a finite abelian group has order dividing the order of the group.

**Proof:** We can view the given group $G$ as a product $G_1 \times \cdots \times G_k$ of cyclic groups $G_i$ of order $n_i$. Then the order of an element $g = (g_1, \ldots, g_k)$ of $G$ is the least common multiple of the orders $o(g_i)$. The least common multiple divides any other common multiple (as one can see by looking at prime factorizations) hence the least common multiple divides $n_1 \cdots n_k$, the order of $G$, since $o(g_i)$ divides $n_i$ for each $i$. \qed

A statement equivalent to the corollary is that if a finite abelian group $G$ has order $n$ then $g^n = 1$ for each $g \in G$. This is because if $g^n = 1$ then the order of $g$ divides $n$ and conversely.

Fermat’s Little Theorem which we encountered in the proof of quadratic reciprocity in Chapter 6 is a special case of this corollary, the case that the group is the group of congruence classes mod $p$ of integers coprime to $p$, for $p$ an odd prime. The group operation is multiplication of congruence classes, and integers coprime to $p$ have multiplicative inverses mod $p$ so one does indeed have a group. The order of the group is $p - 1$, so each element has order dividing $p - 1$ which implies that $a^{p-1} \equiv 1 \pmod{p}$ for each integer $a$ coprime to $p$, as Fermat’s Little Theorem asserts.

The proof we gave for Fermat’s Little Theorem extends easily to give a simple proof of the corollary for any finite abelian group $G$. To see this, suppose $G$ has order $n$, with the elements of $G$ being $g_1, \ldots, g_n$. For an arbitrary element $g$ in $G$ the multiples $gg_1, \ldots, gg_n$ are all distinct since if $gg_i = gg_j$ then multiplying both sides of this equation by $g^{-1}$ gives $g_i = g_j$. Thus the sets $\{g_1, \ldots, g_n\}$ and $\{gg_1, \ldots, gg_n\}$ are equal. Taking the product of all the elements in each of these two sets and using commutativity of the multiplication operation, we have $g_1 \cdots g_n = g^n g_1 \cdots g_n$ which implies $g^n = 1$.

Fermat’s theorem was generalized by Euler to replace the prime $p$ by any number $n$. Here one takes the group of congruence classes mod $n$ of numbers coprime to $n$. As we know, these numbers have multiplicative inverses mod $n$ so we again have a group. Its order is given by Euler’s function $\varphi(n)$, the number of positive integers less than $n$ and coprime to $n$. The statement is then that $a^{\varphi(n)} \equiv 1 \pmod{n}$ for every $a$ coprime to $n$.

The preceding corollary implies that a finite abelian group $G$ of prime order $p$ must be cyclic since any non-identity element of $G$ must have order $p$. This holds
more generally if the order of $G$ is a product of distinct primes since by the theorem $G$ is a product of cyclic groups, and these groups must all have coprime orders so their product will also be cyclic by repeated applications of Proposition 7.10.

By Proposition 7.10, every cyclic group whose order is not a power of a prime can be expressed as a product of two cyclic groups of smaller order. Applying this fact repeatedly, every cyclic group is a product of cyclic groups of prime power order. Hence by Theorem 7.11 every finite abelian group is a product of cyclic groups of prime power order. A cyclic group of prime power order $p^k$ cannot be factored as a product since the factors would have orders $p^l$ for $l < k$ so the elements of the factors would have orders dividing $p^{k-1}$, hence the same would be true for all elements of the product, contradicting the fact that it is cyclic of order $p^k$ and so contains an element of order $p^k$.

**Proposition 7.13.** The factorization of a finite abelian group as a product of cyclic groups of prime power order is unique in the sense that any two such factorizations have the same number of factors of each order.

For example, if we let $C_n$ denote a cyclic group of order $n$, then the only two abelian groups of order 4 are $C_4$ and $C_2 \times C_2$. For order 8 there are three possibilities: $C_8$, $C_4 \times C_2$, and $C_2 \times C_2 \times C_2$. For order 16 there are five possibilities: $C_{16}$, $C_8 \times C_2$, $C_4 \times C_4$, $C_4 \times C_2 \times C_2$, and $C_2 \times C_2 \times C_2 \times C_2$. These examples illustrate the general fact that the abelian groups of order a prime power $p^k$ correspond exactly to the different partitions of $k$ as a sum of numbers from 1 to $k$. In the case of $2^4 = 16$ these were the five partitions $4$, $3 + 1$, $2 + 2$, $2 + 1 + 1$, and $1 + 1 + 1 + 1$. (The order of the terms does not matter, so $2 + 1 + 1$ is regarded as the same partition as $1 + 2 + 1$ and $1 + 1 + 2$.)

For groups whose order is a product of powers of different primes one just combines the various groups of each prime power independently. Thus for order $144 = 9 \cdot 16$ there are ten possibilities, the products of the five groups of order 16 listed above with either of the two groups $C_9$ and $C_3 \times C_3$ of order 9.

Thus we see that the only time that there is only one group of order $n$ is when $n$ is a product of distinct primes, so the group is a product of cyclic groups of distinct prime orders, making the whole group cyclic.

**Proof of Proposition 7.13:** The idea will be to characterize the number of cyclic factors of each prime power order in a way that does not depend on a particular choice of factorization. For a prime $p$ dividing the order of a finite abelian group $G$ let $G(p)$ be the set of elements in $G$ whose order is a power of $p$, including the identity element 1 of order $p^0$. Given a factorization of $G$ as a product $G_1 \times \cdots \times G_k$ of cyclic groups of prime power order, an element $(g_1, \ldots, g_k)$ of $G$ has order a power of $p$ exactly when each coordinate $g_i$ has order a power of $p$, or in other words, when $g_i = 1$ for each $G_i$ of order a power of a prime different from $p$. We can thus regard $G(p)$ as the
product of the factors $G_i$ of order a power of $p$. Thus we have a characterization of the product of these factors $G_i$ that does not depend on the choice of the factorization of $G$.

This gives a reduction of the problem to the case that $G = G(p)$, i.e., $G$ has order $p^n$ for some $n$, so we assume this from now on. It remains to give an intrinsic characterization of the number of cyclic factors of order $p^r$ for each $r$. We will do this by counting the number of elements in the set $G[q]$ of elements of $G$ that are $p^q$th powers of elements of $G$. We have $G = G[0]$ since all elements are $p^0$th powers, and $G[q]$ contains $G[q + 1]$ for each $q$ since $g^{p^{q+1}} = (g^p)^{p^q}$, so a $p^{q+1}$st power is also a $p^q$th power. The identity element 1 belongs to $G[q]$ for all $q$.

A cyclic group of order $p^r$ contains $p$ elements that are $p^{r-1}$st powers since if $g$ is a generator of the group and we set $s = p^{r-1}$ then $g^s, g^{2s}, g^{3s}, \ldots, g^{ps} = 1$ are distinct $p^{r-1}$st powers but after this there are just repetitions, with $g^{(p+1)s} = g^s, g^{(p+2)s} = g^{2s}$, and so on. Similarly there are $p^2$ elements that are $p^{r-2}$nd powers, and generally $p^r$ elements that are $p^{r-t}$th powers. For a product $G_1 \times \cdots \times G_k$ an element $(g_1, \ldots, g_k)$ is a $p^q$th power exactly when each coordinate $g_i$ is a $p^q$th power in $G_i$.

For a given factorization $G = G_1 \times \cdots \times G_k$ let $\phi(r)$ be the number of cyclic factors $G_i$ of order $p^r$ and let $p^m$ be the maximal order of elements of $G$. Thus $G[m] = \{1\}$. From the preceding paragraph we see that the number of elements in $G[m - 1]$ is $p^{\phi(m)}$ since each of the $\phi(m)$ factors $G_i$ of order $p^m$ has $p$ elements that are $p^{m-1}$st powers and in the other factors only the identity is a $p^{m-1}$st power. Since $G[m - 1]$ has $p^{\phi(m)}$ elements it follows that $G[m - 1]$ determines $\phi(m)$. Next, $G[m - 2]$ contains $(p^2)^{\phi(m)} p^{\phi(m-1)} = p^{2\phi(m)+\phi(m-1)}$ elements, with $p^2$ coming from each factor of order $p^m$ and $p$ from each factor of order $p^{m-1}$. Thus $G[m - 2]$ determines $\phi(m - 1)$ since $\phi(m)$ has already been determined. Continuing in the same way, we see that the subsets $G[q]$ determine all the numbers $\phi(r)$. Since the sets $G[q]$ are defined independently of how $G$ is factored as a product of cyclic groups, this finishes the proof. $\square$

The factorization of a finite abelian group as a product of cyclic groups of prime power order is the unique factorization with the largest number of factors since any other factorization with at least as many factors could be factored further into a product with prime-power cyclic factors, contradicting the uniqueness statement in the preceding proposition.

On the other hand there can be different factorizations into cyclic factors with the minimum number of factors. For example, if $p$ and $q$ are distinct primes then $C_{p^2q^2} \times C_{pq}$ and $C_{p^2q} \times C_{pq^2}$ are both the group $C_{p^2} \times C_p \times C_{q^2} \times C_q$. A natural way to factor a group $G$ as a product $G_1 \times \cdots \times G_k$ of cyclic groups with the minimum number of factors is via the following procedure. First factor $G$ as a product of cyclic
groups of prime power order. For each prime $p_i$ dividing the order of $G$ let $G(p_i)$ be the product of the factors of $G$ whose order is a power of $p_i$. For each $p_i$ choose a factor of $G(p_i)$ of maximal order and let $G_1$ be the product of these chosen factors, so $G_1$ has one factor for each prime $p_i$. Now repeat the process for the remaining factors of $G$ to obtain $G_2$, then once again for $G_3$ and so on until all the prime power cyclic factors of $G$ have been exhausted. In the end the number of factors $G_j$ is equal to the maximum number of factors in each of the groups $G(p_i)$.

In the preceding example of a group of order $p^3q^3$ this would yield the factorization $C_{p^2q} \times C_{pq}$. In general this procedure yields a product $C_{n_1} \times \cdots \times C_{n_k}$ with each $n_i$ divisible by $n_{i+1}$. This is the same factorization of $G$ as the one produced in the proof of Theorem 7.11, and is uniquely determined by the condition that each $n_i$ is divisible by $n_{i+1}$.

In the rest of this section we will give two propositions about finite abelian groups that will be applied to class groups in the next two sections. Both propositions have to do with the operation of squaring elements of a group. For the first proposition we consider elements whose square is the identity, that is, elements of order 1 or 2.

**Proposition 7.14.** In a finite abelian group $G$ the number of elements whose order is 1 or 2 is $2^e$ where $e$ is the number of factors of even order in any factorization of $G$ as a product of cyclic groups.

In general, when a finite abelian group $G$ is factored as a product of cyclic groups of prime power order, the number of factors of order a power of the prime $p$ is called the $p$-rank of $G$. The number $e$ in the proposition is thus the 2-rank of $G$. The proposition easily generalizes to the statement that the number of elements of $G$ of order 1 or $p$ is $p^r$ where $r$ is the $p$-rank of $G$.

**Proof:** Let $G = G_1 \times \cdots \times G_k$ be a factorization of $G$ as a product of cyclic groups. An element $(g_1, \ldots, g_k)$ of the product has order 1 or 2 exactly when each $g_i$ has order 1 or 2. (The order is 2 when at least one $g_i$ has order 2.) A cyclic group $C_{2n}$ of even order generated by an element $g$ has just one element of order 2, the element $g^n$, since a power $g^k$ with $0 < k < n$ has $g^{2k} \neq 1$ and the inverses of these elements are the powers $g^k$ with $n < k < 2n$ so these too do not have order 2. A cyclic group of odd order has no elements of order 2 since the order of an element always divides the order of the group. Thus if $e$ is the number of factors $G_i$ of even order, there are $e$ coordinates $g_i$ of $(g_1, \ldots, g_k)$ where we have a choice of two elements of $G_i$ of order 1 or 2 and in the other coordinates we must have $g_i = 1$. The number of elements of $G$ of order 1 or 2 is therefore $2^e$. \hfill \Box

For any abelian group $G$ we can form another group denoted $G/G^2$ whose elements are congruence classes of elements of $G$ mod squares, so $g_1 \equiv g_2$ if $g_2 = g_1g^2$ for some $g \in G$. This is analogous to taking congruence classes of integers mod 2.
except now the group operation is multiplication rather than addition. The multiplication in \( G/G^2 \) comes from multiplication in \( G \), so if we denote the congruence class of \( g \in G \) by \([g]\) then \([g_1][g_2]\) is defined to be \([g_1g_2]\). This is unambiguous since if \( g_1 = g_1' \) and \( g_2 = g_2' \), so \( g_1' = g_1h_1^2 \) and \( g_2' = g_2h_2^2 \) for some \( h_1, h_2 \in G \), then \( g_1g_2 = g_1'g_2' \) since \( g_1'g_2' = g_1g_2(h_1h_2)^2 \). The identity element of \( G/G^2 \) is \([1]\) where \( 1 \) is the identity of \( G \), and \([g]^{-1} = [g^-1]\). Since associativity in \( G/G^2 \) follows automatically from associativity in \( G \), we conclude that \( G/G^2 \) is a group, which is abelian since \( G \) is abelian.

**Proposition 7.15.** For a finite abelian group \( G \) the group \( G/G^2 \) is a product of cyclic groups of order 2 with one factor for each cyclic factor of \( G \) of even order.

**Proof:** Consider a factorization of \( G \) as a product \( G_1 \times \cdots \times G_k \) of cyclic groups. Since the square of an element \((g_1, \cdots, g_k)\) is \((g_1^2, \cdots, g_k^2)\), the group \( G/G^2 \) will be the product of the groups \( G_i/G_i^2 \). Thus the proposition reduces to the special case that \( G \) is a cyclic group. If \( G \) is cyclic of even order \( 2n \) with generator \( g \) then the squares in \( G \) are the even powers \( g^2, g^4, \cdots, g^{2n}, g^{2n+2} = g^2, g^{2n+4} = g^4, \cdots \) which are all congruent to 1. The odd powers \( g, g^3, \cdots, g^{2n-1}, g^{2n+1} = g, g^{2n+3} = g^3, \cdots \) are all congruent to each other but not to any even power of \( g \) so \( G/G^2 \) is cyclic of order 2. If \( G \) is cyclic of odd order \( 2n+1 \) then the squares \( g^2, g^4, \cdots, g^{2n}, g^{2n+2} = g, g^{2n+4} = g^3, \cdots \) form all of \( G \) so \( G/G^2 \) has order 1. \( \square \)

### 4. Symmetry and the Class Group

We have defined the symmetric class number \( h^s_\Delta \) for discriminant \( \Delta \) to be the number of equivalence classes of primitive forms of discriminant \( \Delta \) whose topographs have mirror symmetry. Thus \( h^s_\Delta \) is the number of elements in the class group \( CG(\Delta) \) whose order is 1 or 2 since mirror symmetric forms correspond to elements of \( CG(\Delta) \) satisfying \( Q = Q^{-1} \), which is the same as saying \( Q^2 = 1 \). (For symmetric forms there is no distinction between equivalence and proper equivalence.)

**Proposition 7.16.** (a) The symmetric class number \( h^s_\Delta \) is equal to \( 2^r \) where \( r \) is the 2 \(-\) rank of \( CG(\Delta) \), the number of cyclic factors of \( CG(\Delta) \) of order a power of 2 when \( CG(\Delta) \) is expressed as a product of cyclic groups of prime-power order.

(b) The ordinary class number \( h_\Delta \) is always a multiple of \( h^s_\Delta \), with \( h_\Delta = h^s_\Delta \) exactly when \( CG(\Delta) \) is a product of cyclic groups of order 2, and \( h^s_\Delta = 1 \) exactly when \( h_\Delta \) is odd.

**Proof:** Part (a) is just a restatement of Proposition 7.15 for \( CG(\Delta) \). For (b) consider the factorization of \( CG(\Delta) \) as a product of cyclic groups \( G_i \) of prime-power order. If \( h_\Delta \) is odd there are no even-order factors \( G_i \) so \( h^s_\Delta = 1 \). If \( h_\Delta \) is even, each \( G_i \) of order \( 2^k \) with \( k \geq 1 \) contributes a factor of 2 to \( h^s_\Delta \) while each \( G_i \) of odd order contributes nothing to \( h^s_\Delta \). The rest follows. \( \square \)
Corollary 7.17. If the number of distinct prime divisors of $\Delta$ is $k$ then the 2-rank of $CG(\Delta)$ is $k - 1$ except when $\Delta = 4(4m + 1)$ when the 2-rank is $k - 2$, and when $\Delta = 32m$ when the 2-rank is $k$. In particular the 2-rank is $k - 1$ when $\Delta$ is a fundamental discriminant.

We know that $CG(\Delta)$ is cyclic if the class number is prime or a product of distinct primes, but there are other cases when the structure of $CG(\Delta)$ as a product of cyclic groups is completely determined if one knows the class number as well as the prime factorization of $\Delta$, using the fact that the latter determines the 2-rank of $CG(\Delta)$ as in the preceding corollary. For example if the class number is 4 then $CG(\Delta)$ is either $C_4$ or $C_2 \times C_2$ and these two cases are distinguished by their 2-ranks. We saw this distinction between $C_4$ and $C_2 \times C_2$ for the fundamental discriminants $-56$ and $-84$ both of which have class number 4, but $-56$ has two distinct prime divisors so its class group is $C_4$ while $-84$ has three distinct prime divisors so its class group is $C_2 \times C_2$.

A similar thing works for class number 8 where the group is either $C_8$, $C_4 \times C_2$, or $C_2 \times C_2 \times C_2$, with different 2-ranks. For class number 16 on the other hand there is an ambiguity between $C_8 \times C_2$ and $C_4 \times C_4$. The first negative discriminant with class number 16 is $\Delta = -399 = -3 \cdot 7 \cdot 19$, a fundamental discriminant. Since there are three distinct prime factors of $\Delta$ the 2-rank of $CG(\Delta)$ is 2 so the ambiguity between $C_8 \times C_2$ and $C_4 \times C_4$ arises here. It is easy to compute that there are ten reduced forms of discriminant $-399$:


Labeling these as $Q_1, \ldots, Q_5$ in the top row and $Q_6, \ldots, Q_{10}$ in the second row, we see that there are four forms with mirror symmetry, $Q_1$, $Q_5$, $Q_6$, $Q_8$, the forms with two of their coefficients equal. This is in agreement with the 2-rank being 2. The six without symmetry count double in the class number which is therefore 16. To determine whether the class group is $C_8 \times C_2$ or $C_4 \times C_4$ it suffices to look for elements of order greater than 4. This happens to be very easy in this case if we look at which forms represent powers of 2. In the list above we see that $Q_2$ represents 2, $Q_3$ represents 4, $Q_9$ represents 8, and $Q_8$ represents 16. Since powers of primes not dividing the discriminant are always represented by unique equivalence classes of forms, it follows that $Q_2^2 = Q_3^{\pm 1}$, $Q_2^3 = Q_9^{\pm 1}$, and $Q_2^4 = Q_8$, with no sign ambiguity in the last case since $Q_8$ has mirror symmetry. In particular we see that $Q_2$ must have order greater than 4, so $CG(\Delta)$ is not $C_4 \times C_4$ and hence it must be $C_8 \times C_2$.

The order of $Q_2$ is 8 since there are no elements of order 16 in $C_8 \times C_2$. (This also follows from the fact that $Q_2^4$ has mirror symmetry hence must have order 2.)
As in the proof of Theorem 7.11 we can choose $Q_2$ as a generator of the $C_8$ factor of $CG(\Delta)$, and a generator of the $C_2$ factor can be chosen to be either $Q_5$ or $Q_6$, the two forms with mirror symmetry that are not a power of $Q_2$. Additional work would be needed to compute the remaining products $Q_iQ_j$ such as whether $Q_2^3$ is $Q_3$ or $Q_3^{-1}$. However some products can be determined without calculation, for example the fact that the product of any two of the symmetric forms $Q_5,Q_6,Q_8$ equals the third since these three forms together with $Q_1$ give a subgroup $C_2 \times C_2$ of $CG(\Delta)$, the elements of order 1 or 2.

A similar but even simpler sort of ambiguity occurs for class numbers $p^2$ with $p$ an odd prime, where the choice is between the groups $C_{p^2}$ and $C_p \times C_p$. The first example of this sort among negative discriminants occurs when $\Delta = -199$. The reduced forms are $Q_1 = [1,1,50]$, $Q_2 = [2,1,25]$, $Q_3 = [5,1,10]$, $Q_4 = [4,3,13]$, and $Q_5 = [7,5,8]$. Only $Q_1$ has mirror symmetry so the other four forms count twice in the class number which is therefore 9. To decide whether $CG(\Delta)$ is $C_9$ or $C_3 \times C_3$ we observe that $Q_2$ represents 2, $Q_4$ represents $2^2$, and $Q_5$ represents $2^3$, so $Q_2$ must have order greater than 3 in $CG(\Delta)$. Since the order of $Q_2$ must divide the order of $CG(\Delta)$ we see that $Q_2$ has order 9 and so $CG(\Delta)$ is $C_9$ rather than $C_3 \times C_3$.

The same strategy as in the preceding examples can be used to prove the following qualitative statement:

**Proposition 7.18.** For each number $n > 0$ there exists a negative fundamental discriminant $\Delta$ such that $CG(\Delta)$ contains an element of order greater than $n$.

This can be viewed as a refinement of the fact that the class number can be made arbitrarily large by taking $\Delta$ to have a large number of distinct prime factors, using a product of distinct odd primes if one wants a fundamental discriminant.

**Proof:** Consider the form $Q = [2,1,m]$ of discriminant $\Delta = 1 - 8m$ for a given number $m > 0$. This form is primitive and represents 2. Since $\Delta$ is odd, all powers $2^l$ are also represented by primitive forms of discriminant $\Delta$. The principal form is $[1,1,2m]$ and its topograph has a source vertex surrounded by the labels $1,2m,2m$. This implies that all other numbers in the topograph are greater than $2m$. Thus for a given number $n$, if we choose $m$ so that $2m > 2^n$, then the principal form will not represent $2^n$ or any smaller power of 2, so $Q$ will have order greater than $n$ in $CG(\Delta)$.

To do this with a fundamental discriminant we can start with a large number of distinct primes $p_1, \cdots, p_k$ greater than 10 and let $\Delta = -p_1 \cdots p_k$. If this is not 1 mod 8 we can multiply it by 3, 5, or 7 to ensure that it is, so $\Delta = 1 - 8m$ for some $m$ and the previous argument applies.

If one wanted to do this argument with $\Delta$ even rather than odd, one could use the prime 3 instead of 2, starting with the primitive form $[3,2,m]$ of discriminant $4-12m$ which is congruent to 1 mod 3 so all powers of 3 are represented by primitive
forms of this discriminant. The associated principal form is \([1,0,3m - 1]\), with no labels between 1 and \(3m - 1\) in its topograph. We leave it as an exercise to arrange that \(\Delta\) is a fundamental discriminant. \(\square\)

In general it is a hard question to determine which finite abelian groups occur as class groups. An interesting special case is to determine the values of \(n\) such that the product of \(n\) cyclic groups of order 2 is a class group \(CG(\Delta)\) for some \(\Delta\). By Proposition 7.16 this is equivalent to having \(h_\Delta = h_\Delta^s\), and we have mentioned that there is a list, probably complete, of 101 negative discriminants \(\Delta\) with this property. In these 101 cases the number of \(C_2\) factors of \(CG(\Delta)\) ranges from 0 to 4, so the class number is 1, 2, 4, 8, or 16. Thus it appears that a product of five or more copies of \(C_2\) cannot occur as a class group \(CG(\Delta)\) with \(\Delta < 0\). For \(\Delta > 0\) less seems to be known.

Here is a chart listing the smallest discriminants having class group a given abelian group of order up to 12:

<table>
<thead>
<tr>
<th>(CG(\Delta))</th>
<th>(C_1)</th>
<th>(C_2)</th>
<th>(C_3)</th>
<th>(C_4)</th>
<th>(C_2 \times C_2)</th>
<th>(C_5)</th>
<th>(C_6)</th>
<th>(C_7)</th>
</tr>
</thead>
<tbody>
<tr>
<td>(\Delta &lt; 0)</td>
<td>-3</td>
<td>-15</td>
<td>-23</td>
<td>-39</td>
<td>-84</td>
<td>-47</td>
<td>-87</td>
<td>-71</td>
</tr>
<tr>
<td>(\Delta &gt; 0)</td>
<td>5</td>
<td>12</td>
<td>148</td>
<td>136</td>
<td>60</td>
<td>401</td>
<td>316</td>
<td>577</td>
</tr>
<tr>
<td>(C_8)</td>
<td>-95</td>
<td>-260</td>
<td>-420</td>
<td>-199</td>
<td>-4027</td>
<td>-119</td>
<td>-167</td>
<td>-279</td>
</tr>
<tr>
<td>(C_9)</td>
<td>505</td>
<td>396</td>
<td>480</td>
<td>1129</td>
<td>32009</td>
<td>817</td>
<td>1297</td>
<td>1345</td>
</tr>
</tbody>
</table>

As one can see, for positive discriminants one usually needs to go farther than for negative discriminants to realize a given group.

While positive discriminants are more difficult both computationally and theoretically, they have an extra piece of structure that adds to their interest, namely the operation that sends a form \(Q\) to its negative \(-Q\). This gives a well-defined operation on \(CG(\Delta)\) since if two forms \(Q_1\) and \(Q_2\) are properly equivalent then so are \(-Q_1\) and \(-Q_2\) because an orientation-preserving linear fractional transformation taking the topograph of \(Q_1\) to the topograph of \(Q_2\) does the same for the topographs of \(-Q_1\) and \(-Q_2\). Also, if \(Q\) is primitive then obviously so is \(-Q\).

In \(CG(\Delta)\) the operation sending \(Q\) to \(-Q\) is generally different from the operation which sends \(Q\) to its mirror image form \(Q^{-1}\) in \(CG(\Delta)\). For example when \(\Delta = 12\) the group \(CG(\Delta)\) is cyclic of order 2 consisting of the principal form \(Q = x^2 - 3y^2\) and its negative \(-Q = -x^2 + 3y^2\) which is equivalent to \(3x^2 - y^2\). Thus \(Q\) and \(-Q\) are distinct elements of \(CG(\Delta)\), but both \(Q\) and \(-Q\) have mirror symmetry so \(Q = Q^{-1}\) and \(-Q = -Q^{-1}\). Note that there is never any ambiguity about whether \(-Q^{-1}\) is \(-(Q^{-1})\), the negative of the mirror image of \(Q\), or \((-Q)^{-1}\), the mirror image of the negative of \(Q\), since these are obviously the same.
Proposition 7.19. Inverses and negatives are related to symmetries and skew symmetries in the following ways:

(i) \( Q = Q^{-1} \) in \( CG(\Delta) \) if and only if the topograph of \( Q \) has a mirror symmetry.

(ii) \( Q = -Q \) in \( CG(\Delta) \) if and only if the topograph of \( Q \) has a 180 degree rotational skew symmetry.

(iii) \( Q = -Q^{-1} \) in \( CG(\Delta) \) if and only if the topograph of \( Q \) has a glide-reflection skew symmetry.

Proof: We can identify elements of \( CG(\Delta) \) with the parts of their topographs along the separator line as we usually picture it: A horizontal straight line with vertical edges leading off it either above or below, say at unit intervals, with positive labels on the regions between successive upward edges and negative labels on the regions between successive downward edges. (This data of vertical edges and labels along the separator line must be periodic, as we know.) In the actual topograph the regions have labels \( x/y \) corresponding to vertex labels in the Farey diagram, but we disregard these labels when we pass to proper equivalence classes of forms. We also do not distinguish between a separator line and any parallel translation of itself.

For statement (i) we begin with the fact that the topograph of \( Q^{-1} \) is the mirror image of the topograph of \( Q \). In terms of separator lines this is saying that the separator line of \( Q^{-1} \) is a mirror image of the separator line of \( Q \). Since we only consider horizontal separator lines with positive values above and negative values below, the reflection giving the mirror image must be across a vertical line, taking the right end of the separator line for \( Q \) to the left end of the separator line for \( Q^{-1} \) and vice versa. If \( Q \) and \( Q^{-1} \) are properly equivalent this means that the separator line of \( Q \) is taken to itself by reflecting across a vertical line and possibly translating along the line. The composition of the reflection and the translation can be achieved by a single reflection since every symmetry of a line switching its two ends is a reflection across some perpendicular line, as we saw in our initial discussion of symmetries in Chapter 5. Thus for \( Q \) to be properly equivalent to \( Q^{-1} \) means exactly that the separator line for \( Q \) has mirror symmetry under reflection across some vertical line. This is statement (i).

For (ii), the separator line for the negative of a form \( Q \) is obtained by first changing the sign of all the labels along the separator line for \( Q \) and then rotating the plane by 180 degrees to bring the positive labels back above the separator line. If \( Q \) is properly equivalent to \( -Q \) this means that these two operations of changing signs and rotating produce the same separator line we started with, which means just that the separator line has a rotational skew symmetry.

For (iii), If one form is properly equivalent to the mirror image of the negative of another then we can transform the first separator line to the second by first changing the signs of the labels and rotating it by 180 degrees to get the negative, then reflecting it across a vertical line to take the inverse. The composition of the rotation
and the reflection is a glide-reflection along the separator line. Thus the first spine is transformed into the second by a glide-reflection, which is what (iii) asserts. □

We can picture the relationships between inverses and negatives by the diagram at the right which can be viewed as a picture of a regular tetrahedron. The tetrahedron has three 180 degree rotational symmetries about the three axes passing through midpoints of opposite edges of the tetrahedron. One of these rotations sends each form to its inverse, another sends each form to its negative, and the third sends each form to the negative of its inverse. These rotational symmetries of the tetrahedron are related to symmetries and skew symmetries of forms in the following ways.

1. If $Q$ has mirror symmetry then so does $-Q$ so the top two forms are equal in $CG(\Delta)$ and so are the bottom two. The first of the three rotational symmetries of the tetrahedron realizes these equalities in $CG(\Delta)$.

2. If $Q$ has a rotational skew symmetry then so does $Q^{-1}$ so the two forms on the left are equal in $CG(\Delta)$ and so are the two on the right. These equalities are realized by the second rotation of the tetrahedron.

3. If $Q$ has a glide-reflection skew symmetry then so does $-Q$ so the two forms in each diagonal pair are equal in $CG(\Delta)$, and the third rotation of the tetrahedron gives these equalities.

When $Q$ has two of the three types of symmetries and skew symmetries, it has the third type as well, so all four forms are equal in $CG(\Delta)$. In this case we will say that $Q$ is fully symmetric. For example the principal form always has mirror symmetry and represents 1 so it is fully symmetric exactly when it represents $-1$ since Lemma 6.3 says this is equivalent to its having a skew symmetry.

Now let us see how negation of forms relates to multiplication in $CG(\Delta)$. One might guess that $(-Q_1)Q_2 = -(Q_1Q_2)$ as with numbers, but this turns out to be not quite right.

**Lemma 7.20.** In $CG(\Delta)$ the formula $(-Q_1)Q_2 = -(Q_1Q_2^{-1})$ holds for all $Q_1$ and $Q_2$. In particular, when $Q_1 = Q_2$ we have $(-Q_1)Q_1 = -Q_0$ where $Q_0$ is the principal form.

**Proof:** The forms $Q_1$ and $Q_2$ are properly equivalent to concordant forms $[a_1, b, a_2c]$ and $[a_2, b, a_1c]$. The form $[-a_1, -b, -a_2c] = [(-a_1), -b, a_2(-c)]$ is then concordant to $[a_2, -b, (-a_1)(-c)] = [a_2, -b, a_1c]$. Taking the product of these two concordant forms gives $[-a_1, -b, -a_2c][a_2, -b, a_1c] = [-a_1a_2, -b, -c]$. This says that $(-Q_1)(Q_2^{-1}) = -(Q_1Q_2)$. Replacing $Q_2$ by $Q_2^{-1}$ then gives the claimed formula $(-Q_1)Q_2 = -(Q_1Q_2^{-1})$. □
Proposition 7.21. If one element of $CG(\Delta)$ has a glide-reflection skew symmetry then so do all elements of $CG(\Delta)$. This occurs exactly for those discriminants for which the principal form represents $-1$.

Proof: Suppose that $Q$ is some form with a glide-reflection skew symmetry, so $Q = -Q^{-1}$ or equivalently $-Q = Q^{-1}$. Then if $Q_0$ is the principal form we have $Q_0 = Q^{-1}Q = (-Q)Q$ and this equals $-Q_0$ by the previous lemma. Thus $Q_0 = -Q_0$ if a single form has a glide-reflection skew symmetry. Once one has $Q_0 = -Q_0$, then for arbitrary $Q$ the formula $(-Q)Q = -Q_0$ says that $Q$ is the inverse of $-Q$, so $Q = -Q^{-1}$ which means that $Q$ has a glide-reflection skew symmetry. □

Corollary 7.22. If the class number $h_\Delta$ is odd then all forms in $CG(\Delta)$ have a glide-reflection skew symmetry but only the principal form has a rotational skew symmetry.

Proof: The principal form $Q_0$ has mirror symmetry and therefore so does $-Q_0$. Thus $(-Q_0)^2 = Q_0$. If $CG(\Delta)$ has odd order then it has no elements of order 2 so we must have $-Q_0 = Q_0$. Thus $Q_0$ has a rotational skew symmetry so it must also have a glide-reflection skew symmetry. By the preceding proposition all forms in $CG(\Delta)$ then have a glide-reflection skew symmetry. Any form which had a rotational skew symmetry would therefore also have a mirror symmetry and hence be of order 1 or 2 in $CG(\Delta)$, so it would have to be $Q_0$. □

One might ask whether the “one implies all” property in Proposition 7.21 also holds for the other two types of symmetries and skew symmetries. For mirror symmetries the only time all elements of $CG(\Delta)$ have mirror symmetry is when $CG(\Delta)$ is a product of cyclic groups of order 2, a rather rare occurrence that we have discussed before. For rotational skew symmetries it can happen that some forms have rotational skew symmetry while others do not. We just saw that when $CG(\Delta)$ has odd order only the principal form has rotational skew symmetry. An example where another form has rotational skew symmetry but the principal form does not is $\Delta = 136$. Here it is not hard to compute that there are three equivalence classes of forms: $Q_0 = [1,0,-34]$, $-Q_0 = [-1,0,34]$, and $Q_1 = [3,2,-11]$. Here are the topographs of $Q_0$ and $Q_1$:

\[ Q_0 = x^2 - 34y^2 \]

\[
\begin{array}{cccccccc}
1 & 2 & 1 \\
\end{array}
\]

\[ Q_1 = 3x^2 + 2xy -11y^2 \]

\[
\begin{array}{cccccccc}
3 & 5 & 6 & 11 & 10 \\
-11 & -6 & -5 & -3 & -10 & -11 & -6 & -5 & -3 & -10 & -11
\end{array}
\]
Since $Q_0$ and $-Q_0$ have mirror symmetry while $Q_1$ does not, the class number is 4. The group $CG(\Delta)$ must be $C_4$ rather than $C_2 \times C_2$ since it contains a form $Q_1$ without mirror symmetry, so this form has order 4 rather than 2. Thus $Q_1^2$ has order 2 so it must be the form $-Q_0$, as is confirmed by the fact that $Q_1$ represents 3 and $-Q_0$ represents 9. The topographs show that only $Q_1$ and $Q_1^{-1}$ have a rotational skew symmetry.

When do all primitive forms of discriminant $\Delta$ have a rotational skew symmetry? When this happens then in particular the principal form has a rotational skew symmetry, as well as a mirror symmetry, so it also has a glide-reflection skew symmetry. The previous proposition then says that all primitive forms have a glide-reflection skew symmetry, in addition to the assumed rotational skew symmetry, so they have mirror symmetry as well. Thus the class group is a product of cyclic groups of order 2 and the principal form represents $-1$. Conversely, these two conditions imply that all principal forms have mirror symmetry and glide-reflection skew symmetry, hence also rotational skew symmetry.

Another question one could ask is which discriminants have at least one primitive form with rotational skew symmetry. This turns out to have a very pleasing answer. As we observed in Chapter 5, the pivot points of rotational skew symmetries lie at the midpoints of edges of the separator line where the labels of the adjacent regions in the topograph are $a$ and $-a$. If the edge itself is labeled $b$ then the associated form is $[a, b, -a]$, and all such forms occur this way at pivot points of rotational skew symmetries. The discriminant of the form $[a, b, -a]$ is $b^2 + 4a^2$ so we are looking for solutions of $x^2 + 4y^2 = \Delta$. For $[a, b, -a]$ to be primitive means that the pair $(a, b)$ is primitive, so the question reduces just to finding the numbers represented by the form $x^2 + 4y^2$, excluding squares since we want the resulting forms $[a, b, -a]$ to be hyperbolic. (Squares correspond to 0-hyperbolic forms with rotational skew symmetry.) Here is a portion of the topograph of $x^2 + 4y^2$ showing also the labels $\frac{x}{y} = \frac{b}{a}$ which determine the associated forms $[a, b, -a]$.

The form $x^2 + 4y^2$ has discriminant $-16$ with class number 1. From Theorems 6.19
and 7.8 we can deduce that the numbers represented by \( x^2 + 4y^2 \) are the numbers \( 2^m p_1 \cdots p_k \) where \( m \) is 0, 2, or 3 and each \( p_i \) is a prime congruent to 1 mod 4. This tells us which discriminants have at least one primitive form with rotational skew symmetry.

A more refined question is how many different elements of \( CG(\Delta) \) have rotational skew symmetries. Solutions of \( b^2 + 4a^2 = \Delta \) come in groups of four obtained by varying the signs of \( a \) and \( b \). If we restrict attention just to the solutions with \( a \) positive, the primitive solutions \((a, b)\) correspond exactly to regions in the topograph of \( x^2 + 4y^2 \) labeled \( \Delta \), and these regions come in pairs, one in the upper half of the topograph and one in the lower half. Each topograph of a form with rotational skew symmetry has two pivot points on the separator line in each period. The sign of \( b \), which is the label on the edge with a pivot point, can be specified by orienting all edges of the separator line so that the regions on the left of the separator line have positive labels. Thus the number of proper equivalence classes of primitive forms of discriminant \( \Delta \) with rotational skew symmetry is half the number of regions labeled \( \Delta \) in the topograph of \( x^2 + 4y^2 \), and is therefore equal to the number of such regions in the upper half of the topograph. In other words the number of elements of \( CG(\Delta) \) with rotational skew symmetry equals the number of times that \( \Delta \) appears in the upper half of the topograph of \( x^2 + 4y^2 \). For example, a prime can appear only once in the upper half of the topograph by Proposition 6.3 so prime discriminants have only one form with rotational skew symmetry. This is consistent with the class number often being 1 for prime discriminants. In general the number of rotationally skew symmetric forms can be computed from the prime factorization of \( \Delta \) using methods from the next chapter. The result is that if \( \Delta = 2^m p_1^{e_1} \cdots p_k^{e_k} \) for distinct primes \( p_i \equiv 1 \mod 4 \) with each \( e_i > 0 \) then the number of primitive forms with rotational skew symmetry is \( 2^{k-1} \) when \( m = 0 \) or 2, and \( 2^k \) when \( m = 3 \).

5. Genus and the Class Group

The various genera of forms of discriminant \( \Delta \) are determined by the characters \( \chi \) associated to primes \( p \) dividing \( \Delta \), where \( \chi \) assigns a value \( \chi(n) = \pm 1 \) to each integer \( n \) not divisible by \( p \). Since each character has a constant value on all numbers in a topograph not divisible by \( p \), we can regard characters as functions from \( CG(\Delta) \) to \( \{ \pm 1 \} \). A key property of characters is that they are multiplicative, so \( \chi(n_1 n_2) = \chi(n_1) \chi(n_2) \). This implies that characters are also multiplicative as functions on \( CG(\Delta) \), meaning that \( \chi(Q_1 Q_2) = \chi(Q_1) \chi(Q_2) \) for forms \( Q_1 \) and \( Q_2 \) defining elements of \( CG(\Delta) \). This is because the topographs of \( Q_1 \) and \( Q_2 \) contain numbers \( n_1 \) and \( n_2 \) not divisible by \( p \) and coprime to each other by Lemma 7.4, and then the topograph of \( Q_1 Q_2 \) contains \( n_1 n_2 \). Thus \( \chi(Q_1 Q_2) = \chi(n_1 n_2) = \chi(n_1) \chi(n_2) = \chi(Q_1) \chi(Q_2) \).

Since the values of characters are \( \pm 1 \) this implies that \( \chi(Q^2) = +1 \) for each
primitive form \( Q \). Therefore \( \chi(Q_1 Q^2) = \chi(Q_1) \chi(Q^2) = \chi(Q_1) \) for all \( Q_1 \) and \( Q \). This means that characters define functions on the group \( CG(\Delta)/CG(\Delta)^2 \) of congruence classes of forms modulo squares. Let \( \text{Gen}(\Delta) \) be the set of genera in discriminant \( \Delta \). Since forms that are congruent modulo squares have the same genus, there is a well-defined function \( \Phi \) from \( CG(\Delta)/CG(\Delta)^2 \) to \( \text{Gen}(\Delta) \) sending each congruence class of forms to the genus of these forms.

**Proposition 7.23.** The function \( \Phi \) from \( CG(\Delta)/CG(\Delta)^2 \) to \( \text{Gen}(\Delta) \) is a one-to-one correspondence. Thus two primitive forms \( Q_1 \) and \( Q_2 \) of discriminant \( \Delta \) belong to the same genus if and only if when we regard them as elements of \( CG(\Delta) \) we have \( Q_2 = Q_1 Q_1 \) for some primitive form \( Q \) of discriminant \( \Delta \).

**Proof:** By the definition of genus, every genus contains at least one form, so \( \Phi \) is onto. Let us check that both \( CG(\Delta)/CG(\Delta)^2 \) and \( \text{Gen}(\Delta) \) have the same number of elements. By Corollary 6.17 the number of genera is equal to the number of elements of \( CG(\Delta) \) corresponding to forms with mirror symmetry, or in other words the number of elements of \( CG(\Delta) \) of order 1 or 2. By Propositions 7.14 and 7.15 this equals the number of elements of \( CG(\Delta)/CG(\Delta)^2 \). Finally, since \( \Phi \) is a function between two finite sets of the same size and \( \Phi \) is onto, it must also be one-to-one. \( \square \)

Let us illustrate this result by the example of discriminant \( \Delta = -104 \) which we have already looked at in some detail. In this case \( CG(\Delta) \) is a cyclic group of order 6. We have \( \left( \frac{-104}{p} \right) = \left( \frac{-26}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right) \left( \frac{13}{p} \right) = \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right) \left( \frac{p}{13} \right) \). The product \( \left( \frac{-1}{p} \right) \left( \frac{2}{p} \right) \) is +1 for \( p \equiv 1, 3 \mod 8 \) and −1 for \( p \equiv 5, 7 \mod 8 \) so this is the character we called \( \chi'_8 \) in Chapter 6, while \( \left( \frac{p}{13} \right) \) is \( \chi_{13} \), with the value +1 for \( p \equiv 1, 3, 4, 9, 10, 12 \mod 13 \) and −1 for \( p \equiv 2, 5, 6, 7, 8, 11 \mod 13 \). Using the fact that these characters take the same values on all numbers not divisible by 2 or 13 in the topograph of each form, we see from the topographs that \( Q_1 \) and \( Q_3^{\pm 1} \) belong to one genus where the character values are +1, +1, while \( Q_2 \) and \( Q_4^{\pm 1} \) make up the other genus with character values −1, −1. Expressing the forms as powers of \( Q_4 \) we see that the even powers form one genus and the odd powers the other genus. Thus two forms belong to the same genus exactly when one is a square times the other.

**Corollary 7.24.** Each genus contains the same number of proper equivalence classes of primitive forms.

**Proof:** For convenience we will again not distinguish between a primitive form and its proper equivalence class in \( CG(\Delta) \). Let \( Q_1, \ldots, Q_k \) be the distinct elements of \( CG(\Delta) \) in the genus of the principal form. These are exactly the elements of \( CG(\Delta) \) that are squares. If \( Q \) is any other element of \( CG(\Delta) \) then the genus of \( Q \) consists of \( QQ_1, \ldots, QQ_k \) since these are all the elements of \( CG(\Delta) \) obtained by multiplying \( Q \) by squares. These multiples of \( Q \) are all distinct since if \( QQ_i = QQ_j \) then after
multiplying by $Q^{-1}$ we have $Q_i = Q_j$ so $i = j$. Thus each genus consists of $k$ elements of $CG(\Delta)$.

Two extreme situations are worth keeping in mind:

(A) Each genus consists of a single equivalence class of forms. Among the equivalent ways of stating this condition are the following:

1. The number of genera equals the class number.
2. Every form has mirror symmetry.
3. Every element of $CG(\Delta)$ has order 2.
4. $CG(\Delta)$ is a product of cyclic groups of order 2.
5. The representation problem of determining which numbers are represented by each primitive form has a solution just in terms of congruence classes modulo the discriminant.

(B) The primitive forms of discriminant $\Delta$ all have the same genus, or in other words the number of genera is 1. Again there are equivalent statements:

1. The only primitive forms with mirror symmetry are the forms equivalent to the principal form.
2. $CG(\Delta)$ contains no elements of order 2.
3. $CG(\Delta)$ contains no elements of even order.
4. The class number is odd.

Discriminants where (A) or (B) occurs are rather rare. For (B), Corollary 5.9 says exactly when this happens in terms of the prime factorization of $\Delta$. For (A) there is no such simple characterization.

One final result:

**Proposition 7.25.** If two primitive forms of the same discriminant represent the same number coprime to the conductor then the two forms are in the same genus.

For numbers coprime to the discriminant this is a simple consequence of the definition of genus, but the result is less obvious in the more general situation, and indeed often fails to hold for numbers not coprime to the conductor, as we observed earlier in the chapter.

**Proof:** According to Theorem 7.8 we obtain the various primitive forms representing a number $n$ coprime to the conductor as products $Q_1^{e_1} \cdots Q_k^{e_k}$ where the prime factorization of $n$ is $n = p_1^{e_1} \cdots p_k^{e_k}$ and $Q_i$ represents $p_i$. Changing the exponent of $Q_i$ from $+e_i$ to $-e_i$ amounts to multiplying $Q_i^{e_i}$ by a square $Q_i^{-2e_i}$, and similarly for changing the exponent from $-e_i$ to $+e_i$. As we noted at the beginning of this section, multiplying a form by a square does not change its genus. So any two primitive forms representing $n$ have the same genus. ∎
Exercises

1. Show that all primitive forms of discriminant $\Delta$ representing a nonzero number $n$ coprime to the conductor are equivalent provided this is true whenever $n$ is a product of two primes not dividing $\Delta$.

2. Show that the numbers represented by $x^2 + 4y^2$ are the numbers $2^m p_1 \cdots p_k$ where $m$ is 0, 2, or 3 and each $p_i$ is a prime congruent to 1 mod 4.
Chapter 8. Quadratic Fields

Even when one’s primary interest is in integer solutions to equations, it can sometimes be very helpful to consider more general sorts of numbers. For example, when studying the principal quadratic form \( x^2 - Dy^2 \) of discriminant \( 4D \) it can be a great aid to understanding to allow ourselves to factor this form as \((x + y\sqrt{D})(x - y\sqrt{D})\). Here we allow \( D \) to be negative as well as positive, in which case we would be moving into the realm of complex numbers.

To illustrate this idea, consider the case \( D = -1 \), so the form is \( x^2 + y^2 \) which we are factoring as \((x + yi)(x - yi)\). Writing a number \( n \) as a sum \( a^2 + b^2 \) is then equivalent to factoring it as \((a + bi)(a - bi)\). For example \( 5 = 2^2 + 1^2 = (2 + i)(2 - i) \), and \( 13 = 3^2 + 2^2 = (3 + 2i)(3 - 2i) \), so 5 and 13 are no longer prime when we allow factorizations using numbers \( a + bi \). Sometimes a nonprime number such as 65 can be written as the sum of two squares in more than one way: \( 65 = 8^2 + 1^2 = 4^2 + 7^2 \), so it has factorizations as \((8 + i)(8 - i)\) and \((4 + 7i)(4 - 7i)\). This becomes more understandable if one uses the factorization \( 65 = 5 \cdot 13 = (2 + i)(2 - i)(3 + 2i)(3 - 2i) \).

Thus we will consider the set

\[
\mathbb{Z}[\sqrt{D}] = \{ x + y\sqrt{D} \mid x, y \in \mathbb{Z} \}
\]

which consists of real numbers if \( D > 0 \) and complex numbers if \( D < 0 \). We will always assume \( D \) is not a square, so \( \mathbb{Z}[\sqrt{D}] \) is not just \( \mathbb{Z} \). When \( D = -1 \) we have \( \mathbb{Z}[\sqrt{D}] = \mathbb{Z}[i] \), and numbers \( a + bi \) in \( \mathbb{Z}[i] \) are known as Gaussian integers.

We will also have occasion to consider numbers \( x + y\sqrt{D} \) where \( x \) and \( y \) are allowed to be rational numbers, not just integers. The set of all such numbers is

\[
\mathbb{Q}(\sqrt{D}) = \{ x + y\sqrt{D} \mid x, y \in \mathbb{Q} \}
\]

Here round parentheses are used instead of square brackets as a way of emphasizing that \( \mathbb{Q}(\sqrt{D}) \) is a field while \( \mathbb{Z}[\sqrt{D}] \) is only a ring. In other words, in \( \mathbb{Q}(\sqrt{D}) \) we can perform all four of the basic arithmetic operations of addition, subtraction, multiplication, and division, whereas in \( \mathbb{Z}[\sqrt{D}] \) only the first three operations are possible in general. (Multiplicative inverses of nonzero elements of \( \mathbb{Q}(\sqrt{D}) \) are given by the formula \((x + y\sqrt{D})^{-1} = \frac{x - y\sqrt{D}}{x^2 - Dy^2} \).)
1. Prime Factorization

The ring \( \mathbb{Z}[^{\sqrt{D}}] \) is useful for factoring the form \( x^2 - Dy^2 \) as \( (x + y\sqrt{D})(x - y\sqrt{D}) \). For this form the discriminant \( \Delta = 4D \) is 0 mod 4, and it would be nice to treat also the discriminants \( \Delta = 4d + 1 \equiv 1 \mod 4 \), when the principal form is \( x^2 + xy - dy^2 \). This can be factored as

\[
x^2 + xy - dy^2 = (x + \frac{1 + \sqrt{1+4d}}{2}y)(x + \frac{1 - \sqrt{1+4d}}{2}y)
\]

To simplify the notation we let \( \omega = \frac{(1 + \sqrt{1+4d})}{2} \) and \( \overline{\omega} = \frac{(1 - \sqrt{1+4d})}{2} \), the conjugate of \( \omega \), so the factorization becomes

\[
x^2 + xy - dy^2 = (x + \omega y)(x + \overline{\omega} y)
\]

The quadratic equation satisfied by \( \omega \) is \( \omega^2 - \omega - d = 0 \). Thus \( \omega^2 = \omega + d \) and this allows the product of two numbers of the form \( m + n\omega \) to be written in the same form. In other words, the set

\[
\mathbb{Z}[\omega] = \{ x + y\omega \mid x, y \in \mathbb{Z} \}
\]

is closed under multiplication and hence is a ring, just like \( \mathbb{Z}[^{\sqrt{D}}] \).

For example, when \( d = -1 \) we have \( \omega = \frac{(1 + \sqrt{-3})}{2} \) and the elements of \( \mathbb{Z}[\omega] \) form a lattice of equilateral triangles in the \( xy \)-plane:

The picture for larger negative values of \( d \) is similar but stretched in the vertical direction. In these cases the \( xy \)-plane is just the plane of complex numbers. When \( d \) is positive we can still draw the same figure but this is just a schematic representation of \( \mathbb{Z}[\omega] \) since all the numbers in \( \mathbb{Z}[\omega] \) are real numbers in this case.

Elements of \( \mathbb{Z}[\omega] \) can always be written in the form \( m + n\omega = (a + b\sqrt{1 + 4d})/2 \) for suitable integers \( a \) and \( b \). Here \( a \) and \( b \) must have the same parity since this is true for \( \omega = \frac{(1 + \sqrt{1+4d})}{2} \) and hence for any integer multiple \( n\omega \), and then adding an arbitrary integer \( m \) to \( n\omega \) preserves the equal parity condition since it adds an even integer to \( a \). Conversely, if two integers \( a \) and \( b \) have the same parity then \( (a + b\sqrt{1 + 4d})/2 \) lies in \( \mathbb{Z}[\omega] \) since by adding or subtracting a suitable even integer from \( a \) we can reduce to the case \( a = b \) when one has a multiple of \( \omega \). Notice
that having both \(a\) and \(b\) even is equivalent to \((a + b\sqrt{1 + 4d})/2\) lying in \(\mathbb{Z}[\sqrt{1 + 4d}]\), so \(\mathbb{Z}[\sqrt{1 + 4d}]\) is a subring of \(\mathbb{Z}[\omega]\). In the figure above we can see that \(\mathbb{Z}[\sqrt{1 + 4d}]\) consists of the even rows, the numbers \(m + n\omega\) with \(n\) even.

To have a unified notation for both the cases \(\mathbb{Z}[\sqrt{D}]\) and \(\mathbb{Z}[\omega]\) let us define \(R_\Delta\) to be \(\mathbb{Z}[\sqrt{\Delta}]\) when the discriminant \(\Delta\) is \(4D\) and \(\mathbb{Z}[\omega]\) when \(\Delta = 4d + 1\). We will often write elements of \(R_\Delta\) using lower case Greek letters, for example \(\alpha = x + y\sqrt{\Delta}\) or \(\alpha = x + y\omega\).

The main theme of this section and the next one will be how elements of \(R_\Delta\) factor into ‘primes’ within \(R_\Delta\). For example, if a prime \(p\) in \(\mathbb{Z}\) happens to be representable as \(p = x^2 - Dy^2\) then this is saying that \(p\) is no longer prime in \(\mathbb{Z}[\sqrt{D}]\) since it factors as \(p = (x + y\sqrt{D})(x - y\sqrt{D}) = \alpha\overline{\alpha}\) for \(\alpha = x + y\sqrt{\Delta}\) and \(\overline{\alpha} = x - y\sqrt{\Delta}\). Of course, we should say precisely what we mean by a ‘prime’ in \(\mathbb{Z}[\sqrt{D}]\) or \(\mathbb{Z}[\omega]\). For an ordinary integer \(p > 1\), being prime means that \(p\) is divisible only by itself and 1. If we allow negative numbers, we can ‘factor’ a prime \(p\) as \((-1)(-p)\), but this should not count as a genuine factorization, otherwise there would be no primes at all in \(\mathbb{Z}\). In \(R_\Delta\) things can be a little more complicated because of the existence of units in \(R_\Delta\), the nonzero elements \(\varepsilon\) in \(R_\Delta\) whose inverse \(\varepsilon^{-1}\) also lies in \(R_\Delta\). For example, in the Gaussian integers \(\mathbb{Z}[i]\) there are four obvious units, \(\pm 1\) and \(\pm i\), since \((i)(-i) = 1\). We will see in a little while that these are the only units in \(\mathbb{Z}[i]\). Having four units in \(\mathbb{Z}[i]\) instead of just \(\pm 1\) complicates the factorization issue slightly, but not excessively so.

For positive values of \(\Delta\) things are somewhat less tidy because there are always infinitely many units in \(R_\Delta\). For example, in \(\mathbb{Z}[\sqrt{2}]\) the number \(\varepsilon = 3 + 2\sqrt{2}\) is a unit because \((3 + 2\sqrt{2})(3 - 2\sqrt{2}) = 1\). All the powers of \(3 + 2\sqrt{2}\) are therefore also units, and there are infinitely many of them since \(3 + 2\sqrt{2} > 1\) so \((3 + 2\sqrt{2})^n \to \infty \) as \(n \to \infty\).

Whenever \(\varepsilon\) is a unit in \(R_\Delta\) we can factor any number \(\alpha\) in \(R_\Delta\) as \(\alpha = (\alpha\varepsilon)(\varepsilon^{-1})\). If we allowed this as a genuine factorization there would be no primes in \(R_\Delta\), so it is best not to consider it as a genuine factorization. This leads us to the following definition: An element \(\alpha\) of \(R_\Delta\) is said to be prime in \(R_\Delta\) if it is neither 0 nor a unit, and if whenever we have a factorization of \(\alpha\) as \(\alpha = \beta\gamma\) with both \(\beta\) and \(\gamma\) in \(R_\Delta\), then it must be the case that either \(\beta\) or \(\gamma\) is a unit in \(R_\Delta\). Not allowing units as primes is analogous to the standard practice of not considering 1 to be a prime in \(\mathbb{Z}\).

If we replace \(R_\Delta\) by \(\mathbb{Z}\) in the definition of primeness above, we get the condition that an integer \(a\) in \(\mathbb{Z}\) is prime if its only factorizations are the trivial ones \(a = (a)(1) = (1)(a)\) and \(a = (-a)(-1) = (-1)(-a)\), which is what we would expect. This definition of primeness also means that we are allowing negative primes as the negatives of the positive primes in \(\mathbb{Z}\).

A word of caution: An integer \(p\) in \(\mathbb{Z}\) can be prime in \(\mathbb{Z}\) but not prime in \(\mathbb{Z}[\sqrt{D}]\). For example, in \(\mathbb{Z}[i]\) we have the factorization \(5 = (2 + i)(2 - i)\), and as we will be able to verify soon, neither \(2 + i\) nor \(2 - i\) is a unit in \(\mathbb{Z}[i]\). Hence by our definition 5
is not a prime in \( \mathbb{Z}[i] \), even though it is prime in \( \mathbb{Z} \). Thus one always has to be careful when speaking about primeness to distinguish “prime in \( \mathbb{Z} \)” from “prime in \( R_\Delta \)”.

Having defined what we mean by primes in \( R_\Delta \) it is then natural to ask whether every nonzero element of \( R_\Delta \) that is not a unit can be factored as a product of primes, and if so, is this factorization in any way unique? As we will see, the existence of prime factorizations is fairly easy to prove, but the uniqueness question is much more difficult and subtle. To clarify what the uniqueness question means, notice first that if we have a unit \( \varepsilon \) in \( R_\Delta \) we can always modify a factorization \( \alpha = \beta \gamma \) to give another factorization \( \alpha = (\varepsilon \beta)(\varepsilon^{-1} \gamma) \). This is analogous to writing \( 6 = (2)(3) = (-2)(-3) \) in \( \mathbb{Z} \). This sort of nonuniqueness is unavoidable, but it is also not too serious a problem. So when we speak of factorization in \( R_\Delta \) being unique, we will always mean unique up to multiplying the factors by units.

A fruitful way to study factorizations in \( R_\Delta \) is to relate them to factorizations in \( \mathbb{Z} \) by means of the function \( N : R_\Delta \to \mathbb{Z} \) defined by \( N(\alpha) = \alpha \overline{\alpha} \). Thus in the two cases \( R_\Delta = \mathbb{Z}[\sqrt{D}] \) and \( R_\Delta = \mathbb{Z}[\omega] \) we have

\[
N(x + y\sqrt{D}) = (x + y\sqrt{D})(x - y\sqrt{D}) = x^2 - Dy^2 \\
N(x + y\omega) = (x + y\omega)(x + y\overline{\omega}) = x^2 + xy - dy^2
\]

The number \( N(\alpha) \) is called the norm of \( \alpha \). Notice that when the discriminant is negative, so \( \alpha \) is a complex number which can be written as \( a + bi \) for real numbers \( a \) and \( b \), the norm of \( \alpha \) is just \( \alpha \overline{\alpha} = (a + bi)(a - bi) = a^2 + b^2 \), the square of the distance from \( \alpha \) to the origin in the complex plane. When the discriminant is positive the norm can be negative so it does not have a nice geometric interpretation in terms of distance, but it will be quite useful in spite of this.

The reason the norm is useful for studying factorizations is that it satisfies the following multiplicative property:

**Proposition 8.1.** \( N(\alpha \beta) = N(\alpha)N(\beta) \) for all \( \alpha \) and \( \beta \) in \( R_\Delta \).

**Proof:** We will deduce multiplicativity of the norm from multiplicativity of the conjugation operation, the fact that \( \overline{\alpha \beta} = \overline{\alpha} \overline{\beta} \). The argument will apply more generally to all elements of \( \mathbb{Q}(\sqrt{D}) \) for any integer \( D \) that is not a square. To verify that \( \overline{\alpha \beta} = \overline{\alpha} \overline{\beta} \), write \( \alpha = x + y\sqrt{D} \) and \( \beta = z + w\sqrt{D} \), so that \( \alpha \beta = (xz + ywD) + (xw + yz)\sqrt{D} \). Then

\[
\overline{\alpha \beta} = (xz + ywD) - (xw + yz)\sqrt{D} = (x - y\sqrt{D})(z - w\sqrt{D}) = \overline{\alpha} \overline{\beta}
\]

Now for the norm we have \( N(\alpha \beta) = (\alpha \beta)(\overline{\alpha \beta}) = \alpha \beta \overline{\alpha} \overline{\beta} = \alpha \overline{\alpha} \beta \overline{\beta} = N(\alpha)N(\beta) \)

Using the multiplicative property of the norm we can derive a simple criterion for recognizing units:
Proposition 8.2. An element \( \varepsilon \in R_\Delta \) is a unit if and only if \( N(\varepsilon) = \pm 1 \).

Proof: Suppose \( \varepsilon \) is a unit, so its inverse \( \varepsilon^{-1} \) also lies in \( R_\Delta \). Then \( N(\varepsilon)N(\varepsilon^{-1}) = N(\varepsilon\varepsilon^{-1}) = N(1) = 1 \). Since both \( N(\varepsilon) \) and \( N(\varepsilon^{-1}) \) are elements of \( \mathbb{Z} \), this forces \( N(\varepsilon) \) to be \( \pm 1 \). For the converse, the inverse of an element \( \varepsilon \) in \( R_\Delta \) is \( \varepsilon^{-1} = \varepsilon/N(\varepsilon) \) since multiplying this by \( \varepsilon \) gives 1. Hence if \( N(\varepsilon) = \pm 1 \) we have \( \varepsilon^{-1} = \pm \varepsilon \) which is an element of \( R_\Delta \) if \( \varepsilon \) is, so \( \varepsilon \) is a unit.

When \( \Delta \) is negative there are very few units in \( R_\Delta \). In the case of \( \mathbb{Z}[\sqrt{\Delta}] \) the equation \( N(x + y\sqrt{\Delta}) = x^2 - Dy^2 = \pm 1 \) has very few integer solutions, namely, if \( D = -1 \) there are only the four solutions \( (x, y) = (\pm 1, 0) \) and \( (0, \pm 1) \) while if \( D < -1 \) there are only the two solutions \( (x, y) = (\pm 1, 0) \). Thus the only units in \( \mathbb{Z}[i] \) are \( \pm 1 \) and \( \pm i \), and the only units in \( \mathbb{Z}[\sqrt{\Delta}] \) for \( D < -1 \) are \( \pm 1 \). In the case of \( \mathbb{Z}[\omega] \) one can see from the earlier figure of \( \mathbb{Z}[\omega] \) when \( d = -1 \) that there are six lattice points of distance 1 from the origin, giving the six units \( \pm 1, \pm \omega, \) and \( \pm (\omega - 1) \). These are the powers \( \omega^n \) for \( n = 0, 1, 2, 3, 4, 5 \) since \( \omega^2 = \omega - 1 \) and \( \omega^3 = -1 \), hence \( \omega^6 = 1 \). When \( d < -1 \) the only units in \( \mathbb{Z}[\omega] \) are \( \pm 1 \) since the lattice is stretched vertically so there are only two lattice points of distance 1 from the origin.

The situation for \( R_\Delta \) with \( \Delta \) positive is quite different. For \( \mathbb{Z}[\sqrt{\Delta}] \) we are looking for solutions of \( x^2 - Dy^2 = \pm 1 \) with \( D > 0 \), while for \( \mathbb{Z}[\omega] \) the corresponding equation is \( x^2 + xy - dy^2 = \pm 1 \) with \( d > 0 \). We know from our study of topographs of hyperbolic forms that these equations have infinitely many integer solutions since the value 1 occurs along the periodic separator line in the topograph of the principal form when \( (x, y) = (1, 0) \), so it appears infinitely often by periodicity. For some values of \( D \) or \( d \) the number \( -1 \) also appears along the separator line, and then it too appears infinitely often. Thus when \( \Delta > 0 \) the ring \( R_\Delta \) has infinitely many units \( \varepsilon = x + y\sqrt{\Delta} \) or \( x + y\omega \), with arbitrarily large values of \( x \) and \( y \).

There is a nice interpretation of units in \( R_\Delta \) as symmetries of the topograph of the principal form of discriminant \( \Delta \), as we shall now describe. A unit \( \varepsilon \) in \( R_\Delta \) defines a transformation \( T_\varepsilon \) of \( R_\Delta \) by the formula \( T_\varepsilon(\alpha) = \varepsilon\alpha \). In the case of \( \mathbb{Z}[\sqrt{\Delta}] \), if \( \varepsilon = p + q\sqrt{\Delta} \) then

\[
T_\varepsilon(x + y\sqrt{\Delta}) = (p + q\sqrt{\Delta})(x + y\sqrt{\Delta}) = (px + Dqy) + (qx + p\sqrt{\Delta})
\]

while for \( \mathbb{Z}[\omega] \), if \( \varepsilon = p + q\omega \) we have

\[
T_\varepsilon(x + y\omega) = (p + q\omega)(x + y\omega) = (px + qy\omega^2) + (qx + py)\omega
\]

\[
= (px + dqy) + (qx + (p + q)y)\omega
\]

since \( \omega^2 = \omega - d \). In both cases we see that \( T_\varepsilon \) is a linear transformation of \( x \) and \( y \), with matrix \( \begin{pmatrix} p & Dq \\ q & p \end{pmatrix} \) in the first case and \( \begin{pmatrix} p & dq \\ q & p + q \end{pmatrix} \) in the second case. The determinants in the two cases are \( p^2 - Dq^2 \) and \( p^2 + pq - dq^2 \) which equal \( N(\varepsilon) \) and hence are \( \pm 1 \) since \( \varepsilon \) is a unit. Thus \( T_\varepsilon \) defines a linear fractional transformation giving a symmetry.
of the Farey diagram. Since \(N(\varepsilon \alpha) = N(\varepsilon)N(\alpha)\) we see that \(T_\varepsilon\) is an orientation-preserving symmetry of the topograph of the norm form when \(N(\varepsilon) = +1\) and an orientation-reversing skew symmetry when \(N(\varepsilon) = -1\). The symmetry corresponding to the 'universal' unit \(\varepsilon = -1\) is just the identity since \(\varepsilon = \frac{x}{y}\).

When \(\Delta < 0\) the only interesting cases are \(\Delta = -3\), when \(T_\varepsilon\) for \(\varepsilon = \omega\) is a 120 degree rotation of the topograph, and \(\Delta = -4\) when \(T_\varepsilon\) for \(\varepsilon = i\) rotates the topograph by 180 degrees.

When \(\Delta > 0\) there is a fundamental unit \(\varepsilon\) corresponding to the \(\pm 1\) in the topograph of the norm form at the vertex \(p/q\) with smallest positive values of \(p\) and \(q\). When \(N(\varepsilon) = +1\) the transformation \(T_\varepsilon\) is then the translation giving the periodicity along the separating line since it is an orientation-preserving symmetry. In the opposite case \(N(\varepsilon) = -1\) the transformation \(T_\varepsilon\) is an orientation-reversing skew symmetry so it must be a glide reflection along the separator line by half a period.

**Proposition 8.3.** If \(\Delta > 0\) then the units in \(R_\Delta\) are the elements \(\pm \varepsilon^n\) for \(n \in \mathbb{Z}\), where \(\varepsilon\) is the fundamental unit.

**Proof:** The units appear along the separator line at the regions \(x/y\) where the norm form takes the value \(\pm 1\). From our previous comments, these are the points \(T_\varepsilon^n(1/0)\) as \(n\) varies over \(\mathbb{Z}\). Since \(T_\varepsilon\) is multiplication by \(\varepsilon\), the power \(T_\varepsilon^n\) is multiplication by \(\varepsilon^n\). Thus the values \(\pm 1\) occur at the regions labeled \(x/y\) for \(\varepsilon^n = x + y\sqrt{D}\) or \(\varepsilon^n = x + y\omega\). The units are therefore the elements \(\pm \varepsilon^n\) where the \(\pm\) comes from the fact that the topograph does not distinguish between \((x, y)\) and \((-x, -y)\). \(\Box\)

The conjugation operation in \(R_\Delta\) sending \(\alpha\) to \(\overline{\alpha}\) also gives a symmetry of the topograph of the norm form since \(N(\alpha) = N(\overline{\alpha})\). Conjugation in \(\mathbb{Z}[\sqrt{D}]\) sends \(x + y\sqrt{D}\) to \(x - y\sqrt{D}\) so in the Farey diagram it is reflection across the edge joining 1/0 and 0/1. Conjugation in \(\mathbb{Z}[\omega]\) sends \(x + y\omega\) to \(x + y\overline{\omega} = (x + y) - \omega\) since \(\overline{\omega} = 1 - \omega\), so conjugation fixes the vertex 1/0 and interchanges 0/1 and \(-1/1\) by reflecting across the line perpendicular to the edge from 0/1 to \(-1/1\).

**Proposition 8.4.** All symmetries and skew symmetries of the topograph of the norm form are obtainable as combinations of conjugation and the transformations \(T_\varepsilon\) associated to units \(\varepsilon\) in \(R_\Delta\).

**Proof:** It will suffice to reduce an arbitrary symmetry or skew symmetry \(T\) to the identity by composing with conjugation and transformations \(T_\varepsilon\). If \(T\) is a skew symmetry we must have \(\Delta > 0\) with \(-1\) appearing along the separator line as well as \(+1\). Composing \(T\) with a glide reflection \(T_\varepsilon\) then converts \(T\) into a symmetry, so we may assume \(T\) is a symmetry from now on. If \(T\) reverses orientation of the Farey diagram we may compose it with conjugation to reduce further to the case that \(T\) preserves orientation. When \(\Delta < 0\) the only possibility for \(T\) is then the identity except when \(\Delta = -4\) and \(T = T_\varepsilon\) for \(\varepsilon = i\), or when \(\Delta = -3\) and \(T = T_\varepsilon\) for \(\varepsilon = \omega\) or \(\omega^2\). If
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$\Delta > 0$ the only possibility for $T$ is a translation along the separator line, which is $T_\varepsilon$ for some unit $\varepsilon$.

Now we begin to study primes and prime factorizations in $R_\Delta$. First we have a useful fact:

**Proposition 8.5.** If the norm $N(\alpha)$ of an element $\alpha$ in $R_\Delta$ is prime in $\mathbb{Z}$ then $\alpha$ is prime in $R_\Delta$.

For example, when we factor $5$ as $(2+i)(2-i)$ in $\mathbb{Z}[i]$, this proposition implies that both factors are prime since the norm of each is $5$, which is prime in $\mathbb{Z}$.

**Proof:** Suppose an element $\alpha$ in $R_\Delta$ has a factorization $\alpha = \beta \gamma$, hence $N(\alpha) = N(\beta)N(\gamma)$. If $N(\alpha)$ is prime in $\mathbb{Z}$, this forces one of $N(\beta)$ and $N(\gamma)$ to be $\pm 1$, hence one of $\beta$ and $\gamma$ is a unit. This means $\alpha$ is a prime since it cannot be $0$ or a unit, as its norm is a prime.

The converse of this proposition is not generally true. For example the number $3$ has norm $9$, which is not prime in $\mathbb{Z}$, and yet $3$ is prime in $\mathbb{Z}[i]$ since if we had a factorization $3 = \alpha \beta$ in $\mathbb{Z}[i]$ with neither $\alpha$ nor $\beta$ a unit, then the equation $N(\alpha)N(\beta) = N(3) = 9$ would imply that $N(\alpha) = \pm 3 = N(\beta)$, but there are no elements of $\mathbb{Z}[i]$ with norm $\pm 3$ since the equation $x^2 + y^2 = \pm 3$ has no integer solutions.

Now we can prove that prime factorizations always exist:

**Proposition 8.6.** Every nonzero element of $R_\Delta$ that is not a unit can be factored as a product of primes in $R_\Delta$.

**Proof:** We argue by induction on $|N(\alpha)|$. Since we are excluding $0$ and units, the induction starts with the case $|N(\alpha)| = 2$. In this case $\alpha$ must itself be a prime by the preceding proposition since $2$ is prime in $\mathbb{Z}$. For the induction step, if $\alpha$ is a prime there is nothing to prove. If $\alpha$ is not prime, it factors as $\alpha = \beta \gamma$ with neither $\beta$ nor $\gamma$ a unit, so $|N(\beta)| > 1$ and $|N(\gamma)| > 1$. Since $N(\alpha) = N(\beta)N(\gamma)$, it follows that $|N(\beta)| < |N(\alpha)|$ and $|N(\gamma)| < |N(\alpha)|$. By induction, both $\beta$ and $\gamma$ are products of primes in $R_\Delta$, hence their product $\alpha$ is also a product of primes.

Let us investigate how to compute a prime factorization by looking at the case of $\mathbb{Z}[i]$. Assuming that factorizations of Gaussian integers into primes are unique (up to units), which we will prove later, here is a procedure for finding the prime factorization of a Gaussian integer $\alpha = a + bi$:

1. Factor the integer $N(\alpha) = a^2 + b^2$ into primes $p_k$ in $\mathbb{Z}$.
2. Determine how each $p_k$ factors into primes in $\mathbb{Z}[i]$.
3. By the uniqueness of prime factorizations, the primes found in step (2) will be factors of either $a + bi$ or $a - bi$ since they are factors of $(a + bi)(a - bi)$, so all that remains is to test which of the prime factors of each $p_k$ are factors of $a + bi$.
To illustrate this with a simple example, let us see how $3 + i$ factors in $\mathbb{Z}[i]$. We have $N(3 + i) = (3 + i)(3 - i) = 10 = 2 \cdot 5$. These two numbers factor as $2 = (i + i)(1 - i)$ and $5 = (2 + i)(2 - i)$. These are prime factorizations in $\mathbb{Z}[i]$ since $N(1 \pm i) = 2$ and $N(2 \pm i) = 5$, both of which are primes in $\mathbb{Z}$. Now we test whether for example $1 + i$ divides $3 + i$ by dividing:

$$\frac{3 + i}{1 + i} = \frac{(3 + i)(1 - i)}{(1 + i)(1 - i)} = \frac{4 - 2i}{2} = 2 - i$$

Since the quotient $2 - i$ is a Gaussian integer, we conclude that $1 + i$ is a divisor of $3 + i$ and we have the factorization $3 + i = (1 + i)(2 - i)$. This is the prime factorization of $3 + i$ since we have already noted that both $1 + i$ and $2 - i$ are primes in $\mathbb{Z}[i]$.

For a more complicated example consider $244 + 158i$. For a start, this factors as $2(122 + 79i)$. Since 122 and 79 have no common factors in $\mathbb{Z}$ we can't go any farther by factoring out ordinary integers. We know that 2 factors as $(1 + i)(1 - i)$ and these two factors are prime in $\mathbb{Z}[i]$ since their norm is 2. It remains to factor $122 + 79i$. This has norm $122^2 + 79^2 = 21125 = 5^3 \cdot 13^2$. Both 5 and 13 happen to factor in $\mathbb{Z}[i]$, namely $5 = (2 + i)(2 - i)$ and $13 = (3 + 2i)(3 - 2i)$, and these are prime factorizations since the norms of $2 \pm i$ and $3 \pm 2i$ are 5 and 13, primes in $\mathbb{Z}$. Thus we have the prime factorization

$$(122 + 79i)(122 - 79i) = 5^3 \cdot 13^2 = (2 + i)^3(2 - i)^3(3 + 2i)(3 - 2i)^2$$

Now we look at the factors on the right side of this equation to see which ones are factors of $122 + 79i$. Suppose for example we test whether $2 + i$ divides $122 + 79i$:

$$\frac{122 + 79i}{2 + i} = \frac{(122 + 79i)(2 - i)}{(2 + i)(2 - i)} = \frac{323 + 36i}{5}$$

This is not a Gaussian integer, so $2 + i$ does not divide $122 + 79i$. Let's try $2 - i$ instead:

$$\frac{122 + 79i}{2 - i} = \frac{(122 + 79i)(2 + i)}{(2 - i)(2 + i)} = \frac{165 + 280i}{5} = 33 + 56i$$

So $2 - i$ does divide $122 + 79i$. In fact, we can expect that $(2 - i)^3$ will divide $122 + 79i$, and it can be checked that it does. In a similar way one can check whether $3 + 2i$ or $3 - 2i$ divides $122 + 79i$, and one finds that it is $3 - 2i$ that divides $122 + 79i$, and in fact $(3 - 2i)^2$ divides $122 + 79i$. After these calculations one might expect that $122 + 79i$ was the product $(2 - i)^3(3 - 2i)^2$, but upon multiplying this product out one finds that it is the negative of $122 + 79i$, so

$$122 + 79i = (-1)(2 - i)^3(3 - 2i)^2$$

The factor $-1$ is a unit, so it could be combined with one of the other factors, for example changing one of the factors $2 - i$ to $i - 2$. Alternatively, we could replace the factor $-1$ by $i^2$ and then multiply each $3 - 2i$ factor by $i$ to get the prime factorization

$$122 + 79i = (2 - i)^3(2 + 3i)^2$$
Hence for $244 + 158i$ we have the prime factorization
\[ 244 + 158i = (1 + i)(1 - i)(2 - i)^3(2 + 3i)^2 \]

The method in this example for computing prime factorizations in $\mathbb{Z}[i]$ depended on unique factorization. When unique factorization fails, things are more complicated. One of the simplest instances of this is in $\mathbb{Z}[\sqrt{-5}]$ where we have the factorizations
\[ 6 = (2)(3) = (1 + \sqrt{-5})(1 - \sqrt{-5}) \]

The only units in $\mathbb{Z}[\sqrt{-5}]$ are $\pm 1$, so these two factorizations do not differ just by units. We can see that $2$, $3$, and $1 \pm \sqrt{-5}$ are prime in $\mathbb{Z}[\sqrt{-5}]$ by looking at norms. Using the formula $N(x + y\sqrt{-5}) = x^2 + 5y^2$ we see that the norms of $2$, $3$, and $1 \pm \sqrt{-5}$ are $4$, $9$, and $6$, so if one of $2$, $3$, or $1 \pm \sqrt{-5}$ was not a prime, it would have a factor of norm $2$ or $3$ since these are the only numbers that occur in nontrivial factorizations of $4$, $9$, and $6$ in $\mathbb{Z}$. However, the equations $x^2 + 5y^2 = 2$ and $x^2 + 5y^2 = 3$ obviously have no integer solutions so there are no elements of $\mathbb{Z}[\sqrt{-5}]$ of norm $2$ or $3$. Thus in $\mathbb{Z}[\sqrt{-5}]$ the number $6$ has two prime factorizations that do not differ merely by units.

What is secretly going on in this example is that $x^2 + 5y^2$ is not the only quadratic form of discriminant $-20$, up to equivalence. Another form of the same discriminant is $2x^2 + 2xy + 3y^2$, and this form takes on the values 2 and 3 that the form $x^2 + 5y^2$ omits, even though $x^2 + 5y^2$ does take on the value $6 = 2 \cdot 3$. Here are the topographs of these two forms, with prime values circled.
The boxed nonprime values in the topograph of \( x^2 + 5y^2 \) give rise to nonunique prime factorizations like the two factorizations of 6 given above. For example 14 = \((2)(7) = (3 + \sqrt{-5})(3 - \sqrt{-5})\). Some numbers occur in boxes twice, leading to three different prime factorizations. Thus 21 factors into primes in \( \mathbb{Z}[\sqrt{-5}] \) as \( 3 \cdot 7 \), as \((1 + 2\sqrt{-5})(1 - 2\sqrt{-5})\) and as \((4 + \sqrt{-5})(4 - \sqrt{-5})\). Another example is 69 = \( 3 \cdot 23 = (7 + 2\sqrt{-5})(7 - 2\sqrt{-5}) = (8 + \sqrt{-5})(8 - \sqrt{-5})\).

The question of how a prime \( p \) in \( \mathbb{Z} \) factors in \( R_\Delta \) can be rephrased in terms of the norm form \( x^2 - 5y^2 \) or \( x^2 + xy - 5y^2 \), according to the following result:

**Proposition 8.7.** Let \( p \) be a prime in \( \mathbb{Z} \). Then:

(a) If either \( p \) or \( -p \) is represented by the norm form for \( R_\Delta \), so \( N(\alpha) = \pm p \) for some \( \alpha \) in \( R_\Delta \), then \( p \) factors in \( R_\Delta \) as \( p = \alpha \overline{\alpha} \) and both these factors are prime in \( R_\Delta \).

(b) If neither \( p \) nor \( -p \) is represented by the norm form then \( p \) remains prime in \( R_\Delta \).

In case (a) one says the prime \( p \) splits in \( R_\Delta \). In case (b) the prime \( p \) is said to be inert.

In statement (a) note that when \( \Delta < 0 \) the norm only takes positive values, so if a positive prime \( p \) factors in \( R_\Delta \) it must factor as \( p = \alpha \overline{\alpha} \), never as \( -\alpha \overline{\alpha} \). However for \( \Delta > 0 \) this need not be the case. For example for \( \mathbb{Z}[\sqrt{3}] \) the topograph of \( x^2 - 3y^2 \) (shown in Chapter 4) contains the value \( -2 \) but not 2, so the prime 2 factors as \( -(1 + \sqrt{3})(1 - \sqrt{3}) \) in \( \mathbb{Z}[\sqrt{3}] \) but not as \( \alpha \overline{\alpha} \).

**Proof:** For part (a), if \( p = \pm N(\alpha) \), then \( p \) factors in \( R_\Delta \) as \( p = \pm \alpha \overline{\alpha} = \pm N(\alpha) \). The two factors are prime since their norm is \( \pm p \) which is prime in \( \mathbb{Z} \) by assumption.

For (b), if \( p \) is not a prime in \( R_\Delta \) then it factors in \( R_\Delta \) as \( p = \alpha \beta \) with neither \( \alpha \) nor \( \beta \) a unit. Then \( N(p) = p^2 = N(\alpha)N(\beta) \) with neither \( N(\alpha) \) nor \( N(\beta) \) equal to \( \pm 1 \), hence we must have \( N(\alpha) = N(\beta) = \pm p \). The equation \( N(\alpha) = \pm p \) says that the norm form represents \( \pm p \). Thus if the norm form represents neither \( p \) nor \( -p \) then \( p \) must be prime in \( R_\Delta \).

**Proposition 8.8.** If \( R_\Delta \) has unique factorization into primes then the only primes in \( R_\Delta \) are the primes described in (a) or (b) of the preceding proposition (or units times these primes).

**Proof:** Let \( \alpha \) be an arbitrary prime in \( R_\Delta \). The norm \( n = N(\alpha) = \alpha \overline{\alpha} \) is an integer in \( \mathbb{Z} \) so it can be factored as a product \( n = p_1 \cdots p_k \) of primes in \( \mathbb{Z} \). By the preceding proposition each \( p_i \) either stays prime in \( R_\Delta \) or factors as a product \( \pm \alpha_i \overline{\alpha}_i \) of two primes in \( R_\Delta \). This gives a factorization of \( n \) into primes in \( R_\Delta \). A second factorization of \( n \) into primes in \( R_\Delta \) can be obtained from the formula \( n = \alpha \overline{\alpha} \) by factoring \( \overline{\alpha} \) into primes since the first factor \( \alpha \) is already prime by assumption. (In fact if \( \alpha \) is prime then \( \overline{\alpha} \) will also be a prime, but we don’t need to know this.) If we have
unique factorization in $R_\Delta$ then the prime factor $\alpha$ of the second prime factorization will have to be one of the prime factors in the first prime factorization of $n$, or a unit times one of these primes. Thus $\alpha$ will be a unit times a prime of one of the two types described in the previous proposition.

We have seen that prime factorizations in $R_\Delta$ may not be unique (even up to units) but there is one special situation in which they are:

**Proposition 8.9.** The factorization of a prime $p$ in $\mathbb{Z}$ into primes in $R_\Delta$ is always unique up to units.

**Proof:** If $p$ is prime in $R_\Delta$ then uniqueness is automatic. If $p$ is not prime in $R_\Delta$ then it has a prime factorization $p = \pm \alpha \overline{\alpha}$ and we want to show that any other prime factorization is the same as this one, up to units. Note that a prime factorization of $p$ in $R_\Delta$ cannot have more than two factors since $N(p) = p^2$ is the product of two primes in $\mathbb{Z}$.

Suppose that we have a factorization $p = \beta \gamma$ with $\beta$ and $\gamma$ prime in $R_\Delta$. Both $\beta$ and $\gamma$ have norm $\pm p$, so in the topograph of the norm form the number $p$ or $-p$ appears in the regions corresponding to $\beta$ and $\gamma$ as well as $\alpha$ and $\overline{\alpha}$. By Proposition 6.3 there is a symmetry or skew symmetry of the topograph taking the $\alpha$ region to the $\beta$ region. By Proposition 8.4 this symmetry or skew symmetry is realizable by a combination of conjugation and multiplication by units, hence $\beta$ is a unit times $\alpha$ or $\overline{\alpha}$. Interchanging $\alpha$ and $\overline{\alpha}$ if necessary, we may assume that $\beta$ is a unit times $\alpha$. The equation $\pm \alpha \overline{\alpha} = \beta \gamma$ then implies that $\gamma$ is a unit times $\overline{\alpha}$. Thus the two factorizations differ only by units.

There are two qualitatively different ways in which a prime $p$ can factor as the product of two primes in $R_\Delta$. The distinction is illustrated by the factorizations $2 = (1 + i)(1 - i)$ and $5 = (2 + i)(2 - i)$ in $\mathbb{Z}[i]$. The two factors $1 + i$ and $1 - i$ differ only by multiplication by a unit since $-i(1 + i) = 1 - i$. However, $2 + i$ and $2 - i$ do not differ just by a unit since multiplying $2 + i$ by the units $i$, $-1$, and $-i$ gives $-1 + 2i$, $-2 - i$, and $1 - 2i$, none of which equals $2 - i$. The terminology usually used for this distinction is to say that $2$ is ramified in $\mathbb{Z}[i]$ while $5$ is unramified in $\mathbb{Z}[i]$.

In fact $2$ is the only prime that is ramified in $\mathbb{Z}[i]$, for if $p = (a + bi)(a - bi)$ with $a - bi = \pm (a + bi)$ we must have $a = 0$ or $b = 0$, so $p = a^2$ or $p = b^2$ which is impossible if $p$ is prime, and if $a - bi = \pm i(a + bi)$ then $a = \pm b$ so $p = (a \pm ai)(a \mp ai) = a^2 \cdot 2$ and hence $a = \pm 1$ and $p = 2$.

2. Unique Factorization via the Euclidean Algorithm

Our goal now is to show that unique factorization holds for the Gaussian integers $\mathbb{Z}[i]$, and in a few other cases as well. The plan will be to see that Gaussian integers have a Euclidean algorithm much like the Euclidean algorithm in $\mathbb{Z}$, then deduce unique factorization from this Euclidean algorithm.
In order to prove that prime factorizations are unique we will use the following special property that holds in \( \mathbb{Z} \) and in some of the rings \( R_\Delta \) as well:

\((\star)\) \textit{If a prime } \( p \) \textit{divides a product } \( ab \) \textit{then } \( p \) \textit{must divide either } \( a \) \textit{or } \( b \).

One way to prove this for \( \mathbb{Z} \) would be to consider the prime factorization of \( ab \), which can be obtained by factoring each of \( a \) and \( b \) into primes separately. Then if the prime \( p \) divides \( ab \), it would have to occur in the prime factorization of \( ab \), hence it would occur in the prime factorization of either \( a \) or \( b \), which would say that \( p \) divides \( a \) or \( b \).

This argument assumed implicitly that the prime factorization of \( ab \) was unique. Thus the property \((\star)\) is a consequence of unique factorization into primes. But the property \((\star)\) also implies that prime factorizations are unique. To see why, consider two factorizations of a number \( n > 1 \) into positive primes:

\[
n = p_1 p_2 \cdots p_k = q_1 q_2 \cdots q_l
\]

We can assume \( k \leq l \) by interchanging the \( p_i \)'s and \( q_i \)'s if necessary. We want to argue that if \((\star)\) holds for each \( p_i \), then the \( q_i \)'s are just a permutation of the \( p_i \)'s and in particular \( k = l \). The argument to prove this goes as follows. Consider first the prime \( p_1 \). This divides the product \( q_1 (q_2 \cdots q_l) \) so by property \((\star)\) it divides either \( q_1 \) or \( q_2 q_3 \cdots q_l \). In the latter case, another application of \((\star)\) shows that \( p_1 \) divides either \( q_2 \) or \( q_3 q_4 \cdots q_l \). Repeating this argument as often as necessary, we conclude that \( p_1 \) must divide at least one \( q_i \). After permuting the \( q_i \)'s we can assume that \( p_1 \) divides \( q_1 \). We are assuming all the \( p_i \)'s and \( q_i \)'s are positive, so the fact that the prime \( p_1 \) divides the prime \( q_1 \) implies that \( p_1 \) equals \( q_1 \). We can then cancel \( p_1 \) and \( q_1 \) from the equation \( p_1 p_2 \cdots p_k = q_1 q_2 \cdots q_l \) to get \( p_2 \cdots p_k = q_2 \cdots q_l \). Now repeat the argument to show that \( p_2 \) equals some remaining \( q_i \) which we can assume is \( q_2 \) after a permutation. After further repetitions we eventually reach the point that the final \( p_k \) is a product of the remaining \( q_i \)'s. But then since \( p_k \) is prime there could only be one remaining \( q_i \), so we would have \( k = l \) and \( p_k = q_k \), finishing the argument.

If we knew the analog of property \((\star)\) held for primes in \( R_\Delta \) we could make essentially the same argument to show that unique factorization holds in \( R_\Delta \). The only difference in the argument would be that we would have to take units into account. The argument would be exactly the same up to the point where we concluded that \( p_1 \) divides \( q_1 \). Then the fact that \( q_1 \) is prime would not say that \( p_1 \) and \( q_1 \) were equal, but only that \( q_1 \) is a unit times \( p_1 \), so we would have an equation \( q_1 = ep_1 \) with \( e \) a unit. Then we would have \( p_1 p_2 \cdots p_k = ep_1 q_2 \cdots q_l \). Canceling \( p_1 \) would then yield \( p_2 p_3 \cdots p_k = eq_2 q_3 \cdots q_l \). The product \( eq_2 \) is prime if \( q_2 \) is prime, so if we let \( q'_2 = eq_2 \) we would have \( p_2 p_3 \cdots p_k = q'_2 q_3 \cdots q_l \). The argument could then be repeated to show eventually that the \( q_i \)'s are the same as the \( p_i \)'s up to permutation and multiplication by units, which is what unique factorization means.
Since the property \((\ast)\) implies unique factorization, it will not hold in \(R_\Delta\) when \(R_\Delta\) does not have unique factorization. For a concrete example consider \(\mathbb{Z}[\sqrt{-5}]\). Here we had nonunique prime factorizations \(6 = 2 \cdot 3 = (1 + \sqrt{-5})(1 - \sqrt{-5})\). The prime 2 thus divides the product \((1 + \sqrt{-5})(1 - \sqrt{-5})\) but it does not divide either factor \(1 \pm \sqrt{-5}\) since \((1 \pm \sqrt{-5})/2\) is not an element of \(\mathbb{Z}[\sqrt{-5}]\).

Our task now is to prove the property \((\ast)\) without using unique factorization. As we saw in Chapter 2, an equation \(ax + by = 1\) always has integer solutions \((x, y)\) whenever \(a\) and \(b\) are coprime integers. This fact can be used to show that property \((\ast)\) holds in \(\mathbb{Z}\). To see how, suppose that a prime \(p\) divides a product \(ab\). It will suffice to show that if \(p\) does not divide \(a\) then it must divide \(b\). If \(p\) does not divide \(a\), then since \(p\) is prime, \(p\) and \(a\) are coprime. This implies that the equation \(px + ay = 1\) is solvable with integers \(x\) and \(y\). Now multiply this equation by \(b\) to get an equation \(b = pbx + aby\). The number \(p\) divides the right side of this equation since it obviously divides \(pbx\) and it divides \(ab\) by assumption. Hence \(p\) divides \(b\), which is what we wanted to show.

The fact that equations \(ax + by = 1\) in \(\mathbb{Z}\) are solvable whenever \(a\) and \(b\) are coprime can be deduced from the Euclidean algorithm, in the following way. What the Euclidean algorithm gives is a method for starting with two positive integers \(\alpha_0\) and \(\alpha_1\) and constructing a sequence of positive integers \(\alpha_i\) and \(\beta_i\) satisfying the equations

\[
\alpha_0 = \beta_1 \alpha_1 + \alpha_2 \\
\alpha_1 = \beta_2 \alpha_2 + \alpha_3 \\
\vdots \\
\alpha_{n-2} = \beta_{n-1} \alpha_{n-1} + \alpha_n \\
\alpha_{n-1} = \beta_n \alpha_n + \alpha_{n+1} \\
\alpha_n = \beta_{n+1} \alpha_{n+1}
\]

From these equations we can deduce two consequences:

1. \(\alpha_{n+1}\) divides \(\alpha_0\) and \(\alpha_1\).
2. The equation \(\alpha_{n+1} = \alpha_0 x + \alpha_1 y\) is solvable in \(\mathbb{Z}\).

To see why (1) is true, note that the last equation implies that \(\alpha_{n+1}\) divides \(\alpha_n\). Then the next-to-last equation implies that \(\alpha_{n+1}\) divides \(\alpha_{n-1}\), and the equation before this then implies that \(\alpha_{n+1}\) divides \(\alpha_{n-2}\), and so on until one deduces that \(\alpha_{n+1}\) divides all the \(\alpha_i\)'s and in particular \(\alpha_0\) and \(\alpha_1\).

To see why (2) is true, observe that each equation before the last one allows an \(\alpha_i\) to be expressed as a linear combination of \(\alpha_{i-1}\) and \(\alpha_{i-2}\), so by repeatedly substituting in, one can express each \(\alpha_i\) in terms of \(\alpha_0\) and \(\alpha_1\) as a linear combination \(x\alpha_0 + y\alpha_1\) with integer coefficients \(x\) and \(y\), so in particular \(\alpha_{n+1}\) can be represented in this way, which says that the equation \(\alpha_{n+1} = \alpha_0 x + \alpha_1 y\) is solvable in \(\mathbb{Z}\).
Now if we assume that \( \alpha_0 \) and \( \alpha_1 \) are coprime then \( \alpha_{n+1} \) must be 1 by statement (1), and by statement (2) we get integers \( x \) and \( y \) such that \( \alpha_0 x + \alpha_1 y = 1 \), as we wanted.

Putting all the preceding arguments together, we see that the Euclidean algorithm in \( \mathbb{Z} \) implies unique factorization in \( \mathbb{Z} \).

A very similar argument works in \( R_\Delta \) provided that one has a Euclidean algorithm to produce the sequence of equations above starting with any nonzero pair of elements \( \alpha_0 \) and \( \alpha_1 \) in \( R_\Delta \). The only difference in the more general case is that \( \alpha_{n+1} \) might not be 1, but only a unit in \( R_\Delta \). Thus one would apply statements (1) and (2) to a pair \( \alpha_0, \alpha_1 \) whose only common divisors were units, hence \( \alpha_{n+1} \) would be a unit, and then the equation \( \alpha_{n+1} = \alpha_0 x + \alpha_1 y \) could be modified by multiplying through by \( \alpha_{n+1}^{-1} \) to get an equation \( 1 = \alpha_0 x + \alpha_1 y \) with a solution \( x, y \) in \( R_\Delta \). As we have seen, this would imply unique factorization in \( R_\Delta \).

Let us show now that there is a Euclidean algorithm in the Gaussian integers \( \mathbb{Z}[i] \). The key step is to be able to find, for each pair of nonzero Gaussian integers \( \alpha_0 \) and \( \alpha_1 \), two more Gaussian integers \( \beta_1 \) and \( \alpha_2 \) such that \( \alpha_0 = \beta_1 \alpha_1 + \alpha_2 \) with \( \alpha_2 \) being ‘smaller’ than \( \alpha_1 \). We measure ‘smallness’ of complex numbers by computing their distance to the origin in the complex plane. For a complex number \( \alpha = x + yi \) this distance is \( \sqrt{x^2 + y^2} \). Here \( x^2 + y^2 \) is just the norm \( N(\alpha) \) when \( x \) and \( y \) are integers, so we could measure the size of a Gaussian integer \( \alpha \) by \( \sqrt{N(\alpha)} \). However it is simpler just to use \( N(\alpha) \) without the square root, so this is what we will do.

Thus our goal is to find an equation \( \alpha_0 = \beta_1 \alpha_1 + \alpha_2 \) with \( N(\alpha_2) < N(\alpha_1) \), starting from two given nonzero Gaussian integers \( \alpha_0 \) and \( \alpha_1 \). If we can always do this, then by repeating the process we can construct a sequence of \( \alpha_i \)’s and \( \beta_i \)’s where the successive \( \alpha_i \)’s have smaller and smaller norms. Since these norms are positive integers, they cannot keep decreasing infinitely often, so eventually the process will reach an \( \alpha_i \) of norm 0, hence this \( \alpha_i \) will be 0 and the Euclidean algorithm will end in a finite number of steps, as it should.

The equation \( \alpha_0 = \beta_1 \alpha_1 + \alpha_2 \) is saying that when we divide \( \alpha_1 \) into \( \alpha_0 \), we obtain a quotient \( \beta_1 \) and a remainder \( \alpha_2 \). What we want is for the remainder \( \alpha_2 \) to have a smaller norm than \( \alpha_1 \). To get an idea how we can do this let us look instead at the equivalent equation

\[
\frac{\alpha_0}{\alpha_1} = \beta_1 + \frac{\alpha_2}{\alpha_1}
\]

If we were working with ordinary integers, the quotient \( \beta_1 \) would be the integer part of the rational number \( \alpha_0/\alpha_1 \) and \( \alpha_2/\alpha_1 \) would be the remaining fractional part. For Gaussian integers we do something similar, but instead of taking \( \beta_1 \) to be the ‘integer part’ of \( \alpha_0/\alpha_1 \) we take it to be the closest Gaussian integer to \( \alpha_0/\alpha_1 \).

Here is an example, where we choose \( \alpha_0 \) to be \( 12 + 15i \) and \( \alpha_1 \) to be \( 5 + 2i \).
Then:
\[
\frac{\alpha_0}{\alpha_1} = \frac{12 + 15i}{5 + 2i} = \frac{(12 + 15i)(5 - 2i)}{(5 + 2i)(5 - 2i)} = \frac{90 + 51i}{29} = (3 + 2i) + \frac{3 - 7i}{29}
\]

Here in the last step we chose \(3 + 2i\) as \(\beta_1\) because \(3\) is the closest integer to \(90/29\) and \(2\) is the closest integer to \(51/29\). Having found a likely candidate for \(\beta_1\), we can use the equation \(\alpha_0 = \beta_1 \alpha_1 + \alpha_2\) to find \(\alpha_2\). This equation is
\[
12 + 15i = (3 + 2i)(5 + 2i) + \alpha_2 = (11 + 16i) + \alpha_2
\]

hence \(\alpha_2 = 1 - i\). Notice that \(N(1 - i) = 2 < N(5 + 2i) = 29\) so we have \(N(\alpha_2) < N(\alpha_1)\) as we wanted.

Will the process of choosing \(\beta_1\) as the nearest Gaussian integer to the ‘Gaussian rational’ \(\alpha_0/\alpha_1\) always lead to an \(\alpha_2\) with \(N(\alpha_2) < N(\alpha_1)\)? The answer is yes because if we write the quotient \(\alpha_2/\alpha_1\) in the form \(x + yi\) for rational numbers \(x\) and \(y\) (so in the example above we have \(x + yi = \frac{3}{29} + \frac{7}{29}i\)) then having \(\beta_1\) the closest Gaussian integer to \(\alpha_0/\alpha_1\) says that \(|x| \leq \frac{1}{2}\) and \(|y| \leq \frac{1}{2}\), so
\[
N\left(\frac{\alpha_2}{\alpha_1}\right) = x^2 + y^2 \leq \frac{1}{4} + \frac{1}{4} < 1
\]

and hence
\[
N(\alpha_2) = N\left(\frac{\alpha_2}{\alpha_1} \cdot \alpha_1\right) = N\left(\frac{\alpha_2}{\alpha_1}\right)N(\alpha_1) < N(\alpha_1)
\]

This shows that there is a general Euclidean algorithm in \(\mathbb{Z}[i]\), hence \(\mathbb{Z}[i]\) has unique factorization.

Just as an exercise let us finish carrying out the Euclidean algorithm for \(\alpha_0 = 12 + 15i\) and \(\alpha_1 = 5 + 2i\). The next step is to divide \(\alpha_2 = 1 - i\) into \(\alpha_1 = 5 + 2i\):
\[
\frac{5 + 2i}{1 - i} = \frac{(5 + 2i)(1 + i)}{(1 - i)(1 + i)} = \frac{3 + 7i}{2} = (1 + 3i) + \frac{1 + i}{2}
\]

Notice that the fractions \(3/2\) and \(7/2\) are exactly halfway between two consecutive integers, so instead of choosing \(1 + 3i\) for the closest integer to \((3 + 7i)/2\) we could equally well have chosen \(2 + 3i\), \(1 + 4i\), or \(2 + 4i\). Let us stick with the choice \(1 + 3i\) and use this to calculate the next \(\alpha_i\):
\[
5 + 2i = (1 + 3i)(1 - i) + \alpha_3 = (4 + 2i) + \alpha_3
\]

hence \(\alpha_3 = 1\). The final step would be simply to write \(1 - i = (1 - i)1 + 0\). Thus the full Euclidean algorithm gives the following equations:
\[
12 + 15i = (3 + 2i)(5 + 2i) + (1 - i)
\]
\[
5 + 2i = (1 + 3i)(1 - i) + 1
\]
\[
1 - i = (1 - i)(1) + 0
\]

In particular, since the last nonzero remainder is \(1\), a unit in \(\mathbb{Z}[i]\), we deduce that this is the greatest common divisor of \(12 + 15i\) and \(5 + 2i\), where ‘greatest’ means ‘of
greatest norm’. In other words \(12 + 15i\) and \(5 + 2i\) have no common divisors other than units.

The equations that display the results of carrying out the Euclidean algorithm can be used to express the last nonzero remainder in terms of the original two numbers:

\[
1 = (5 + 2i) - (1 + 3i)(1 - i) \\
= (5 + 2i) - (1 + 3i)[(12 + 15i) - (3 + 2i)(5 + 2i)] \\
= -(1 + 3i)(12 + 15i) + (-2 + 11i)(5 + 2i)
\]

If it had happened that the last nonzero remainder was a unit other than 1, we could have expressed this unit in terms of the original two Gaussian integers, and then multiplied the equation by the inverse of the unit to get an expression for 1 in terms of the original two Gaussian integers.

Having shown that prime factorizations in \(\mathbb{Z}[i]\) are unique, let us see what this implies about the representation problem for the norm form \(x^2 + y^2\). The equation \(x^2 + y^2 = n\) can be written as \((x + yi)(x - yi) = n\). If the prime factorization of \(x + yi\) in \(\mathbb{Z}[i]\) is \(x + yi = \alpha_1 \cdots \alpha_l\) and the prime factorization of \(n\) in \(\mathbb{Z}\) is \(n = p_1 \cdots p_m\) then the equation \(x^2 + y^2 = n\) becomes \(\alpha_1 \overline{\alpha}_1 \cdots \alpha_l \overline{\alpha}_l = p_1 \cdots p_m\).

A prime \(\alpha_j\) which is a prime \(p = 4k + 3\) in \(\mathbb{Z}\) (recall that such a \(p\) is called inert) is matched with \(\overline{\alpha}_j\) which also equals \(p\) so this corresponds to two factors of \(p_1 \cdots p_m\) being this same \(p\). Inert primes must therefore occur to an even power in \(n\). On the other hand each pair \(\alpha_j, \overline{\alpha}_j\) that does not lie in \(\mathbb{Z}\) corresponds to a single split prime factor of \(n\), so these split primes are either 2 or primes \(p = 4k + 1\). Thus we see that the equation \(x^2 + y^2 = n\) has an integer solution \((x, y)\) exactly when each inert prime factor \(p = 4k + 3\) of \(n\) occurs with an even exponent in \(n\). This agrees with the answer we got in Chapter 6, but the only nontrivial result from that chapter we have used here is the fact that all primes \(p = 4k + 1\) are represented by \(x^2 + y^2\).

Going further, we can also answer the more subtle question of when the equation \(x^2 + y^2 = n\) has a solution with \(x\) and \(y\) coprime. For \(x\) and \(y\) not to be coprime means they are both divisible by some integer greater than 1, which we can take to be a prime \(p\). For \(x\) and \(y\) to be divisible by \(p\) is the same as saying that \(x + yi\) is divisible by \(p\) in \(\mathbb{Z}[i]\), and likewise for \(x - yi\). If a prime factor \(p\) of \(n\) in \(\mathbb{Z}\) is inert in \(\mathbb{Z}[i]\) then in the factorization \(n = (x + yi)(x - yi)\) we will have \(p\) as a prime factor of \(x + yi\) or \(x - yi\) in \(\mathbb{Z}[i]\), so \(x\) and \(y\) will not be coprime. Thus \(n\) must be a product of primes \(p\) that split as \(p = \alpha \overline{\alpha}\) in \(\mathbb{Z}[i]\). We cannot have both \(\alpha\) and \(\overline{\alpha}\) as factors of \(x + yi\), otherwise \(p\) would divide \(x + yi\). Thus if \(p\) appears to the \(k\)th power in \(n\), we can have \(\alpha^k\) as a factor of \(x + yi\) and \(\overline{\alpha}^k\) as a factor of \(x - yi\) or vice versa, at least when \(p\) is unramified in \(\mathbb{Z}[i]\). If \(p\) is ramified, so \(\overline{\alpha}\) is a unit times \(\alpha\), then we must have \(k = 1\), otherwise \(x + yi\) or \(x - yi\) would have \(p\) as a factor. We noted earlier that 2 is the only prime in \(\mathbb{Z}\) that is ramified in \(\mathbb{Z}[i]\), so the final result is that \(x^2 + y^2 = n\) has a solution with \(x\) and \(y\) coprime exactly when
n is a product of primes \( p = 4k + 1 \) (we can assume \( n > 1 \)) or twice such a product. This too agrees with what we showed in Chapter 6.

For negative discriminants it is not difficult to figure out exactly when \( R_\Delta \) has a Euclidean algorithm. Recall that this means that for each pair of nonzero elements \( \alpha_0 \) and \( \alpha_1 \) in \( R_\Delta \) there should exist elements \( \beta \) and \( \alpha_2 \) such that \( \alpha_0 = \beta \alpha_1 + \alpha_2 \) and \( N(\alpha_2) < N(\alpha_1) \). Since \( \alpha_2 \) is determined by \( \alpha_0, \alpha_1, \) and \( \beta \), this is equivalent to saying that there should exist an element \( \beta \) in \( R_\Delta \) such that \( N(\alpha_0 - \beta \alpha_1) < N(\alpha_1) \).

The last inequality can be rewritten as \( N(\frac{\alpha_0}{\alpha_1} - \beta) < 1 \). Geometrically this is saying that every point \( \frac{\alpha_0}{\alpha_1} \) in the plane should be within a distance less than 1 of a point in the lattice \( R_\Delta \). We can check this by seeing whether the interiors of all circles of radius 1 centered at points of \( R_\Delta \) completely cover the plane.

For \( \mathbb{Z}[\sqrt{D}] \) with \( D < 0 \) the critical case \( D = -3 \) is shown in the first figure below, where the triangle is an equilateral triangle of side length 1.

For \( \mathbb{Z}[\sqrt{-2}] \) does have a Euclidean algorithm.

The case of \( \mathbb{Z}[\omega] \) for \( \omega = (1 + \sqrt{1 + 4d})/2 \) with \( d < 0 \) is illustrated in the right half of the figure above. Here the second row of disks are at height \( \sqrt{|1 + 4d|}/2 \) above the first row, so from the figure we see that the condition we need is that this height should be less than \( 1 + \frac{\sqrt{3}}{2} \). Thus we need \( \sqrt{|1 + 4d|} < 2 + \sqrt{3} \). Squaring both sides gives \( |1 + 4d| < 7 + 4\sqrt{3} \) which is satisfied only in the cases \( d = -1, -2, -3 \).

In summary, we have shown the following result:

**Proposition 8.10.** The only negative discriminants \( \Delta \) for which \( R_\Delta \) has a Euclidean algorithm are \( \Delta = -3, -4, -7, -8, -11 \).

Notice that these are the first five negative discriminants.

For even discriminants \( \Delta = 4D \) it is easy to prove that unique factorization fails in \( R_\Delta = \mathbb{Z}[\sqrt{D}] \) in all cases when \( \Delta \) is negative and there is no Euclidean algorithm:
Proposition 8.11. Unique factorization fails in \(\mathbb{Z}[\sqrt{D}]\) whenever \(D < -2\), and it also fails when \(D > 0\) and \(D \equiv 1\) modulo 4.

Proof: The number \(D^2 - D\) factors in \(\mathbb{Z}[\sqrt{D}]\) as \((D + \sqrt{D})(D - \sqrt{D})\), and it also factors as \(D(D - 1)\). The number 2 divides either \(D\) or \(D - 1\) since one of these two consecutive integers must be even. However, 2 does not divide either \(D + \sqrt{D}\) or \(D - \sqrt{D}\) in \(\mathbb{Z}[\sqrt{D}]\) since \((D \pm \sqrt{D})/2\) is not an element of \(\mathbb{Z}[\sqrt{D}]\) as the coefficient of \(\sqrt{D}\) in this quotient is not an integer. If we knew that 2 was prime in \(\mathbb{Z}[\sqrt{D}]\) we would then have two distinct factorizations of \(D^2 - D\) into primes in \(\mathbb{Z}[\sqrt{D}]\): One obtained by combining prime factorizations of \(D\) and \(D - 1\) in \(\mathbb{Z}[\sqrt{D}]\) and the other obtained by combining prime factorizations of \(D + \sqrt{D}\) and \(D - \sqrt{D}\). The first factorization would contain the prime 2 and the second would not.

It remains to check that 2 is a prime in \(\mathbb{Z}[\sqrt{D}]\) in the cases listed. If it is not a prime, then it factors as \(2 = \alpha\beta\) with neither \(\alpha\) nor \(\beta\) a unit, so we would have \(N(\alpha) = N(\beta) = \pm 2\). Thus the equation \(x^2 - Dy^2 = \pm 2\) would have an integer solution \((x, y)\). This is clearly impossible if \(D = -3\) or any negative integer less than \(-3\). If \(D > 0\) and \(D \equiv 1\) modulo 4 then if we look at the equation \(x^2 - Dy^2 = \pm 2\) modulo 4 it becomes \(x^2 - y^2 \equiv 2\), but this is impossible since \(x^2\) and \(y^2\) are congruent to 0 or 1 modulo 4, so \(x^2 - y^2\) is congruent to 0, 1, or \(-1\).

This proposition says in particular that unique factorization fails in \(\mathbb{Z}[\sqrt{-3}]\), \(\mathbb{Z}[\sqrt{-7}]\), and \(\mathbb{Z}[\sqrt{-11}]\), but when we enlarge these three rings to \(\mathbb{Z}[\omega]\) for \(\omega\) equal to \(\frac{1 + \sqrt{-3}}{2}\), \(\frac{1 + \sqrt{-7}}{2}\), and \(\frac{1 + \sqrt{-11}}{2}\) we do have unique factorization. A similar thing happens when we enlarge \(\mathbb{Z}[\sqrt{-8}]\) to \(\mathbb{Z}[\sqrt{-2}]\). In all these cases the enlargement replaces a nonfundamental discriminant by one which is fundamental.

One might wonder whether there are other ways to enlarge \(\mathbb{Z}[\sqrt{D}]\) to make prime factorization unique when it is not unique in \(\mathbb{Z}[\sqrt{D}]\) itself. Without changing things too drastically, suppose we just tried a different choice of \(\omega\) besides \((1 + \sqrt{1 + 4d})/2\). In order to do multiplication within the set \(\mathbb{Z}[\omega]\) of numbers \(x + y\omega\) with \(x\) and \(y\) integers one must be able to express \(\omega^2\) as \(m\omega + n\), so \(\omega\) must satisfy a quadratic equation \(\omega^2 - m\omega - n = 0\). This has roots \((m \pm \sqrt{m^2 + 4n})/2\), so we see that larger denominators than 2 in the definition of \(\omega\) will not work. If \(m\) is even, say \(m = 2k\), then \(\omega\) becomes \(k + \sqrt{k^2 + n}\), with no denominators at all and we are back in the situation of \(\mathbb{Z}[\sqrt{D}]\). If \(m\) is odd, say \(m = 2k + 1\), then \(\omega\) is \((2k + 1 + \sqrt{4k^2 + 4k + 1 + 4n})/2\), which can be written as \(k + (1 + \sqrt{1 + 4d})/2\) so the ring \(\mathbb{Z}[\omega]\) in this case would be the same as when we chose \(\omega = (1 + \sqrt{1 + 4d})/2\).

It is known that there are only nine negative discriminants for which \(R_\Delta\) has unique factorization, the discriminants

\[
\Delta = -3, -4, -7, -8, -11, -19, -43, -67, -163
\]

These are exactly the nine negative discriminants for which all quadratic forms of that discriminant are equivalent. This is not an accident since the usual way one
determines whether unique factorization holds is by proving that unique factorization holds precisely when all forms of the given discriminant are equivalent, as we will see later in the chapter.

For positive discriminants the norm form is hyperbolic so it takes on both positive and negative values. The Euclidean algorithm is then modified so that in the equations \( \alpha_{i-1} = \beta_i \alpha_i + \alpha_{i+1} \) it is required that \(|N(\alpha_{i+1})| < |N(\alpha_i)|\). It is known that there are exactly 16 positive fundamental discriminants for which there is a Euclidean algorithm in \( R_\Delta \):

\[
\Delta = 5, 8, 12, 13, 17, 21, 24, 28, 29, 33, 37, 41, 44, 57, 73, 76
\]

The determination of this list is much more difficult than for negative discriminants since the norm no longer has the nice geometric meaning of the square of the distance to the origin in the plane.

There are many positive fundamental discriminants for which \( R_\Delta \) has unique factorization even though there is no Euclidean algorithm. The fundamental discriminants less than 100 with this property are 53, 56, 61, 69, 77, 88, 89, 92, 93, 97.

To conclude this section we will illustrate further the usefulness of unique prime factorization in \( R_\Delta \) by an extended example in which we finish the determination of which numbers are represented by the form \( x^2 + 7y^2 \). We studied this problem in an example in Section 6.4 and found that a necessary condition for a number \( n \) to be represented by \( x^2 + 7y^2 \) was that \( n \) factors into primes as \( n = 2^a 7^b p_1^{k_1} \cdots p_m^{k_m} \) where \( a = 0 \) or \( a \geq 3 \), \( b = 1 \), and the \( p_i \)'s are distinct odd primes congruent to 1, 2, or 4 mod 7. Now we will show the converse, that each \( n \) having such a factorization is represented by \( x^2 + 7y^2 \). This includes the curious fact that all powers of 2 starting with 8 have such a representation. For example \( 8 = 1^2 + 7 \cdot 1^2 \), \( 16 = 3^2 + 7 \cdot 1^2 \), \( 32 = 5^2 + 7 \cdot 1^2 \), \( 64 = 1^2 + 7 \cdot 3^2 \), \( 128 = 11^2 + 7 \cdot 1^2 \), \( 256 = 9^2 + 7 \cdot 5^2 \), \( 512 = 13^2 + 7 \cdot 2^2 \), \( 1024 = 31^2 + 7 \cdot 3^2 \), \( 2048 = 5^2 + 7 \cdot 17^2 \), \( 4096 = 57^2 + 7 \cdot 11^2 \), \( 8192 = 67^2 + 7 \cdot 23^2 \), \( 16384 = 47^2 + 7 \cdot 45^2 \), and \( 32768 = 181^2 + 7 \cdot 1^2 \). As we will see, these are the unique primitive solutions with positive \( x \) and \( y \) in each case. Without the primitivity requirement solutions are much easier to construct just by repeated doubling of the solutions for \( n = 8 \) and \( n = 16 \), so for example \( 32 = 2^2 + 7 \cdot 2^2 \), \( 64 = 6^2 + 7 \cdot 2^2 \), \( 128 = 4^2 + 7 \cdot 4^2 \), and so on.

The form \( x^2 + 7y^2 \) is the norm form in \( \mathbb{Z}[\sqrt{-7}] \), but rather than using \( \mathbb{Z}[\sqrt{-7}] \) which does not have the unique factorization property we will use the larger ring \( \mathbb{Z}[\omega] \) for \( \omega = (1 + \sqrt{-7})/2 \) which does have unique factorization. This implies in particular that all primes in \( \mathbb{Z}[\omega] \) are obtained by factoring primes in \( \mathbb{Z} \). Primes in \( \mathbb{Z} \) that do not appear in the topograph of the norm form \( x^2 + xy + 2y^2 \) remain prime in \( \mathbb{Z}[\omega] \) while primes that do appear factor as \( p = \alpha \overline{\alpha} \). The methods of Chapter 6 show that the primes \( p \) that factor as \( \alpha \overline{\alpha} \) are 7 (the only prime dividing the discriminant) which factors as \( (\sqrt{-7})(-\sqrt{-7}) \) and primes \( p \equiv 1, 2, 4 \mod 7 \). This includes \( p = 2 \) which
has the factorization \( 2 = \omega \overline{\omega} \). For odd primes the factorization actually takes place in \( \mathbb{Z}[\sqrt{-7}] \) since otherwise we would have \( p = \frac{a+b\sqrt{-7}}{2} \cdot \frac{a-b\sqrt{-7}}{2} = \frac{a^2+7b^2}{4} \) with \( a \) and \( b \) both odd, so \( a^2 \) and \( b^2 \) would be 1 mod 8 making \( a^2 + 7b^2 \) zero mod 8 which would mean \( p \) was even.

For factorizations \( p = \alpha \overline{\alpha} \) we will need to know whether \( \alpha \) and \( \overline{\alpha} \) differ only by a unit. Since the only units in \( \mathbb{Z}[\omega] \) are \( \pm 1 \) this is easy to figure out. In the factorization of 7 as \( (\sqrt{-7})(-\sqrt{-7}) \) the two factors differ by a unit. In all other cases the two factors \( \frac{a+b\sqrt{-7}}{2} \) and \( \frac{a-b\sqrt{-7}}{2} \) do not differ just by a unit (that is, just by a sign) since both \( a \) and \( b \) must be nonzero.

With \( n = 2^a 7^b p_1^{k_1} \cdots p_m^{k_m} \) as before, let \( p_i \) factor as \( p_i = \alpha_i \overline{\alpha}_i \). In the cases \( a \geq 3 \) let \( x \) and \( y \) be defined by the equation

\[
x + y\sqrt{-7} = 2\omega^{a-2}(\sqrt{-7})^{b} \alpha_1^{k_1} \cdots \alpha_m^{k_m}
\]

The reason for including the factor 2 on the right side of the equation is to guarantee that \( x \) and \( y \) are integers since \( \omega^{a-2} \), as an element of \( \mathbb{Z}[\omega] \), might have a denominator of 2. (We already noted that each \( \alpha_i \) lies in \( \mathbb{Z}[\sqrt{-7}] \).) When we take the norm of \( x + y\sqrt{-7} \) we get

\[
x^2 + 7y^2 = 4 \cdot 2^{a-2} 7^b p_1^{k_1} \cdots p_m^{k_m}
\]

so we have a solution of \( x^2 + 7y^2 = n \). All that remains to check is that this is a primitive solution when \( a \geq 3 \) and \( b \leq 1 \). This is where the unique factorization property in \( \mathbb{Z}[\omega] \) will be used strongly.

Suppose that some prime \( p \) divides both \( x \) and \( y \), hence \( p \) divides \( x + y\sqrt{-7} \) in \( \mathbb{Z}[\omega] \). There are several cases to consider:

- If \( p \) remains prime in \( \mathbb{Z}[\omega] \) then it cannot divide \( x + y\sqrt{-7} \) since it does not divide any of the prime factors of \( x + y\sqrt{-7} \).
- If \( p = p_i \) for some \( i \) then since \( p = \alpha_i \overline{\alpha}_i \), for \( p \) to divide \( x + y\sqrt{-7} \) implies that \( \overline{\alpha}_i \) divides \( x + y\sqrt{-7} \), which it doesn’t since \( \overline{\alpha}_i \) is not a unit times \( \alpha_i \).
- If \( p = 7 \) then it does not divide \( x + y\sqrt{-7} \) since we assume \( b \leq 1 \).
- If \( p = 2 \) we can divide the formula for \( x + y\sqrt{-7} \) by 2 to obtain a formula \( w + z\sqrt{-7} = \omega^{a-2}(\sqrt{-7})^{b} \alpha_1^{k_1} \cdots \alpha_m^{k_m} \) with both \( w \) and \( z \) integers. Since we assume \( a \geq 3 \) this means \( N(x + y\sqrt{-7}) \) is divisible by 8 so \( N(w + z\sqrt{-7}) = w^2 + 7z^2 \) is divisible by 2. This implies that \( w \) and \( z \) have the same parity. If they are both even then \( 2 = \omega \overline{\omega} \) would divide the right side of the formula for \( w + z\sqrt{-7} \), hence \( \overline{\omega} \) would divide it, a contradiction. If \( w \) and \( z \) are both odd then \( \frac{w+z\sqrt{-7}}{2} \) would be an element of \( \mathbb{Z}[\omega] \) so 2 would again divide \( w + z\sqrt{-7} \) leading to the same contradiction.

Thus we have found a primitive solution of \( x^2 + 7y^2 = n \) in the cases \( a \geq 3 \). When \( a = 0 \) we can omit the term \( 2\omega^{a-2} \) from the formula for \( x + y\sqrt{-7} \) and follow the same line of reasoning, with a few resulting simplifications in the argument. The only difference occurs at the very last step, where we check that \( 2 \) does not divide \( x + y\sqrt{-7} \) by noting that neither \( \omega \) nor \( \overline{\omega} \) divides \( (\sqrt{-7})^{b} \alpha_1^{k_1} \cdots \alpha_m^{k_m} \).
We stated earlier that the equation \( x^2 + 7y^2 = 2^k \) has only one primitive solution with \( x \) and \( y \) positive. To see this, factor this equation as \( (x + y\sqrt{-7})(x - y\sqrt{-7}) = \omega^k\overline{\omega}^k \). By unique factorization, \( x + y\sqrt{-7} \) must equal \( \pm \omega^l\overline{\omega}^m \) for some \( l \) and \( m \) with \( l + m = k \). Choosing \( m = 1 \) gives the primitive solution \( \pm 2\omega^k \) that we found above, and the conjugate of this solution with \( l = 1 \) is another primitive solution. Choosing \( l = 0 \) or \( m = 0 \) gives a solution in \( \mathbb{Z}[\omega] \) but not in \( \mathbb{Z}[\sqrt{-7}] \), otherwise multiplying this solution by \( 2 = \omega\overline{\omega} \) would give the nonprimitive solution \( \pm 2\omega^k \) or \( \pm 2\overline{\omega}^k \) of \( x^2 + 7y^2 = 2^{k+2} \), contradicting our earlier proof that this solution is primitive. When \( l > 1 \) and \( m > 1 \) with \( l \neq m \) we get nonprimitive solutions obtained from the known primitive solutions for a smaller value of \( k \) by multiplying by a power of 2. When \( l = m \) we get the nonprimitive solution \( x = 2^l, y = 0 \). Thus we have shown that there are at most four primitive solutions. The signs of \( x \) and \( y \) can be changed arbitrarily, so if we take \( x \) and \( y \) to be positive the number of primitive solutions reduces to one.

A historical footnote: In the early 1900s the number theorist Ramanujan observed that the Diophantine equation \( x^2 + 7y^2 = 2^k \) has solutions for \( k = 3, 4, 5, 7, 15 \) and he conjectured that there were no solutions for larger \( k \). In terms of the preceding example this is saying that the only solutions of \( x^2 + 7y^2 = 2^k \) with \( y = 1 \) occur in these five cases. Ramanujan's conjecture was later proved in a paper by Skolem, Chowla, and Lewis published in 1959.

### 3. The Correspondence Between Forms and Ideals

So far in this chapter we have focused on principal forms, and now we begin to extend what we have done to arbitrary forms. For principal forms we began by factoring them as a product of two linear factors whose coefficients involved square roots, for example the factorization \( x^2 - Dy^2 = (x + \sqrt{D}y)(x - \sqrt{D}y) \) in the case of discriminant \( \Delta = 4D \). For a general form \( Q(x, y) = ax^2 + bxy + cy^2 \) of discriminant \( \Delta \) the corresponding factorization is \( a(x - \alpha y)(x - \overline{\alpha}y) \) where \( \alpha \) is a root of the quadratic equation \( ax^2 + bxy + c = 0 \). Thus we have

\[
ax^2 + bxy + cy^2 = a\left(x - \frac{-b + \sqrt{\Delta}}{2a}y\right)\left(x - \frac{-b - \sqrt{\Delta}}{2a}y\right)
\]

An equivalent equation that will be more convenient for our purposes is obtained by multiplying both sides by the coefficient \( a \) to obtain

\[
a(ax^2 + bxy + cy^2) = \left(ax + \frac{b + \sqrt{\Delta}}{2}y\right)\left(ax + \frac{b - \sqrt{\Delta}}{2}y\right)
\]

Notice that now in each of the two linear factors on the right the coefficients of \( x \) and \( y \) lie in the ring \( R_\Delta \) since \( b \) must have the same parity as \( \Delta \), so if \( \Delta = 4D \) we can eliminate the denominator 2 in the coefficient of \( y \) to obtain an element of \( \mathbb{Z}[\sqrt{D}] \) while if \( \Delta = 4d+1 \) the fraction lies in \( \mathbb{Z}[\omega] \) since \( b \) is odd. Another thing to observe is
that the right side of the equation is just the norm $N(a x + \frac{b + \sqrt{\Delta}}{2} y)$, so the displayed equation above can be written more concisely as $aQ(x, y) = N(a x + \frac{b + \sqrt{\Delta}}{2} y)$.

For a form $Q(x, y) = ax^2 + bxy + cy^2$ the set of numbers $ax + \frac{b + \sqrt{\Delta}}{2} y$ as $x$ and $y$ range over all integers forms a lattice contained in the larger lattice $R_\Delta$ in the plane. Here we use the term lattice to refer to a set of numbers of the form $\alpha x + \beta y$ for fixed nonzero elements $\alpha$ and $\beta$ of $R_\Delta$, with $x$ and $y$ varying over $\mathbb{Z}$, and we assume that $\alpha$ and $\beta$ do not lie on the same line through the origin. We denote this lattice by $L(\alpha, \beta)$ and call $\alpha$ and $\beta$ a basis for this lattice.

In particular, associated to the form $Q$ we have the lattice $L_Q = L(a, \frac{b + \sqrt{\Delta}}{2})$ consisting of all the numbers $ax + \frac{b + \sqrt{\Delta}}{2} y$ for integers $x$ and $y$. The earlier equation $aQ(x, y) = N(a x + \frac{b + \sqrt{\Delta}}{2} y)$ then says that the form $Q$ is obtained from the lattice $L_Q$ by taking the norms of all its elements and multiplying by the constant factor $1/a$, which can be regarded as a sort of normalization constant as we will see in more detail later.

Let us look at some examples to see what $L_Q$ can look like in the case $\Delta = -4$ so $R_\Delta = \mathbb{Z}[i]$, the Gaussian integers. In this case we have $ax + \frac{b + \sqrt{\Delta}}{2} y = ax + (b' + i) y$ where $b' = b/2$, an integer since $b$ always has the same parity as $\Delta$. For the principal form $x^2 + y^2$ we have $a = 1$ and $b' = 0$ so $L_Q = L(1, i) = \mathbb{Z}[i]$. Four more cases are shown in the figures below.
In each case the lattice forms a grid of squares, rotated and expanded from the square grid formed by $\mathbb{Z}[i]$ itself. Not all lattices in $\mathbb{Z}[i]$ form square grids since for example one could have a lattice of long thin rectangles such as $L(10, i)$.

A 90 degree rotation of the plane about the origin takes a square lattice to itself. Conversely, a lattice that is taken to itself by a 90 degree rotation about the origin must be a square lattice. To see this, observe first that the 90 degree rotation takes the closest lattice point to the origin to another closest lattice point, with the sum of these two lattice points giving another lattice point that is the fourth vertex of a square of lattice points. There can be no lattice points in the interior of this square since such a point would be closer to a corner of the square than the length of the side of the square, which is impossible since the minimum distance between any two points in a lattice equals the minimum distance from the origin to a lattice point.

Since 90 degree rotation is the same as multiplication of complex numbers by $i$, we could also say that square lattices are those that are taken to themselves by multiplication by $i$. Once a lattice has this property it follows that multiplication by an arbitrary element of $\mathbb{Z}[i]$ takes the lattice into itself. Namely, if we know that $i\alpha$ is in a lattice $L$ whenever $\alpha$ is in $L$, then for arbitrary integers $m$ and $n$ it follows that $m\alpha$ and $n i\alpha$ are in $L$ and hence also $(m + ni)\alpha$ is in $L$.

There is a standard term for this concept. A lattice $L$ in $\mathbb{R} \Delta$ is called an ideal if for each element $\alpha$ in $L$ and each $\beta$ in $\mathbb{R} \Delta$ the product $\beta\alpha$ is in $L$. In other words, $L$ is taken to itself by multiplication by every element of $\mathbb{R} \Delta$. The term ‘ideal’ may seem like an odd name, but it originally arose in a slightly different context where it seems more natural, as we will see later in the chapter. For now we can just imagine that ideals are the best kind of lattices, ‘ideal lattices’. 

The fact that all lattices $L_Q$ in $\mathbb{Z}[i]$ are square lattices is a special case of the following general fact:

**Proposition 8.12.** Every lattice $L_Q$ associated to a quadratic form $Q$ of discriminant $\Delta$ is an ideal in $R_\Delta$.

**Proof:** To cover all discriminants at once we can write $R_\Delta$ as $\mathbb{Z}[\tau]$ for $\tau = \frac{e + \sqrt{\Delta}}{2}$ where $e$ is 0 if $\Delta = 4D$ and 1 if $\Delta = 4d + 1$. What we need to check in order to verify that the lattice $L_Q = L(a, \frac{b + \sqrt{\Delta}}{2})$ is an ideal is that both of the products $\tau \cdot a$ and $\tau \cdot \frac{b + \sqrt{\Delta}}{2}$
are elements of $L_Q$. For the product $\tau \cdot a$ this means we want to solve the equation
\[
\frac{e + \sqrt{\Delta}}{2} \cdot a = ax + \frac{b + \sqrt{\Delta}}{2} y
\]
for integers $x$ and $y$. Comparing the coefficients of $\sqrt{\Delta}$ on both sides of the equation, we get $y = a$, an integer. Substituting $y = a$ into the equation then gives $\frac{ea}{2} = ax + \frac{ba}{2}$ so $x = \frac{e-b}{2}$. This is an integer since both $e$ and $b$ have the same parity as $\Delta$.

For the other product $\tau \cdot \frac{b+\sqrt{\Delta}}{2}$ we have a similar equation
\[
\frac{e + \sqrt{\Delta}}{2} \cdot \frac{b + \sqrt{\Delta}}{2} = ax + \frac{b + \sqrt{\Delta}}{2} y
\]
which we can rewrite as
\[
\frac{eb + \Delta + (e+b)\sqrt{\Delta}}{4} = ax + \frac{b + \sqrt{\Delta}}{2} y
\]
From the coefficients of $\sqrt{\Delta}$ we get $y = \frac{e+b}{2}$ which is an integer since $e$ and $b$ have the same parity. Then the equation becomes
\[
\frac{eb + \Delta}{4} = ax + \frac{eb + b^2}{4}
\]
which simplifies to $ax = \frac{\Delta - b^2}{4}$. Since $\Delta = b^2 - 4ac$ we have the integer solution $x = c$. \qed

We saw in the case of $\mathbb{Z}[i]$ that all ideals are square lattices, so they are obtained from $\mathbb{Z}[i]$ by rotation about the origin and expansion. There are a few other negative discriminants where the same thing happens and all ideals differ only by rotation and rescaling, either expansion or contraction. One example is when $\Delta = -8$ so we have $R_\Delta = \mathbb{Z}[\sqrt{-2}]$ which forms a rectangular lattice with rectangles of side lengths 1 and $\sqrt{2}$. For an arbitrary ideal $L$ in $\mathbb{Z}[\sqrt{-2}]$ let $\alpha$ be a nonzero point in $L$ closest to the origin. Since $L$ is an ideal, the product $\sqrt{-2} \alpha$ must also be in $L$. Since multiplication by $\sqrt{-2}$ rotates the plane by 90 degrees and expands it by a factor of $\sqrt{2}$, the set of all linear combinations $\alpha x + \sqrt{-2} \alpha y$ for integers $x$ and $y$ forms a rectangular sublattice $L'$ of $L$ obtained from $\mathbb{Z}[\sqrt{-2}]$ by rotation and expansion. Since we chose $\alpha$ as the closest point of $L$ to the origin, say of distance $A$ to the origin, there can be no points of $L$ within a distance less than $A$ of any point of $L'$. In other words, if one takes the union of the interiors of all disks of radius $A$ centered at points of $L'$, this union intersects $L$ just in $L'$. However, this union is the whole plane since the ratio of the side lengths of the rectangles of $L'$ is $\sqrt{2}$. Thus $L$ equals the rectangular lattice $L'$. 

![Diagram.png]
This is essentially the same geometric argument we used to show that \( \mathbb{Z}[\sqrt{-2}] \) has a Euclidean algorithm. There were five negative discriminants \( \Delta \) for which \( R_\Delta \) has a Euclidean algorithm, \( \Delta = -3, -4, -7, -8, -11 \). The argument in the preceding paragraph shows that in each of these cases all ideals in \( R_\Delta \) are equivalent under rotation and rescaling. In the case \( \Delta = -3 \) the Eisenstein integers \( \mathbb{Z}[\omega] \) form a grid of equilateral triangles so all ideals are also grids of equilateral triangles that are taken to themselves by multiplication by \( \omega \), rotating the plane by 60 degrees.

For \( \Delta = -7 \) and \( -11 \) the lattice \( R_\Delta = \mathbb{Z}[\omega] \) for \( \Delta = -3 \) is stretched vertically to form a grid of isosceles triangles and all ideals are also grids of isosceles triangles, rotated and rescaled from the triangles in \( R_\Delta \).

We have been using the fact that multiplication by a fixed nonzero complex number \( \alpha \) always has the effect of rotating and rescaling the plane, keeping the origin fixed. Since multiplication by \( \alpha \) sends 1 to \( \alpha \), the rescaling factor is the distance from \( \alpha \) to the origin and the angle of rotation is the angle between the positive \( x \) axis and the ray from the origin to \( \alpha \). Since \( \alpha \) can be any nonzero complex number, every rotation and rescaling is realizable as multiplication by a suitably chosen \( \alpha \).

Let us look at some examples of discriminants where not all forms are equivalent to see whether there is more variety in the shapes of the lattices \( L_Q \), so they are not all obtained from \( R_\Delta \) by rotation and rescaling. First consider the lattices \( L_Q \) in \( \mathbb{Z}[\sqrt{-6}] \) for the two non-equivalent forms \( x^2 + 6y^2 \) and \( 2x^2 + 3y^2 \) of discriminant \( -24 \). These are the lattices \( L(1, \sqrt{-6}) = \mathbb{Z}[-\sqrt{6}] \) and \( L(2, \sqrt{-6}) \) shown below.

\[
x^2 + 6y^2 \rightarrow L(1, \sqrt{-6}) \quad 2x^2 + 3y^2 \rightarrow L(2, \sqrt{-6})
\]

The two lattices do not appear to differ just by rotation and rescaling, and we can verify this by computing the ratio of the distances from the origin to the closest lattice point and to the next-closest lattice point on a different line through the origin. For the lattice \( \mathbb{Z}[\sqrt{-6}] \) this ratio is \( 1/\sqrt{6} \) while for the other lattice it is \( 2/\sqrt{6} \). If the lattices differed only by rotation and rescaling the ratios would be the same.
As another example, consider the lattices $L_Q$ in $\mathbb{Z}[\sqrt{-5}]$ for the non-equivalent forms $x^2 + 5y^2$ and $2x^2 + 2xy + 3y^2$ of discriminant $-20$. The two lattices are $L(1, \sqrt{-5}) = \mathbb{Z}[\sqrt{-5}]$ and $L(2, 1 + \sqrt{-5})$.

\[
\begin{align*}
  x^2 + 5y^2 & \rightarrow L(1, \sqrt{-5}) \\
  2x^2 + 2xy + 3y^2 & \rightarrow L(2, 1 + \sqrt{-5})
\end{align*}
\]

Again we can check the lattices are not related just by rotation and rescaling by computing the same ratios of distances, which are $1/\sqrt{5}$ and $2/\sqrt{6}$ for the two lattices. It is also clear visually that the first lattice is rectangular while the second is not.

A slightly more complicated example is $\mathbb{Z}[\sqrt{-14}]$ with $\Delta = -56$ where there are four proper equivalence classes of forms given by $x^2 + 14y^2$, $2x^2 + 7y^2$, $3x^2 + 2xy + 5y^2$, and $3x^2 - 2xy + 5y^2$. The corresponding lattices are $L(1, \sqrt{-14})$, $L(2, \sqrt{-14})$, $L(3, 1 + \sqrt{-14})$, and $L(3, -1 + \sqrt{-14})$ shown below.

Here the third and fourth forms are equivalent but not properly equivalent since their topographs have a source vertex surrounded by the three smallest values of the form. If the form is reduced, so that $0 \leq b \leq a \leq c$, these are the values $a$, $c$, and $a - b + c$ for $(x, y)$ equal to $(1, 0)$, $(0, 1)$, and $(-1, 1)$, and we have $a \leq c \leq a - b + c$. In the lattice $L_Q = L(a, \frac{b + \sqrt{\Delta}}{2})$.

An elliptic form $Q = ax^2 + bxy + cy^2$ has a source vertex or edge in its topograph, surrounded by the three smallest values of the form. If the form is reduced, so that $0 \leq b \leq a \leq c$, these are the values $a$, $c$, and $a - b + c$ for $(x, y)$ equal to $(1, 0)$, $(0, 1)$, and $(-1, 1)$, and we have $a \leq c \leq a - b + c$. In the lattice $L_Q = L(a, \frac{b + \sqrt{\Delta}}{2})$. 

these numbers determine the distances from the origin to the closest points in the lattice lying on three different lines through the origin via the formula $aQ(x, y) = N(ax + \frac{b+\sqrt{\Delta}}{2}y)$ with $(x, y)$ equal to $(1, 0)$, $(0, 1)$, or $(-1, 1)$. Thus the three norms are $a$ times each of $a$, $c$, and $a - b + c$. Taking square roots, the distances to the origin are $\sqrt{a^2} = a$, $\sqrt{ac}$, and $\sqrt{a(a - b + c)}$. These are the lengths of the sides of the triangles when the grid of parallelograms determined by $L_Q$ with the basis $a, \frac{b+\sqrt{\Delta}}{2}$ is subdivided into a grid of triangles:

When $b = 0$ the form $Q$ is $ax^2 + cy^2$ with a source edge rather than a source vertex and the parallelograms become rectangles. The three smallest values of $Q$ are $a$, $c$, and $a + c$ which occurs twice, for $(x, y)$ equal to $(1, 1)$ as well as $(-1, 1)$. These two choices for $(x, y)$ correspond to subdividing the grid of rectangles into a grid of right triangles by either the positive slope diagonals of the rectangles or the negative slope diagonals.

The examples we have seen so far may lead one to ask how exact a correspondence there is between proper equivalence classes of forms of a given discriminant $\Delta$ and the shapes of lattices that are ideals in $R_{\Delta}$, where two lattices that differ only by rotation and rescaling are regarded as having the same shape. The main theorem will be that this is an exact one-to-one correspondence for negative discriminants, while for positive discriminants there is an analogous one-to-one correspondence with a more subtle notion of ‘shape’ for lattices.

As a first step in this direction let us ask whether every ideal in $R_{\Delta}$ is equal to $L_Q$ for some form $Q$ of discriminant $\Delta$. One way to see that this is not true is to observe that the lattices $L_Q = L(a, \frac{b+\sqrt{\Delta}}{2})$ have the special property that they contain an element $\frac{b+\sqrt{\Delta}}{2}$ lying in the first row of the lattice $R_{\Delta}$ above the $x$ axis, but this is not the case for all ideals since we can expand a lattice $L$ by a positive integer factor $n$ to get a new lattice $nL$ consisting of all elements $n\alpha$ for $\alpha$ in $L$, and $nL$ is an ideal if $L$ is. Thus if we start with an ideal $L_Q$ we obtain another ideal $nL_Q$ which has no elements in the first row of $R_{\Delta}$ above the $x$ axis if $n > 1$. However, nothing more complicated than this can happen:

**Proposition 8.13.** Every ideal in $R_{\Delta}$ is equal to $nL_Q$ for some positive integer $n$ and
some form \( Q(x, y) = ax^2 + bxy + cy^2 \) of discriminant \( \Delta \) with \( a > 0 \).

**Proof:** We write \( R_\Delta \) as \( \mathbb{Z}[\tau] \) for \( \tau = \sqrt{D} \) when \( \Delta = 4D \) and \( \tau = \frac{1 + \sqrt{\Delta}}{2} \) when \( \Delta = 4d + 1 \). Let \( L \) be a lattice in \( \mathbb{Z}[\tau] \). Since \( L \) is not entirely contained in the \( x \) axis there exist elements \( m + n\tau \) in \( L \) with \( n > 0 \). Choose such an element \( \alpha = m + n\tau \) with minimum positive \( n \), so \( \alpha \) lies in the \( n^{th} \) row of \( \mathbb{Z}[\tau] \) and there are no elements of \( L \) in any row between the \( 0^{th} \) and the \( n^{th} \) rows. Since \( L \) is a lattice all elements of \( L \) must then lie in rows numbered an integer multiple of \( n \). For example the element \( k\alpha \) lies in the \( kn^{th} \) row for each integer \( k \). These elements \( k\alpha \) lie on a line through the origin, and \( L \) must also contain elements not on this line, so some \( kn^{th} \) row must contain another element \( \beta \) of \( L \) besides \( k\alpha \). The difference \( \beta - k\alpha \) then lies in the \( x \) axis and is a nonzero integer in \( L \). Choosing a minimal positive integer \( p \) in \( L \), the lattice property of \( L \) implies that the integers in \( L \) are precisely the integer multiples of \( p \). It follows that \( L \) contains the lattice \( L(p, \alpha) = L(p, m + n\tau) \), and in fact \( L \) is equal to \( L(p, m + n\tau) \) otherwise either \( p \) or \( n \) would not be minimal.

Now let us assume that \( L \) is an ideal in \( \mathbb{Z}[\tau] \), not just a lattice. Then \( p\tau \) lies in \( L \) since \( p \) does. Since \( p\tau \) is in the \( p^{th} \) row of \( \mathbb{Z}[\tau] \) we must have \( p = an \) for some positive integer \( a \). We also know that \( \alpha\tau \) must lie in \( L \) where \( \alpha = m + n\tau \) as above. In the case \( \Delta = 4D \) we have \( \tau = \sqrt{D} \) so \( \alpha\tau = m\tau + n\tau^2 = m\tau + nD \). This is in the \( m^{th} \) row of \( \mathbb{Z}[\tau] \) so \( n \) must divide \( m \), say \( m = nq \). In the case \( \Delta = 4d + 1 \) we have \( \tau^2 = \tau + d \) so \( \alpha\tau = (m + n)\tau + nd \). This is in the \( (m + n)^{th} \) row of \( \mathbb{Z}[\tau] \) so \( n \) divides \( m + n \) and hence also \( n \) so we can again write \( m = nq \). Thus \( L = L(p, m + n\tau) = L(na, nq + n\tau) = nL(a, q + \tau) \). Note that \( L(a, q + \tau) \) is an ideal since \( L \) is an ideal.

To finish the proof we would like to find integers \( b \) and \( c \) such that \( q + \tau = \frac{b + \sqrt{\Delta}}{2} \) and \( \Delta = b^2 - 4ac \) since then \( L(a, q + \tau) \) will be \( L_Q \) for \( Q = ax^2 + bxy + cy^2 \) with discriminant \( \Delta \). Consider first the case \( \Delta = 4D \) so \( q + \tau = q + \sqrt{D} \). This is an element of the ideal \( L(a, q + \sqrt{D}) \) so if we multiply it by \( q - \sqrt{D} \) we get an integer \( q^2 - D \) lying in \( L(a, q + \sqrt{D}) \). Thus \( q^2 - D \) must be a multiple of \( a \) which is the smallest positive integer in \( L(a, q + \sqrt{D}) \), so we get an equation \( q^2 - D = ac \) for some integer \( c \). Hence we have \( (2q)^2 - 4D = 4ac \), and since \( 4D = \Delta \) this can be rewritten as \( \Delta = b^2 - 4ac \) for \( b = 2q \). We also have \( q + \tau = q + \sqrt{D} = \frac{b + \sqrt{\Delta}}{2} \) so this finishes the case \( \Delta = 4D \).

In the other case \( \Delta = 4d + 1 \) we again look at the product \( (q + \tau)(q - \tau) \). By the same reasoning this must be a multiple of \( a \), so \( (q + \tau)(q - \tau) = ac \) for some integer \( c \). Writing this out, we have

\[
\left(q + \frac{1 + \sqrt{\Delta}}{2}\right)\left(q + \frac{1 - \sqrt{\Delta}}{2}\right) = ac
\]

\[
(2q + 1 + \sqrt{\Delta})(2q + 1 - \sqrt{\Delta}) = 4ac
\]

\[
(2q + 1)^2 - \Delta = 4ac
\]

so if we take \( b = 2q + 1 \) we have \( \Delta = b^2 - 4ac \) and \( q + \tau = q + \frac{1 + \sqrt{\Delta}}{2} = \frac{b + \sqrt{\Delta}}{2} \). \( \square \)
We have seen how to go from a quadratic form $Q$ to an ideal $L_Q$, and it will be useful to go in the opposite direction as well, from an ideal $L$ in $R_\Delta$ to a quadratic form $Q_L$ of discriminant $\Delta$. As motivation we can start with the earlier formula $aQ(x, y) = N(ax + b\sqrt{\Delta} y)$ which says that, up to the constant factor $a$, the form $Q$ is given by restricting the usual norm in $R_\Delta$ to the elements $ax + b\sqrt{\Delta} y$ in the ideal $L_Q$. We can try the same thing for any lattice $L = L(\alpha, \beta)$ in $R_\Delta$, defining a quadratic form $Q(x, y) = N(\alpha x + \beta y) = (\alpha x + \beta y)(\bar{\alpha}x + \bar{\beta}y) = \alpha\bar{\alpha}x^2 + (\alpha\bar{\beta} + \bar{\alpha}\beta)xy + \beta\bar{\beta}y^2$

Here the coefficients of $x^2$, $xy$, and $y^2$ are integers since they are equal to their conjugates. The form $Q$ depends on the choice of the basis $\alpha, \beta$ for $L$. Another basis $\alpha', \beta'$ can be expressed as linear combinations $\alpha' = p\alpha + q\beta$, $\beta' = r\alpha + s\beta$ with integer coefficients. Since the change of basis can be reversed, going from $\alpha', \beta'$ back to $\alpha, \beta$, the $2 \times 2$ matrix $\begin{pmatrix} p & q \\ r & s \end{pmatrix}$ has determinant $\pm 1$, and conversely any such matrix gives a valid change of basis for $L$. Changing the basis also produces a change of variables in the form $Q(x, y)$ since $N(\alpha' x + \beta' y) = N((p\alpha + q\beta)x + (r\alpha + s\beta)y) = N(\alpha(px + ry) + (\beta(qx + sy))) = Q(px + ry, qx + sy)$. Here the matrix is the transpose $\begin{pmatrix} p & r \\ q & s \end{pmatrix}$, with the same determinant $\pm 1$. Thus changing the basis for $L$ produces an equivalent form, and every equivalent form can be realized by some change of basis for $L$.

The form $N(\alpha x + \beta y)$ depends on the ordering for the two basis elements $\alpha$ and $\beta$ since reversing their order interchanges $x$ and $y$, which gives a mirror image topograph. We can eliminate this ambiguity by ordering $\alpha$ and $\beta$ so that the angle from the ray from 0 passing through $\alpha$ to the ray from 0 passing through $\beta$ is between 0 and $\pi$. We call this order the positive order. If we only use positively ordered bases, then the change of basis matrices have determinant $+1$ since they correspond to orientation-preserving linear transformations of the plane. Thus if we always use positively ordered bases, the lattice $L$ gives rise to a proper equivalence class of quadratic forms.

The norm form $N(\alpha x + \beta y)$ associated to a lattice $L = L(\alpha, \beta)$ in $R_\Delta$ might not have discriminant $\Delta$. For example, if we replace $L$ by $nL = L(n\alpha, n\beta)$ this multiplies the norm form by $n^2$ and so the discriminant is multiplied by $n^4$. We can always rescale a form to have any discriminant we want just by multiplying it by a suitable positive constant, but this may lead to forms with non-integer coefficients. To illustrate this potential difficulty, suppose we take $\Delta = -4$ so $R_\Delta = \mathbb{Z}[i]$. The lattice $L(2, i)$ in $\mathbb{Z}[i]$ yields the form $N(2x + iy) = 4x^2 + y^2$ of discriminant $-16$, but to rescale this to have discriminant $-4$ we would have to take the form $2x^2 + \frac{1}{2}y^2$.

Fortunately this problem does not occur if we consider only lattices that are ideals. By Proposition 8.13 each ideal $L$ in $R_\Delta$ is equal to a multiple $nL_Q = L(na, n\frac{b+\sqrt{\Delta}}{2})$ for some form $Q(x, y) = ax^2 + bxy + cy^2$ of discriminant $\Delta$ with $a > 0$. We have
aQ(x, y) = N(ax + b + \sqrt{\Delta} y), hence \( n^2 aQ(x, y) = N(nax + nb + \sqrt{\Delta} y) \) which is the norm form for \( L \) in the basis \( na, nb + \sqrt{\Delta} \). This basis is positively ordered since \( a > 0 \). By dividing this norm form for \( L \) by \( n^2 a \) we get a form with integer coefficients and discriminant \( \Delta \), namely the form \( Q \). If we change the basis \( na, n b + \sqrt{\Delta} \) for \( L \) to some other positively ordered basis \( \alpha, \beta \), it is still true that the form \( \frac{1}{n^2 a} N(a \alpha + \beta y) \) has integer coefficients and discriminant \( \Delta \) since this just changes \( Q \) to a properly equivalent form.

The scaling factor \( n^2 a \) has a nice geometric interpretation as the number of parallel translates of the lattice \( L \) (including \( L \) itself) that are needed to completely cover the larger lattice \( R_\Delta \). In the special case \( n = 1 \) this is obvious since a basis for \( L \) in this case is \( a, b + \sqrt{\Delta} \) with the latter point lying in the first row of \( R_\Delta \) above the \( x \) axis, so we only need to translate \( L \) horizontally by the numbers \( 0, 1, \cdots, a - 1 \) to cover all of \( R_\Delta \). The case \( n > 1 \) follows from this case since this just amounts to rescaling the whole plane by a factor of \( n \), and it takes \( n^2 \) parallel copies of \( nR_\Delta \) to cover all of \( R_\Delta \).

For a lattice \( L \) in \( R_\Delta \) the number of parallel translates of \( L \) needed to cover all of \( R_\Delta \) is called the norm of \( L \) and written \( N(L) \). Notice that this does not depend on choosing a basis for \( L \). The preceding arguments prove:

**Proposition 8.14.** For an ideal \( L \) in \( R_\Delta \) with basis \( \alpha, \beta \) the form \( \frac{1}{N(L)} N(\alpha x + \beta y) \) has integer coefficients and discriminant \( \Delta \).

For an ideal \( L \) with positively ordered basis \( \alpha, \beta \) the form \( \frac{1}{N(L)} N(\alpha x + \beta y) \) will be denoted by \( Q_L \), although a more precise notation might include \( \alpha \) and \( \beta \) since the form depends on the choice of basis.

Different ideals \( L \) in \( R_\Delta \) can give properly equivalent forms \( Q_L \). Obviously a rescaling \( nL \) of \( L \) gives the same form \( Q_L \). More generally, suppose we multiply all elements of an ideal \( L(\alpha, \beta) \) by a fixed nonzero element \( y \) of \( R_\Delta \) to get a new lattice \( yL(\alpha, \beta) = L(y \alpha, y \beta) \) which is an ideal since \( L(\alpha, \beta) \) was an ideal. Taking norms, we have \( N(y \alpha x + \beta y) = N(y)N(\alpha x + \beta y) \), so if \( N(y) > 0 \) the new form \( N(y \alpha x + \beta y) \) is a rescaling of \( N(\alpha x + \beta y) \), with rescaling factor \( N(y) \). Thus after rescaling to get discriminant \( \Delta \) we have \( Q_{yL} = Q_L \) when \( N(y) > 0 \). As a technical point, we should check that \( y \alpha, y \beta \) is positively oriented if \( \alpha, \beta \) is, which we can do in the following way. In \( \mathbb{Q}(\sqrt{\Delta}) \) we have \((a + b\sqrt{\Delta}) (x + y\sqrt{\Delta}) = (ax + b\Delta y) + (ay + bx)\sqrt{\Delta} \) so multiplication by \( a + b\sqrt{\Delta} \) is a linear transformation with matrix \( \begin{pmatrix} a & b \\ b & a \end{pmatrix} \). This has determinant \( a^2 - b^2 \Delta = N(a + b\sqrt{\Delta}) \) so it preserves orientation when \( N(a + b\sqrt{\Delta}) > 0 \). This means that \( y \alpha, y \beta \) is positively oriented if \( \alpha, \beta \) is positively oriented and \( N(y) > 0 \).

In view of these observations we would like to regard the ideals \( L \) and \( yL \) as equivalent when \( N(y) > 0 \). Now, any reasonable notion of equivalence should have the property that two things equivalent to the same thing are equivalent to each other, but this doesn’t seem to hold for the notion of equivalence that we just considered
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since if two ideals $L$ and $L'$ are equivalent to the same ideal $γL = γ'L'$ for some $γ$ and $γ'$ in $R_Δ$, then it doesn’t seem to follow that $L' = δL$ or $L = δL'$ for some $δ$ in $R_Δ$ since the quotients $γ/γ'$ and $γ'/γ$ might not lie in $R_Δ$.

This difficulty can be avoided by defining two ideals $L$ and $L'$ in $R_Δ$ to be equivalent, written $L \sim L'$, if $γL = γ'L'$ for some nonzero elements $γ, γ'$ in $R_Δ$. If in addition $N(γ) > 0$ and $N(γ') > 0$ then we say $L$ and $L'$ are strictly equivalent and write $L \approx L'$. In particular we have $L \sim γL$ for each $γ \neq 0$ in $R_Δ$ since if we let $L' = γL$ and $γ' = 1$ then the equation $γL = γ'L'$ becomes just $γL = L'$. Similarly $L \approx γL$ for every $γ$ with $N(γ) > 0$.

Conversely, a general equivalence $L \sim L'$ can be realized as a pair of equivalences of the special type originally considered, namely $L \sim γL = γ'L' \sim L'$ and likewise for strict equivalences. Thus we have not really changed the underlying idea by defining $\sim$ and $\approx$ as we did. What we have gained is the property that two things equivalent to the same thing are equivalent to each other, which can be expressed as the assertion that if $L \sim L'$ and $L' \sim L''$ then $L \sim L''$. This holds since if $γL = γ'L'$ and $δL' = δ'L''$ then $δγL = δγ'L' = δ'γ'L''$ so $L \sim L''$. This reasoning also works with $\approx$ in place of $\sim$ by adding the condition that all of $γ, γ', δ, δ'$ have positive norm, hence also all their products.

For negative discriminants there is no difference between equivalence and strict equivalence of ideals since norms of nonzero elements of $R_Δ$ are always positive when $Δ < 0$. For positive discriminants there can be a difference, however:

**Proposition 8.15.** For positive discriminants $Δ$ the relations of equivalence and strict equivalence of ideals in $R_Δ$ are the same if and only if there is a unit in $R_Δ$ of norm $−1$.

Note that it suffices to consider only the fundamental unit since if this has norm $+1$ then all units have norm $+1$.

**Proof:** Suppose there is a unit $ε$ in $R_Δ$ of norm $−1$ and suppose two ideals $L$ and $M$ are equivalent via an equality $αL = βM$. If the norms of $α$ and $β$ are both negative then the equation $α^2L = αβM$ shows that $L$ and $M$ are strictly equivalent. If the norms of $α$ and $β$ have opposite sign, say $N(α) < 0$ and $N(β) > 0$, then $N(εα) > 0$ and $εαL = αL$ so $L$ and $M$ are strictly equivalent via the equality $εαL = βM$. Thus equivalence implies strict equivalence when there is a unit of norm $−1$.

For the converse, suppose equivalence is the same as strict equivalence. Since we assume $Δ > 0$, there exist elements $α$ in $R_Δ$ with $N(α) < 0$. The ideals $R_Δ$ and $αR_Δ$ are equivalent so by hypothesis they are strictly equivalent. This means $βR_Δ = γαR_Δ$ for some elements $β$ and $γ$ in $R_Δ$ of positive norm. Since $β$ is in $βR_Δ = γαR_Δ$ we have $β = γαδ$ for some $δ$ in $R_Δ$. Also $γα$ is in $γαR_Δ = βR_Δ$ so $γα = βε$ for some $ε$ in $R_Δ$. Thus $β = γαδ = βεδ$ and hence $1 = εδ$ since $β \neq 0$. Thus $δ$ and $ε$ are units. The equation $γα = βε$ implies that $N(ε) < 0$ since $N(α) < 0$, $N(β) > 0$, and $N(γ) > 0$. Since $ε$ is a unit, its norm is then $−1$. □
Now we come to the main result in this section:

**Theorem 8.16.** There is a one-to-one correspondence between the set of strict equivalence classes of ideals in \( R_\Delta \) and the set of proper equivalence classes of quadratic forms of discriminant \( \Delta \). Under this correspondence an ideal \( L \) with a positively ordered basis \( \alpha, \beta \) corresponds to the form \( Q_L(x, y) = \frac{1}{N(L)}N(\alpha x + \beta y) \), and a form \( Q(x, y) = ax^2 + bxy + cy^2 \) corresponds to the ideal \( L_Q = L(a, \frac{b + \sqrt{\Delta}}{2}) \). (When \( \Delta < 0 \) we are restricting attention just to forms with positive values, as usual.)

Thus for negative discriminants the proper equivalence classes of forms correspond to the various shapes of lattices that are ideals, up to rotation and rescaling since this is how two lattices \( L \) and \( \alpha L \) are related when \( \alpha \) is a complex number. For example, for the nine negative discriminants where all forms are equivalent (hence properly equivalent), all ideals have the same shape. We already saw this for the discriminants \(-3, -4, -7, -8, -11\) where there is a Euclidean algorithm, but the theorem says it is also true for discriminants \(-19, -43, -67, -163\).

**Proof of Theorem 8.16:** We have seen that there is a well-defined function \( \Phi \) from strict equivalence classes of ideals to proper equivalence classes of forms, induced by sending an ideal \( L \) to the form \( Q_L \). To see that \( \Phi \) is onto note first that in each proper equivalence class of forms there are forms \( Q(x, y) = ax^2 + bxy + cy^2 \) with \( a > 0 \) since the topograph always contains some positive numbers, so we can arrange that \( Q(1, 0) > 0 \). Then \( Q = Q_L \) for the ideal \( L = L_Q = L(a, \frac{b + \sqrt{\Delta}}{2}) \) since \( Q_L = \frac{N(\alpha x + \beta y)}{N(L)}N(L) = ax^2 + bxy + cy^2 \), using the fact that \( N(L) = a \).

To show that \( \Phi \) is one-to-one, suppose we have two ideals \( L \) and \( L' \) with positively ordered bases \( \alpha, \beta \) and \( \alpha', \beta' \) such that the forms \( Q_L \) and \( Q_{L'} \) with respect to these bases are properly equivalent. We can assume the basis \( \alpha, \beta \) is chosen so that \( Q_L(1, 0) > 0 \). Since \( Q_L \) and \( Q_{L'} \) are properly equivalent we can then choose \( \alpha', \beta' \) so that we have actual equality \( Q_L(x, y) = Q_{L'}(x, y) \) for all \( x \) and \( y \).

The norm forms \( N(\alpha x + \beta y) \) and \( N(\alpha' x + \beta' y) \) are rescalings of each other since they rescale to \( Q_L(x, y) \) and \( Q_{L'}(x, y) \) which are equal. We have \( N(\alpha) > 0 \) and \( N(\alpha') > 0 \) since \( Q_L(1, 0) = Q_{L'}(1, 0) > 0 \). Let \( y = \beta/\alpha \) and \( y' = \beta'/\alpha' \), elements of \( \mathbb{Q}(\sqrt{\Delta}) \). We have \( N(\alpha x + \beta y) = N(\alpha)N(x + yy) \) and likewise \( N(\alpha' x + \beta' y) = N(\alpha')N(x + y'y) \) so the two forms \( N(x + yy) = N(\alpha x + \beta y)/N(\alpha) \) and \( N(x + y'y) = N(\alpha' x + \beta' y)/N(\alpha') \) are also rescalings of each other. Note that these two forms have rational coefficients, not necessarily integers. Since the forms \( N(x + yy) \) and \( N(x + y'y) \) are rescalings of each other and take the same value at \( (x, y) = (1, 0) \), namely \( N(1) = 1 \), they must actually be equal.

Next we show that in fact \( y = y' \). Let \( y = r + s\sqrt{\Delta} \) and \( y' = r' + s'\sqrt{\Delta} \) with \( r, s, r', s' \) in \( \mathbb{Q} \). Consider the difference \( N(1+y) - N(y) \) which equals \((r+1)^2 - s^2\Delta - (r'^2 - s'^2\Delta) = 2r + 1 \). Similarly \( N(1+y') - N(y') = 2r' + 1 \). Since the forms \( N(x + yy) \) and \( N(x + y'y) \) are equal we must therefore have \( r = r' \). We also have \( N(y) = N(y') \).
by evaluating both forms at \((x, y) = (0, 1)\). Thus \(r^2 - s^2\Delta = r'^2 - s'^2\Delta\), and from \(r = r'\) we deduce that \(s = \pm s'\). The bases 1, \(y\) and 1, \(y'\) are positively ordered since this was true for \(\alpha, \beta\) and \(\alpha', \beta'\) and multiplication by \(\alpha\) and \(\alpha'\) preserves orientation of the plane since \(N(\alpha) > 0\) and \(N(\alpha') > 0\). Since both 1, \(y\) and 1, \(y'\) are positively ordered we must have \(s = +s'\). Thus \(y = y'\) as claimed.

The lattice \(L(1, y)\) may not lie in \(R_\Delta\) since \(y\) is only an element of \(\mathbb{Q}(\sqrt{\Delta})\), but we can rescale \(L(1, y)\) to a lattice \(nL(1, y) = L(n, ny)\) in \(R_\Delta\) by multiplying by a positive integer \(n\) such that \(ny\) is in \(R_\Delta\). Then \(nL(\alpha, \beta) = \alpha L(n, ny)\) since \(L(\alpha, \beta) = \alpha L(1, y)\). Likewise we have \(nL(\alpha', \beta') = \alpha' L(n, ny')\) with the same scaling factor \(n\) since \(y = y'\). Thus we have \(L = L(\alpha, \beta) \approx L(n, ny) = L(n, ny') \approx L(\alpha', \beta') = L'\). This finishes the proof that \(\Phi\) is one-to-one.

The correspondence between forms and ideals includes nonprimitive forms as well as primitive forms, but the ideals corresponding to primitive and nonprimitive forms behave somewhat differently. Let us illustrate this by the example of discriminant \(\Delta = -12\) where there are two equivalence classes of forms, given by the primitive form \(x^2 + 3y^2\) and the nonprimitive form \(2x^2 + 2xy + 2y^2\). The ideal for \(x^2 + 3y^2\) is \(L(1, \sqrt{-3}) = \mathbb{Z}[\sqrt{-3}] = R_\Delta\) while for \(2x^2 + 2xy + 2y^2\) the ideal is \(L(2, 1 + \sqrt{-3})\).

\[\begin{array}{cccccccccccc}
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The ideal for \(2x^2 + 2xy + 2y^2\) is a lattice of equilateral triangles, and it has the special property that it is taken to itself not just by multiplication by elements of \(\mathbb{Z}[\sqrt{-3}]\) but also by the 60 degree rotation given by multiplication by the element \(\omega = (1 + \sqrt{-3})/2\) in the larger ring \(\mathbb{Z}[\omega]\) which is \(R_\Delta\) for \(\Delta = -3\). Hence the lattice \(L(2, 1 + \sqrt{-3})\) is taken to itself by all elements of \(\mathbb{Z}[\omega]\) and so this lattice is an ideal in \(\mathbb{Z}[\omega]\), not just in the original ring \(\mathbb{Z}[\sqrt{-3}] = R_\Delta\).

More generally, suppose we start with a form \(Q = ax^2 + bxy + cy^2\) of discriminant \(\Delta\) and then consider the nonprimitive form \(kQ = kax^2 + kbx + kcy^2\) of discriminant \(k^2\Delta\) for some integer \(k > 1\). The associated ideal \(L_{kQ}\) is then \(L(ka, \frac{kb + k\sqrt{\Delta}}{2}) = kL(a, \frac{b + \sqrt{\Delta}}{2}) = kL_Q\). This is an ideal not just in \(R_{k^2\Delta}\) but also in the larger ring \(R_\Delta\) since it is \(k\) times an ideal in \(R_\Delta\), namely \(k\) times \(L_Q\).

Let us say that an element \(\alpha\) in \(\mathbb{Q}(\sqrt{\Delta})\) stabilizes an ideal \(L\) in \(R_\Delta\) if \(\alpha L\) is contained in \(L\), and let us call the set of all such elements \(\alpha\) the stabilizer of \(L\). The stabilizer of \(L\) contains \(R_\Delta\) and is a ring itself since if two elements \(\alpha\) and \(\beta\) in \(\mathbb{Q}(\sqrt{\Delta})\) stabilize \(L\) then so do \(\alpha \pm \beta\) and \(\alpha \beta\). If the stabilizer of \(L\) is exactly \(R_\Delta\) then we will say that \(L\) is stable.
Proposition 8.17. A form $Q$ of discriminant $\Delta$ is primitive if and only if the corresponding ideal $L_Q$ in $R_\Delta$ is stable.

Proof: We observed above that a nonprimitive form $Q$ of discriminant $\Delta$ gives an ideal $L_Q$ with stabilizer larger than $R_\Delta$. For the converse we wish to show that if $Q = ax^2 + bxy + cy^2$ is a primitive form of discriminant $\Delta$ then $L_Q$ is not an ideal in any larger ring than $R_\Delta$ in $\mathbb{Q}(\sqrt{\Delta})$. Let us write $L_Q = L(a, \tau)$ for $\tau = (b + \sqrt{\Delta})/2$.

Note that $R_\Delta = \mathbb{Z}[\tau]$ since $b$ has the same parity as $\Delta$. Also $\mathbb{Q}(\sqrt{\Delta}) = \mathbb{Q}(\tau)$.

Suppose we have an element $\alpha = r + s\tau$ in $\mathbb{Q}(\tau)$ such that $\alpha L(a, \tau)$ is contained in $L(a, \tau)$. Here $r$ and $s$ are rational numbers. Our goal is to show that $Q$ being primitive forces $r$ and $s$ to be integers. This will say that $\alpha$ is in $\mathbb{Z}[\tau] = R_\Delta$, and hence that $R_\Delta$ is the stabilizer of $L(a, \tau)$.

For $\alpha L(a, \tau)$ to be contained in $L(a, \tau)$ means that both $\alpha a$ and $\alpha \tau$ are in $L(a, \tau)$. We have $\alpha a = ra + sa\tau$, and for this to be in $L(a, \tau)$ which consists of the linear combinations $xa + y\tau$ with $x$ and $y$ integers means that $r$ is an integer and $sa$ is an integer. It remains to see that $s$ itself is an integer, using the condition that $\alpha \tau$ is in $L(a, \tau)$.

To compute $\alpha \tau$ we use the fact that $\tau$ is a root of the equation $x^2 - bx + ac = 0$ so $\tau^2 = b\tau - ac$. Then we have

$$\alpha \tau = r\tau + s\tau^2 = r\tau + s(b\tau - ac) = sac + (r + sb)\tau$$

For this to be in $L(a, \tau)$ means that $sc$ and $r + sb$ are integers. We already know that $r$ is an integer, so $r + sb$ being an integer is equivalent to $sb$ being an integer. Thus we know that all three of $sa$, $sb$, and $sc$ are integers. From this we can deduce that $s$ is an integer since $a$, $b$, and $c$ have no common divisor greater than 1 by the assumption that the form $Q$ is primitive. Namely, let us write $s$ as a fraction $\frac{m}{n}$ in lowest terms. Then $sa$ being an integer says that $n$ divides $a$. Similarly $sb$ and $sc$ being integers says that $n$ divides $b$ and $c$. But 1 is the only common divisor of $a$, $b$ and $c$ so $n = 1$. Thus $s$ is an integer and we are done. \(\square\)

Let us go into a little more detail about the shapes of lattices in the plane. This is mostly relevant to the case of negative discriminants. Recall that we say two lattices have the same shape if one can be transformed into the other by rotation and rescaling of the plane. With this definition of shape one can ask whether it is possible to characterize exactly all the different shapes of lattices. We will give such a characterization and see how this relates to forms of negative discriminant.

Given a lattice $L$, choose a point in $L$ that is closest to the origin, other than the origin itself. We can rotate $L$ about the origin until this point lies on the positive $x$ axis, and then we can rescale $L$ until this point is at distance 1 from the origin, so it is the point $(1, 0)$, or in other words the complex number 1. Now choose a point $\alpha$ in $L$ closest to the origin among all points of $L$ above the $x$ axis. Thus $\alpha$ lies on or outside
the unit circle $x^2 + y^2 = 1$. Also, $\alpha$ lies in the vertical strip consisting of points $x + yi$ with $-1/2 \leq x \leq 1/2$, otherwise there would be another point of $L$ inside this strip that had the same $y$ coordinate as $\alpha$ and was closer to the origin than $\alpha$. This is because all points of $L$ lie in horizontal rows of points of distance 1 apart. The lattice $L(1, \alpha)$ is contained in $L$ and in fact must equal $L$ by the way that we chose $\alpha$. (There are no other points of $L$ above the $x$ axis and inside the circle $x^2 + y^2 = r^2$ passing through $\alpha$.)

Let $R$ be the region of the plane consisting of the points $\alpha$ as above, that is, all $\alpha = x + yi$ with $x^2 + y^2 \geq 1$, $-1/2 \leq x \leq 1/2$, and $y > 0$.

**Proposition 8.18.** The lattices $L(1, \alpha)$ with $\alpha$ in $R$ realize all lattice shapes, and of these lattices the only ones having the same shape are the pairs $L(1, 1/2 + yi)$ and $L(1, -1/2 + yi)$ and the pairs $L(1, x + yi)$ and $L(1, -x + yi)$ with $x^2 + y^2 = 1$.

Note that these pairs all lie on the boundary of $R$, either on the vertical edges or on the circular arc forming the lower edge of $R$. The two points of each pair are mirror reflections of each other across the $y$ axis.

**Proof:** We have already seen that all lattices have the shape of a lattice $L(1, \alpha)$ for some $\alpha$ in $R$, and it remains to see when two of these lattices $L(1, \alpha)$ have the same shape. A more basic question is when two of the lattices $L(1, \alpha)$ and $L(1, \beta)$ with $\alpha$ and $\beta$ in $R$ are the same lattice. If this happens, the $y$ coordinates of $\alpha$ and $\beta$ must be the same since this is the coordinate of points in the first row of the lattice above the $x$ axis. The $x$ coordinates of $\alpha$ and $\beta$ must then differ by an integer if $L(1, \alpha) = L(1, \beta)$, so if $\alpha$ and $\beta$ are both in $R$ the only possibility is that $\alpha$ and $\beta$ are points $1/2 + yi$ and $-1/2 + yi$ on the two vertical edges of $R$.

For $L(1, \alpha)$ and $L(1, \beta)$ to have the same shape means that there is a rotation and rescaling taking one to the other. However, there can be no rescaling since the smallest distance from nonzero points in these two lattices to the origin is 1 in both cases. To see what sorts of rotations are possible consider the subsets $C_\alpha$ of $L(1, \alpha)$ and $C_\beta$ of $L(1, \beta)$ consisting of the lattice points at distance 1 from the origin. If there is a rotation taking $L(1, \alpha)$ to $L(1, \beta)$ then this rotation carries $C_\alpha$ onto $C_\beta$. In particular, $C_\alpha$ and $C_\beta$ must have the same number of points. The points 1 and $-1$ always belong to $C_\alpha$ and $C_\beta$. If these are the only points in $C_\alpha$ and $C_\beta$ then the only rotations taking $C_\alpha$ to $C_\beta$ are rotations by 0 and 180 degrees, but these do not affect the lattices so we must have $L(1, \alpha) = L(1, \beta)$ in this case. If $C_\alpha$ and $C_\beta$ have more than two points then these points will include $\pm \alpha \in C_\alpha$ and $\pm \beta \in C_\beta$. If $C_\alpha = \{\pm 1, \pm \alpha\}$ and $C_\beta = \{\pm 1, \pm \beta\}$ then the only way for $C_\alpha$ to be a rotation of $C_\beta$ is for the two arcs of the unit circle from $\alpha$ to 1 and $-1$ to have the same lengths as the
two arcs from \( \beta \) to 1 and \(-1\). This implies that \( \beta \) is equal to either \( \alpha \) or the reflection of \( \alpha \) across the \( y \) axis. Thus \( L(1, \alpha) \) and \( L(1, \beta) \) are \( L(1, x + yi) \) and \( L(1, -x + yi) \) for some \( x \) and \( y \) with \( x^2 + y^2 = 1 \).

The remaining possibility is that \( C_\alpha \) and \( C_\beta \) contain more that four points, but this only happens when they are the vertices of regular hexagons inscribed in the unit circle since the points of \( C_\alpha \) must be of distance at least 1 apart, and likewise for \( C_\beta \). In this hexagonal case we have \( L(1, \alpha) = L(1, \beta) \), finishing the proof. 

Let us see now how the lattices \( L_Q = L(a, \frac{b+\sqrt{\Delta}}{2a}) \) associated to elliptic forms \( Q = ax^2 + bxy + cy^2 \) fit into this picture. Here \( a \) and \( c \) are positive since we only consider positive elliptic forms. Rescaling \( L(a, \frac{b+\sqrt{\Delta}}{2a}) \) to a lattice \( L(1, \alpha) \) produces \( L(1, \frac{b+\sqrt{\Delta}}{2a}) \). For \( \frac{b+\sqrt{\Delta}}{2a} \) to be in the region \( R \) means two things: that its norm \( \frac{b^2 - \Delta}{4a^2} = \frac{c}{a} \) should be greater than or equal to 1, so \( a \leq c \), and its real part \( \frac{b}{2a} \) should be between \(-\frac{1}{2}\) and \( \frac{1}{2} \), so \( |b| \leq a \). Thus for \( \frac{b+\sqrt{\Delta}}{2a} \) to be in the right half of \( R \) just amounts to the inequalities \( 0 \leq b \leq a \leq c \) that define what it means for the form \( ax^2 + bxy + cy^2 \) to be reduced. The left half of \( R \) is then obtained by changing the sign of \( b \), taking the mirror image form.

The reduced forms with mirror symmetric topographs are those where one of the inequalities \( 0 \leq b \leq a \leq c \) becomes an equality. When \( b = 0 \) we have the forms \( ax^2 + cy^2 \) corresponding to the lattices \( L(1, \frac{\sqrt{\Delta}}{2a}) \) along the \( y \) axis in \( R \). These are the rectangular lattices, with mirror symmetry across the \( y \) axis. When \( b = a \) we have the forms \( ax^2 + axy + cy^2 \) whose associated lattices \( L(1, \frac{a+\sqrt{\Delta}}{2a}) = L(1, \frac{1}{2} + \frac{\sqrt{\Delta}}{2a}) \) lie along the right-hand edge of \( R \). These lattices also have mirror symmetry across the \( y \) axis since they equal their mirror image lattices \( L(1, -\frac{1}{2} + \frac{\sqrt{\Delta}}{2a}) \). Finally, if \( a = c \) we have forms \( ax^2 + bxy + ay^2 \) corresponding to lattices \( L(1, \frac{b+\sqrt{\Delta}}{2a}) \) with \( \frac{b+\sqrt{\Delta}}{2a} \) on the arc of the unit circle forming the bottom border of \( R \) where the norm \( \frac{c}{a} \) of \( \frac{b+\sqrt{\Delta}}{2a} \) equals 1. These lattices also have mirror symmetry since they form grids of rhombuses, the distances from both generators 1 and \( \frac{b+\sqrt{\Delta}}{2a} \) to the origin being equal.

Thus forms with mirror symmetric topographs give rise to mirror symmetric lattices. The converse is also true since none of the lattices \( L(1, \alpha) \) with \( \alpha \) in the interior of \( R \) but not on the \( y \) axis have mirror symmetry. One can see this by noting that for points \( \alpha \) in the interior of \( R \) the only points in lattices \( L(1, \alpha) \) of unit distance apart lie on horizontal lines, so mirror symmetries of these lattices must take horizontal lines to horizontal lines, which forces these symmetries to be reflections across either horizontal or vertical lines. The only time such a reflection takes a lattice \( L(1, \alpha) \) to itself for some \( \alpha \) in the interior of \( R \) is when \( \alpha \) is on the \( y \) axis, so the lattice is rectangular.

Not all lattices \( L(1, \alpha) \) associated to points \( \alpha \) in \( R \) come from forms. When
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$L(1, \alpha)$ comes from a form $ax^2 + bxy + cy^2$ we have $\alpha = \frac{b + \sqrt{\Delta}}{2a}$ so the $x$ coordinate $\frac{b}{2a}$ is a rational number. Also, the norm of $\alpha$ which is the square of the distance to the origin ranges over all real numbers $r \geq 1$ as $\alpha$ varies over $R$, but for a lattice $L(1, b + \frac{\sqrt{\Delta}}{2a})$ the norm of $b + \frac{\sqrt{\Delta}}{2a}$ is $\frac{2}{\alpha}$, again a rational number. Thus if either the norm of $\alpha$ or its $x$ coordinate is irrational then $L(1, \alpha)$ does not come from a form. However, the points $b + \frac{\sqrt{\Delta}}{2a}$ coming from forms are dense in $R$, by the following argument. A point $\alpha$ in $R$ is determined by specifying its norm $N(\alpha)$ and its $x$ coordinate. For an arbitrary $\alpha$ in $R$ we can approximate $N(\alpha)$ as closely as we like by a rational number $r \geq 1$ and the $x$ coordinate of $\alpha$ by a rational number $s$ between $-\frac{1}{2}$ and $\frac{1}{2}$. After finding a common denominator $a$ for $r$ and $s$ we can write them as $r = \frac{c}{a}$ and $s = \frac{b}{2a}$, doubling the numerator of $s$ to make its denominator $2a$. Then $\frac{b + \sqrt{\Delta}}{2a}$ will be close to $\alpha$, for $\Delta = b^2 - 4ac$.

It is interesting to compare the picture of the region $R$ with a figure in Section 5.6 showing the location of reduced elliptic forms in a triangle inside the Farey diagram. Here is the relevant part of this figure, first as it appeared in Chapter 5 and then reflected across a 45 degree line:

The three sides of the triangle are specified by the equations $a = c$, $a = b$, and $b = 0$, so we see that the triangle corresponds exactly to the right half of the region $R$, with the edge $a = b$ corresponding to the right edge of $R$, the edge $a = c$ to an arc of the unit circle, and the edge $b = 0$ to the central vertical axis of $R$.

For negative discriminants the relation of strict equivalence of ideals corresponds geometrically to rotation and rescaling of lattices. There is an analogous interpretation for positive discriminants but it involves replacing rotations by somewhat more complicated motions of the plane, as we shall now see.

What we want is a geometric description of the transformation $T_y$ of $Q(\sqrt{\Delta})$ defined by multiplying by a fixed nonzero element $\gamma$, so $T_y(\alpha) = \gamma \alpha$. For a positive discriminant $\Delta$ we are regarding $Q(\sqrt{\Delta})$ as a subset of the plane by giving an element $\alpha = a + b \sqrt{\Delta}$ the coordinates $(x, y) = (a, b \sqrt{\Delta})$. The norm $N(\alpha) = a^2 - \Delta b^2$ is then equal to $x^2 - y^2$ and $T_y$ takes each hyperbola $x^2 - y^2 = k$ to a hyperbola $x^2 - y^2 = N(y)k$ since $N(\gamma \alpha) = N(\gamma)N(\alpha)$.

To understand linear transformations of the plane that take hyperbolas $x^2 - y^2 =
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k to hyperbolas $x^2 - y^2 = k'$ it is convenient to changes the coordinates \(x\) and \(y\) to \(X = x + y\) and \(Y = x - y\). This changes the hyperbolas $x^2 - y^2 = k$ to the hyperbolas $XY = k$ whose asymptotes are the \(X\) and \(Y\) axes, at a 45 degree angle from the \(x\) and \(y\) axes.

Notice that since \((x, y) = (a, b\sqrt{\Delta})\), the coordinate \(X = x + y\) is just \(a + b\sqrt{\Delta}\), the real number \(\alpha\) we started with, while \(Y = x - y = a - b\sqrt{\Delta}\) is its conjugate \(\overline{\alpha}\).

The transformation \(T_y\) sends \(\alpha\) to \(\gamma\alpha\) so \(T_y\) multiplies the \(X\) coordinate \(\alpha\) by \(y\). To see how \(T_y\) acts on the \(Y\) coordinate, observe that since the \(Y\) coordinate of \(T_y(\alpha)\) is \(\overline{\gamma\alpha} = \overline{\gamma\alpha}\), which means that the \(Y\) coordinate of \(T_y(\alpha)\) is \(\overline{\gamma}\) times the \(Y\) coordinate of \(\alpha\). Thus \(T_y\) multiplies the \(Y\) coordinate by \(\overline{\gamma}\), so we have the simple formula \(T_y(X, Y) = (\gamma X, \gamma Y)\).

A consequence of the formula \(T_y(X, Y) = (\gamma X, \gamma Y)\) is that \(T_y\) takes the \(X\) axis to itself since the \(X\) axis is the points \((X, Y)\) with \(Y = 0\). Similarly, \(T_y\) takes the \(Y\) axis to itself, the points where \(X = 0\). In general, linear transformations taking both the \(X\) and \(Y\) axes to themselves have the form \(T(X, Y) = (\lambda X, \mu Y)\) for real constants \(\lambda\) and \(\mu\). As a special case, when \(\mu = \lambda^{-1}\) the transformations \(T(X, Y) = (\lambda X, \lambda^{-1} Y)\) take each hyperbola \(XY = k\) to itself. When \(\lambda > 1\) the \(X\) axis is stretched by a factor of \(\lambda\) and the \(Y\) axis is shrunk by \(\lambda\). Thus we are sliding each hyperbola along itself in the direction indicated by the arrows in the figure above. When \(\lambda\) is between 0 and 1 the situation is reversed and the \(Y\) axis is stretched while the \(X\) axis is shrunk.

When \(\lambda > 0\) and \(\mu > 0\) we can rescale the transformation \(T(X, Y) = (\lambda X, \mu Y)\) to \(\frac{1}{\sqrt{\lambda\mu}} T(X, Y) = (\sqrt{\lambda\mu}X, \sqrt{\mu\lambda}Y)\) which is a transformation of the type considered in the preceding paragraph, sliding each hyperbola along itself. Thus a transformation \(T(X, Y) = (\lambda X, \mu Y)\) with \(\lambda\) and \(\mu\) positive is a composition of a 'hyperbola-slide' and a rescaling. This is analogous to compositions of rotations and rescalings in the situation of negative discriminants. Allowing \(\lambda\) or \(\mu\) to be negative then allows reflections across the \(X\) or \(Y\) axes as well. If both \(\lambda\) and \(\mu\) are negative the composition of these two reflections is a 180 degree rotation of the plane.

Now we specialize to the situation of a transformation \(T_y\) of \(R_{\Delta}\) given by multiplication by an element \(\gamma\) in \(R_{\Delta}\) with \(N(\gamma) > 0\). The condition \(N(\gamma) > 0\) implies that \(T_y\) preserves the orientation of the plane and also the sign of the norm so it
either takes each quadrant of the \( XY \) plane (north, south, east, or west) to itself or to the opposite quadrant. In the former case \( T_γ \) is a composition of a hyperbola-slide and a rescaling, while in the latter case there is also a composition with a 180 degree rotation of the plane, which is just \( T_γ \) for \( γ = -1 \). The sign of \( γ \) distinguishes these two cases since if \( γ > 0 \) the transformation \( T_γ \) takes positive numbers to positive numbers so the positive \( X \) axis goes to itself, while if \( γ < 0 \) the positive \( X \) axis goes to the negative \( X \) axis.

If \( γ \) is a unit with \( N(γ) = +1 \) then each hyperbola \( x^2 - y^2 = k \) is taken to itself by \( T_γ \). The two branches of the hyperbola are distinguished by the sign of \( X \), so if \( γ \) is positive then \( T_γ \) slides each branch along itself while if \( γ \) is negative this slide is combined with a 180 degree rotation of the plane. If we choose \( γ \) to be the smallest unit greater than 1 with \( N(γ) = +1 \) then the powers \( γ^n \) for integers \( n \) lie along the right-hand branch of the hyperbola \( x^2 - y^2 = 1 \), becoming farther and farther apart as one moves away from the origin, and \( T_γ \) slides each one of these points along the hyperbola to the next one, increasing the \( X \) coordinate. The case \( Δ = 12 \) is shown in the first figure below, with \( R_Δ = \mathbb{Z}[\sqrt{3}] \). The basic unit \( γ \) is \( 2 + \sqrt{3} \), and the figure shows the units \( \pm γ^n \) for \( |n| \leq 2 \) positioned along the hyperbola \( x^2 - y^2 = 1 \), with \( y^2 = 7 + 4\sqrt{3} \) in the upper right corner of the figure.

For some discriminants there are units \( γ \) with \( N(γ) = -1 \) in addition to those with \( N(γ) = +1 \). The transformation \( T_γ \) for the smallest \( γ > 1 \) of norm \(-1\) is a composition of a hyperbola slide and reflection across the \( X \) axis. The powers \( γ^n \) then lie alternately on \( x^2 - y^2 = +1 \) and \( x^2 - y^2 = -1 \). This happens for example in \( \mathbb{Z}[\sqrt{2}] \) with \( γ = 1 + \sqrt{2} \) as shown in the second figure above, where \( y^2 = 3 + 2\sqrt{2} \) and \( y^3 = 7 + 5\sqrt{2} \).

Each ideal in \( R_Δ \) is taken into itself by the transformations \( T_γ \) for \( γ \) in \( R_Δ \), but when \( γ \) is a unit each ideal is taken \textit{onto} itself since the inverse transformation \((T_γ)^{-1}\) is just \( T_{γ^{-1}} \) which also takes the ideal to itself. Thus all ideals in \( R_Δ \) have “hyperbolic symmetries”, the hyperbola-preserving transformations \( T_γ \) for units \( γ \).

Although we can describe geometrically in terms of hyperbola slides and rescaling how the ideals corresponding to properly equivalent quadratic forms of positive
discriminant are related, the result is somehow less satisfying than in the negative discriminant case. Hyperbola slides are not nearly as simple visually as rotations, making it harder to see at a glance whether two lattices are related by hyperbola slides and rescaling or not. This may be a reflection of the fact that hyperbolic forms do not have a canonical reduced form as elliptic forms do, making it a little more difficult to determine whether two hyperbolic forms are equivalent.

4. The Ideal Class Group

An important feature of ideals is that there is a natural way to define a multiplicative structure in the set of all ideals in $R_{\Delta}$. Thus every pair of ideals $L$ and $M$ in $R_{\Delta}$ has a product $LM$ which is again an ideal in $R_{\Delta}$. We will see that this leads to a group structure on the set of proper equivalence classes of stable ideals, which, under the correspondence between ideals and forms, turns out to be the same as the group structure on the class group of forms studied in the previous chapter. If the procedure for defining the product of forms seemed perhaps a little complicated, the viewpoint of ideals provides an alternative that may seem more obvious and direct.

In order to form the product $LM$ of two ideals $L$ and $M$ in $R_{\Delta}$ one's first guess might be to let $LM$ consist of all products $\alpha \beta$ of elements $\alpha$ in $L$ and $\beta$ in $M$. This sometimes works but not always, as we will see in an example later. The difficulty is that for two products $\alpha_1 \beta_1$ and $\alpha_2 \beta_2$ the sum $\alpha_1 \beta_1 + \alpha_2 \beta_2$ might not be equal to a product $\alpha \beta$ of an element of $L$ with an element of $M$, as it would have to be if the set of all products $\alpha \beta$ was to be an ideal. This difficulty can be avoided by defining $LM$ to be the set of all sums $\sum_i \alpha_i \beta_i$ with each $\alpha_i$ in $L$ and each $\beta_i$ in $M$. With this definition $LM$ is obviously closed under addition as well as subtraction. Also, multiplying such a sum $\sum_i \alpha_i \beta_i$ by an element $\gamma$ in $R_{\Delta}$ gives an element of $LM$ since $\gamma \sum_i \alpha_i \beta_i = \sum_i (\gamma \alpha_i) \beta_i$ and the latter sum is in $LM$ since each term $\gamma \alpha_i$ is in $L$ because $L$ is an ideal. To finish the verification that $LM$ is an ideal we need to check that it is a lattice since we defined ideals in $R_{\Delta}$ to be lattices that are taken to themselves by multiplication by arbitrary elements of $R_{\Delta}$. To check that $LM$ is a lattice we need to explain a few more things about lattices.

We defined a lattice in $R_{\Delta}$ to be a set $L(\alpha, \beta)$ of elements $x\alpha + y\beta$ as $x$ and $y$ range over all integers, where $\alpha$ and $\beta$ are two fixed nonzero elements of $R_{\Delta}$ that do not lie on the same line through the origin. More generally we could define $L(\alpha_1, \cdots, \alpha_n)$ to be the set of all linear combinations $x_1 \alpha_1 + \cdots + x_n \alpha_n$ with coefficients $x_i$ in $\mathbb{Z}$, where not all the $\alpha_i$’s lie on the same line through the origin (so in particular at least two $\alpha_i$’s are nonzero). It is not immediately obvious that $L(\alpha_1, \cdots, \alpha_n)$ is a lattice, but this is true and can be proved by the following procedure which also gives a way of computing what the lattice is.

There are three ways in which the set of generators $\alpha_i$ for $L(\alpha_1, \cdots, \alpha_n)$ can be modified without changing the set $L(\alpha_1, \cdots, \alpha_n)$:
(1) Replace one generator \( \alpha_i \) with \( \alpha_i + k\alpha_j \), adding an integer \( k \) times some other \( \alpha_j \) to \( \alpha_i \).

(2) Replace some \( \alpha_i \) by \(-\alpha_i\).

(3) Interchange two generators \( \alpha_i \) and \( \alpha_j \), or more generally permute the \( \alpha_i \)'s in any way.

After a modification of type (1) each integer linear combination of the new generators is also a linear combination of the old generators so the new \( L(\alpha_1, \ldots, \alpha_n) \) is a subset of the old one, but the process can be reversed by another type (1) operation subtracting \( k\alpha_j \) from the new \( \alpha_i \) so the new \( L(\alpha_1, \ldots, \alpha_n) \) also contains the old one hence must equal it. For the operations (2) and (3) this is also true, more obviously.

**Lemma 8.19.** By applying some sequence of operations (1)–(3) to a set of generators \( \alpha_i \) for \( L(\alpha_1, \ldots, \alpha_n) \) it is always possible to produce a new set of generators \( \beta_1, \ldots, \beta_n \) which are all zero except for \( \beta_1 \) and \( \beta_2 \). In particular \( L(\alpha_1, \ldots, \alpha_n) \) is a lattice.

**Proof:** Let us write \( R_\Delta \) as \( \mathbb{Z}[\tau] \) in the usual way. Each \( \alpha_i \) can be written as \( a_i + b_i\tau \) for integers \( a_i \) and \( b_i \). We can then form a \( 2 \times n \) matrix \( \begin{pmatrix} a_1 & \cdots & a_n \\ b_1 & \cdots & b_n \end{pmatrix} \) whose columns correspond to the \( \alpha_i \)'s. The operations (1)–(3) correspond to adding an integer times one column to another column, changing the sign of a column, and permuting columns.

Our aim is to use these three column operations to simplify the matrix until only the first two columns are nonzero. First we focus on the second row. This must have a nonzero entry since the \( \alpha_i \)'s are not all contained in the \( x \) axis. The nonzero entries in the second row can be made all positive by changing the sign of some columns. Choose a column with smallest positive entry \( b_i \). By subtracting suitable multiples of this column from other columns with positive \( b_j \)'s we can make all other \( b_j \)'s either zero or positive integers smaller than \( b_i \). This process can be repeated using columns with successively smaller second entries until only one nonzero \( b_i \) remains. Switching this column with the first column, we can then assume that \( b_i = 0 \) for all \( i > 1 \).

Now we do the same procedure for columns 2 through \( n \) using the entries \( a_i \) rather than \( b_i \). Since these columns have \( b_i = 0 \), nothing changes in the second row. After this step is finished, only the first two columns will be nonzero. Note that neither of these columns can have both entries zero since otherwise \( L(\alpha_1, \ldots, \alpha_n) \) would be entirely contained in a line through the origin.

If we apply the procedure just described to a lattice \( L = L(\alpha, \beta) \) that already has just two generators, it will produce a \( 2 \times 2 \) matrix with one entry in the second row equal to zero, so we may assume the matrix is \( \begin{pmatrix} a & b \\ 0 & c \end{pmatrix} \), and we can further assume that \( a > 0 \) and \( c > 0 \). By adding a suitable integer multiple of the first column to the
second column we can then arrange that $0 \leq b < a$. This gives a basis $a, b + c\tau$ for $L$ that is called a reduced basis. A reduced basis is unique since $a$ is the smallest positive integer in $L$ and the first row of $L$ above the $x$-axis is in the $c^{th}$ row of $\mathbb{Z}[\tau]$, with the elements of $L$ in this row equally spaced $a$ units apart so there is a unique such element $b + c\tau$ with $0 \leq b < a$. Thus one can determine whether two lattices are equal by computing a reduced basis for each and seeing whether these are equal. (The reader might compare the procedure we have just described with what we did in the first paragraph of the proof of Proposition 8.13 by a more geometric argument.)

For a lattice $L$ with reduced basis $a, b + c\tau$ it is easy to compute the norm $N(L)$ as the product $ac$. This can be seen in two steps. First, it takes $a$ horizontal translates of $L$ to cover all the integers on the $x$-axis, and these translates also cover the $c^{th}$ row of $R_\Delta$ as well as all rows labeled by an integer multiple of $c$. All the rows between these rows contain no elements of $L$ so to cover all of $R_\Delta$ it then takes $c$ translates in the direction of $\tau$ of all the rows covered so far. Thus a total of $ac$ translates of $L$ are needed to cover all of $R_\Delta$, which means that $N(L) = ac$.

Notice that $ac$ is the determinant of the matrix $\left( \begin{array}{cc} a & b \\ 0 & c \end{array} \right)$ associated to the reduced basis $a, b + c\tau$. More generally for an arbitrary basis $\alpha = a + b\tau, \beta = c + d\tau$ for $L = L(\alpha, \beta)$ we have $N(L) = |ad - bc|$, the absolute value of the determinant of the associated matrix $\left( \begin{array}{cc} a & c \\ b & d \end{array} \right)$. This is because the operations (1)–(3) that transform an arbitrary basis into a reduced basis leave the determinant unchanged, up to sign, and $N(L)$ is always positive by its definition as the number of translates of $L$ needed to cover $R_\Delta$.

Let us restrict attention now to lattices that are ideals. One way to generate such a lattice is to start with elements $\alpha_1, \cdots, \alpha_n$ in $R_\Delta$ which can assume are nonzero and then consider the set of all elements $\sum_i y_i \alpha_i$ for arbitrary coefficients $y_i$ in $R_\Delta$ rather than just taking integer coefficients as we would be doing for the lattice $L(\alpha_1, \cdots, \alpha_n)$. The usual notation for this set of all sums $\sum_i y_i \alpha_i$ is $(\alpha_1, \cdots, \alpha_n)$. As a lattice this is the same as $L(\alpha_1, \alpha_1\tau, \alpha_2\tau, \cdots, \alpha_n\tau)$ where $R_\Delta = \mathbb{Z}[\tau]$ since each coefficient $y_i$ can be written as $x_i + y_i\tau$ for integers $x_i$ and $y_i$. To be sure that $(\alpha_1, \cdots, \alpha_n)$ really is a lattice, even when $n = 1$, we can check that $\alpha_1$ and $\alpha_1\tau$ do not lie on the same line through the origin. To see that this is so, observe first that all points of the full lattice $R_\Delta$ that lie on the line through the origin passing through $\alpha_1$ are rational multiples of $\alpha_1$. However, if $\alpha_1\tau$ were a rational multiple $r\alpha_1$ we would have $\tau = r$ since we assume $\alpha_1$ is not zero. However this is impossible since $\tau$ is not a rational number.

Observe that if a lattice $L(\alpha_1, \cdots, \alpha_n)$ is an ideal, then $L(\alpha_1, \cdots, \alpha_n)$ is equal to $(\alpha_1, \cdots, \alpha_n)$ since every product $y\alpha_i$ with $y$ in $R_\Delta$ can be rewritten as an integer linear combination of $\alpha_1, \cdots, \alpha_n$ if $L(\alpha_1, \cdots, \alpha_n)$ is an ideal. A consequence of this, using Lemma 8.19, is that every ideal $(\alpha_1, \cdots, \alpha_n)$ with $n > 2$ can be rewritten as an ideal $(\beta_1, \beta_2)$.
While every ideal in \( R_\Delta \) can be written as \((\alpha, \beta)\) with two generators, only certain special ideals can be written as \((\alpha)\) with a single generator, so \((\alpha) = \alpha R_\Delta\). Such an ideal \((\alpha)\) is called a principal ideal. Under the correspondence between ideals in \( R_\Delta \) and quadratic forms of discriminant \( \Delta \), principal ideals \((\alpha)\) with \( N(\alpha) > 0 \) correspond to forms equivalent to the principal form since the principal form \( ax^2 + bxy + cy^2 \) has \( a = 1 \) so the associated ideal \((a, b + \sqrt{\Delta})\) contains 1 and hence equals \( R_\Delta = (1) \), and then for a form \( Q \) equivalent to the principal form (hence properly equivalent to it since the principal form has mirror symmetry) the associated ideal \( L_Q \) will be \( \alpha R_\Delta = (\alpha) \) for some \( \alpha \) with \( N(\alpha) > 0 \). In the case of negative discriminant the condition \( N(\alpha) > 0 \) holds automatically, and a principal ideal is a lattice having the same shape as the full lattice \( R_\Delta \) since the lattice \( \alpha R_\Delta \) is obtained from \( R_\Delta \) by rotation and rescaling.

For positive discriminants a principal ideal \((\alpha)\) can have \( N(\alpha) < 0 \) and then the corresponding form might not be equivalent to the principal form. For example for \( \Delta = 12 \) with \( R_\Delta = \mathbb{Z}[\sqrt{3}] \) the form \( 3x^2 - y^2 \) is not equivalent to the principal form \( x^2 - 3y^2 \) but rather to its negative, and the ideal associated to \( 3x^2 - y^2 \) is \((3, \sqrt{3})\) which equals the principal ideal \((\sqrt{3})\) since \( 3 = \sqrt{3} \sqrt{3} \) lies in \((\sqrt{3})\). Here \( N(\sqrt{3}) = -3 < 0 \). By contrast, when \( \Delta = 8 \) with \( R_\Delta = \mathbb{Z}[\sqrt{2}] \) the principal ideal \((\sqrt{2})\) has \( N(\sqrt{2}) < 0 \) but the associated form \( 2x^2 - y^2 \) does happen to be equivalent to the principal form. This can be explained by the fact that the fundamental unit in \( \mathbb{Z}[\sqrt{2}] \) is \( 1 + \sqrt{2} \) which has norm \(-1\). In general, if we multiply an ideal \( L \) by a unit \( \varepsilon \) then \( \varepsilon L = L \), so multiplying the ideal \((\sqrt{2})\) by the unit \( 1 + \sqrt{2} \) gives the same ideal back again but now expressed as \((2 + \sqrt{2})\) with \( N(2 + \sqrt{2}) > 0 \). In the previous case of \( \mathbb{Z}[\sqrt{3}] \) the fundamental unit is \( 2 + \sqrt{3} \) of norm 1 so all units have norm 1. This means that we cannot express \((\sqrt{3})\) as \((\alpha)\) with \( N(\alpha) > 0 \) since it is a general fact about principal ideals that \((\alpha) = (\beta)\) exactly when \( \beta = \varepsilon \alpha \) for some unit \( \varepsilon \). (This is easy to show.)

When the class number for a given discriminant is 1 all forms are equivalent to the principal form, hence properly equivalent to it, so all ideals are strictly equivalent and can therefore be expressed as \( \alpha(1) = (\alpha) \) with \( N(\alpha) > 0 \) hence are principal ideals. For negative discriminants the converse is also true since the condition \( N(\alpha) > 0 \) is automatic, but for positive discriminants it can happen that all ideals are principal even though the class number is greater than 1. An example is the case \( \Delta = 12 \) that we just considered. Here the class number is 2 and every form is equivalent to either \( x^2 - 3y^2 \) or \( 3x^2 - y^2 \) so every ideal is either \( \alpha(1) \) or \( \alpha(\sqrt{3}) \) for some \( \alpha \) with \( N(\alpha) > 0 \) so in particular every ideal is principal.

Now we return to products of ideals. For ideals \( L = (\alpha_1, \alpha_2) \) and \( M = (\beta_1, \beta_2) \) the product \( LM \) is the ideal \((\alpha_1 \beta_1, \alpha_1 \beta_2, \alpha_2 \beta_1, \alpha_2 \beta_2)\) since each of the four products \( \alpha_i \beta_j \) is in \( LM \) and every element of \( LM \) is a sum of terms \( \alpha \beta \) for \( \alpha = \gamma_1 \alpha_1 + \gamma_2 \alpha_2 \) and \( \beta = \delta_1 \beta_1 + \delta_2 \beta_2 \), so \( \alpha \beta \) is a linear combination of the products \( \alpha_i \beta_j \) with coefficients
in \( R_\Delta \). Similarly the product of ideals \((\alpha_1, \cdots, \alpha_n)\) and \((\beta_1, \cdots, \beta_k)\) is the ideal generated by all the products \(\alpha_i\beta_j\).

As an example let us compute the product of the ideals \(L = (2, 1 + \sqrt{-5})\) and \(M = (2, 1 - \sqrt{-5})\) in \(\mathbb{Z}[\sqrt{-5}]\). The product \(LM = (2, 1 + \sqrt{-5})(2, 1 - \sqrt{-5})\) is then the ideal \((4, 2 + 2\sqrt{-5}, 2 - 2\sqrt{-5}, 6)\). In this ideal each generator is a multiple of 2 so we can pull out a factor of 2 to get \(LM = 2(2, 1 + \sqrt{-5}, 1 - \sqrt{-5}, 3)\). The ideal \((2, 1 + \sqrt{-5}, 1 - \sqrt{-5}, 3)\) contains 3 and 2 so it contains their difference 1. Once an ideal contains 1 it must be the whole ring, so \((2, 1 + \sqrt{-5}, 1 - \sqrt{-5}, 3) = (1) = \mathbb{Z}[\sqrt{-5}]\) hence \(LM = 2(1) = (2)\). The figure below shows these ideals as lattices, with \((2, 1 + \sqrt{-5})\) indicated by the heavy dots. This happens to be the same as \((2, 1 - \sqrt{-5})\), so \((2)\) is the square of the ideal \((2, 1 + \sqrt{-5})\). This is the sublattice indicated by the dots in squares.

This example illustrates the general fact that a product \(LM\) of two ideals \(L\) and \(M\) is always a sublattice of both \(L\) and \(M\) since each term of a typical element \(\sum_i \alpha_i\beta_i\) of \(LM\) lies in both \(L\) and \(M\) by the defining property of ideals.

This example also illustrates the fact that a product \(LM\) of two ideals need not consist merely of all products \(\alpha\beta\) of an element of \(L\) with an element of \(M\) since the number 2 belongs to \(LM\) but if we had \(2 = \alpha\beta\) then, computing norms, we would have \(4 = N(\alpha)N(\beta)\). There are no elements of \(\mathbb{Z}[\sqrt{-5}]\) of norm \(\pm 2\) since \(N(x + y\sqrt{-5}) = x^2 + 5y^2 = \pm 2\) has no integer solutions. Thus either \(\alpha\) or \(\beta\) would have norm \(\pm 1\) and hence be a unit \(\pm 1\) in \(\mathbb{Z}[\sqrt{-5}]\), but neither 1 nor \(-1\) is in \((2, 1 \pm \sqrt{-5})\), as one can see in the figure.

We have defined the norm of an ideal \(L\) in \(R_\Delta\) geometrically as the number of parallel translates of \(L\), including \(L\) itself, that are needed to fill up all of \(R_\Delta\), but for the ideals we will be most interested in, namely the stable ideals in Proposition 8.17, there is another interpretation of the norm \(N(L)\) that is more like the definition of the norm of an element \(\alpha\) as \(N(\alpha) = \alpha\overline{\alpha}\). This will be in terms of the product \(L\overline{L}\) where \(\overline{L}\) is the ideal consisting of the conjugates of all elements of \(L\).

**Proposition 8.20.** If the ideal \(L\) in \(R_\Delta\) is stable then \(L\overline{L} = (N(L))\), the principal ideal generated by the norm \(N(L)\).

*Proof:* By Proposition 8.13 the ideal \(L\) is equal to \(nL(a, \frac{b + \sqrt{\Delta}}{2})\) for some integer \(n \geq 1\) and some form \(ax^2 + bxy + cy^2\) of discriminant \(\Delta\) with \(a > 0\). It will suffice to prove
the Proposition in the case \( n = 1 \) since replacing an ideal \( L \) by \( nL \) does not affect the stabilizer and it multiplies \( N(L) \) by \( n^2 \), so both sides of the equation \( LL = (N(L)) \) are multiplied by \( n^2 \). Thus we may take \( L = L(a, \frac{b+\sqrt{\Delta}}{2}) \) for the rest of the proof. Since we assume \( L \) is stable, the form \( ax^2 + bx\gamma + c\gamma^2 \) is primitive by Proposition 8.17.

Let \( \tau = \frac{b+\sqrt{\Delta}}{2} \) so \( \tau \) is a root of the equation \( x^2 - bx + ac = 0 \). Then \( L = (a, \tau) \) and \( \overline{L} = (a, \overline{\tau}) \) so

\[
LL = (a^2, a\tau, a\overline{\tau}, \tau\overline{\tau}) = (a^2, a\tau, a\overline{\tau}, ac) = (a, \tau, \overline{\tau}, c)
\]

The ideal \((a, \tau, \overline{\tau}, c)\) contains the ideal \((a, \tau + \overline{\tau}, c) = (a, b, c)\). The latter ideal is all of \( R_\Delta \) since it contains all integral linear combinations \( ma + nb + qc \) and there is one such combination that equals 1 since the greatest common divisor of \( a, b, \) and \( c \) is 1 because the form \( ax^2 + bx\gamma + c\gamma^2 \) is primitive. (We know from Chapter 2 that the greatest common divisor \( d \) of \( a \) and \( b \) can be written as \( d = ma + nb \), and then the greatest common divisor of \( d \) and \( c \), which is the greatest common divisor of \( a, b, \) and \( c \), can be written as an integral linear combination of \( d \) and \( c \) and hence also of \( a, b, \) and \( c \).)

Thus the ideal \((a, \tau, \overline{\tau}, c)\) contains \( R_\Delta \) and so must equal it. Hence we have \( LL = aR_\Delta = (a) \) and this equals \((N(L)) \) since \( N(L) = a \) for \( L = L(a, \frac{b+\sqrt{\Delta}}{2}) \).

\[\Box\]

**Proposition 8.21.** An ideal \( L \) in \( R_\Delta \) is stable if and only if there exists an ideal \( M \) in \( R_\Delta \) such that \( LM \) is a principal ideal.

**Proof:** The forward implication follows from Proposition 8.20 by choosing \( M = \overline{L} \). For the opposite implication, suppose that \( LM = (\alpha) \), and let \( \beta \) be an element of \( \mathbb{Q}(\sqrt{\Delta}) \) such that \( \beta L \) is contained in \( L \). Then \( \beta(\alpha) = \beta LM \) is contained in \( LM = (\alpha) \). In particular this says that \( \beta \alpha \) is in \((\alpha) \) so \( \beta \alpha = \gamma \alpha \) for some element \( \gamma \) of \( R_\Delta \). Since \( \alpha \) is nonzero this implies \( \beta = \gamma \) and so \( \beta \) is an element of \( R_\Delta \). This shows that the stabilizer of \( L \) is \( R_\Delta \), so \( L \) is stable.

\[\Box\]

**Proposition 8.22.** If \( L \) and \( M \) are stable ideals in \( R_\Delta \) then \( N(LM) = N(L)N(M) \).

**Proof:** If \( L \) and \( M \) are stable then so is \( LM \) by Proposition 8.21 since the product of two principal ideals is principal. Thus \( LL = (N(L)), \overline{LM} = (N(M)) \), and \( LMLM = (N(LM)) \). Since \( LM\overline{LM} = LM\overline{LM} = LL\overline{LM} \) we therefore have \((N(LM)) = (N(L))(N(M)) = (N(L)N(M)) \). This implies \( N(LM) = N(L)N(M) \) since if \((a) = (b)\) for positive integers \( a \) and \( b \) then \( a = b \), as is evident from the lattices \((a) = L(a, a\tau) \) and \((b) = L(b, b\tau) \) for \( R_\Delta = \mathbb{Z}[\tau] \).

Interestingly enough, the formula \( LL = (N(L)) \) and the multiplicative property \( N(LM) = N(L)N(M) \) can fail to hold for ideals with stabilizer larger than \( R_\Delta \). A simple example is provided by taking \( L \) to be the ideal \((2, 1 + \sqrt{-3}) \) in \( \mathbb{Z}[\sqrt{-3}] \) which we considered earlier in this section, before Proposition 8.17, as an example of an ideal
corresponding to the nonprimitive form $2x^2 + 2xy + 2y^2$ of discriminant $-12$. Here $L = \mathcal{L}$ and the ideal $L^2 = \mathcal{L}\mathcal{L}$ is $(2, 1 + \sqrt{-3})(2, 1 - \sqrt{-3}) = (4, 2 + 2\sqrt{-3}, 2 - 2\sqrt{-3}, 4)$. Of these four generators we can obviously drop the repeated 4, and we can also omit the third generator which is expressible as the first generator minus the second. We are left with the ideal $(4, 2 + 2\sqrt{-3}) \mathcal{L} = (2, 1 + \sqrt{-3})(4, 2 + 2\sqrt{-3}, 2 - 2\sqrt{-3}, 4)$.

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Looking at the figure below, we see that $N(L) = 2$ and hence $N(2L) = 2^2 N(L) = 8$ so $N(L^2) \neq N(L)^2 = 4$. This shows that $N(LM)$ need not equal $N(L)N(M)$ in general.

Also we see from the figure that $L \mathcal{L} \neq (N(L))$ since $2L \neq (2)$. In fact $L \mathcal{L}$ is not even a principal ideal since $2L$ is a lattice of equilateral triangles while principal ideals have the same shape as the rectangular lattice $\mathbb{Z}[\sqrt{-3}]$.

Now at last we come to the construction of the ideal class group which we will denote $ICG(\Delta)$ until we show that it coincides with the class group $CG(\Delta)$ defined in Chapter 7 in terms of forms. Let $[L]$ denote the strict equivalence class of a stable ideal $L$ in $R_\Delta$ and let $ICG(\Delta)$ be the set of such classes $[L]$. The multiplication operation in $ICG(\Delta)$ is defined by taking products of ideals, so we set $[L][M] = [LM]$, recalling the fact that the product of two stable ideals is stable by Proposition 8.21.

To check that this product in $ICG(\Delta)$ is well defined we need to see that choosing different ideals $L'$ and $M'$ in the classes $[L]$ and $[M]$ does not affect $[LM]$. This is true because $[L] = [L']$ means $\alpha L = \alpha' L'$ for some $\alpha$ and $\alpha'$, and $[M] = [M']$ means $\beta M = \beta' M'$ for some $\beta$ and $\beta'$, hence $\alpha\beta LM = \alpha'\beta' L'M'$, so $[LM] = [LM']$. Here we are dealing with strict equivalence classes of ideals so we are assuming all of $\alpha, \beta, \alpha', \beta'$ have positive norms, hence so do $\alpha\beta$ and $\alpha'\beta'$. (As always this condition is automatic when $\Delta$ is negative.)

**Proposition 8.23.** $ICG(\Delta)$ is a commutative group with respect to the multiplication $[L][M] = [LM]$.

**Proof:** The commutativity property $[L][M] = [M][L]$ is easy since this amounts to saying $[LM] = [ML]$, which holds since multiplication of ideals is commutative, $LM = ML$, because multiplication in $R_\Delta$ is commutative.

To have a group there are three things to check. First, the multiplication should be associative, so $([L][M])[N] = [L]([M][N])$. By the definition of the product in $ICG(\Delta)$ this is equivalent to saying $[LM][N] = [L][MN]$ which in turn means $((LM)N) = [L(MN)]$, so it suffices to check that multiplication of ideals is associative, $(LM)N = L(MN)$. The claim is that each of these two products consists of all
the finite sums $\sum_i \alpha_i \beta_i y_i$ with $\alpha_i$, $\beta_i$, and $y_i$ elements of $L$, $M$, and $N$ respectively. Every such sum is in both $(LM)N$ and $L(MN)$ since each term $\alpha_i \beta_i y_i$ is in both of the ideals $(LM)N$ and $L(MN)$. Conversely, each element of $(LM)N$ is a sum of terms $(\sum_j \alpha_j \beta_j) y$ so it can be written as a sum $\sum_i \alpha_i \beta_i y_i$, and similarly each element of $L(MN)$ can be written as a sum $\sum_i \alpha_i \beta_i y_i$. Thus we have $(LM)N = L(MN)$.

Next, a group must have an identity element, and the class $[(1)]$ of the ideal $(1) = \mathbb{Z}_\Delta$ obviously serves this purpose since $(1)L = L$ for all ideals $L$, hence $[(1)][L] = [L]$. There is no need to check that $[L][(1)] = [L]$ as one would have to do for a noncommutative group since we have already observed that multiplication in $ICG(\Delta)$ is commutative.

The last thing to check is that each element of $ICG(\Delta)$ has a multiplicative inverse, and this is where we use the condition that we are considering only stable ideals in the definition of $ICG(\Delta)$. As we showed in Proposition 8.20, each stable ideal $L$ satisfies $LL = (n)$ where the integer $n$ is the norm of $L$. Then we have $[L][L] = [(n)] = [(1)]$ where this last equality holds since the ideals $(n)$ and $(1)$ are strictly equivalent, the norm of $n$ being $n^2$, a positive integer. Thus the multiplicative inverse of $[L]$ is $[\overline{L}]$. Again commutativity of the multiplication means that we do not have to check that $[\overline{L}]$ is an inverse for $[L]$ for multiplication both on the left and on the right.

There is a variant of the ideal class group in which the relation of strict equivalence of ideals is modified by deleting the word "strict", so an ideal $L$ is considered equivalent to $\alpha L$ for all nonzero elements $\alpha$ of $\mathbb{Z}_\Delta$ without the condition that $N(\alpha) > 0$. The preceding proof that $ICG(\Delta)$ is a group applies equally well in this setting by just omitting any mention of norms being positive. Sometimes the resulting group is called the class group while $ICG(\Delta)$ is called the strict class group or narrow class group. However, for studying quadratic forms the more appropriate notion is strict equivalence, which is why we are using this for the class group $ICG(\Delta)$.

Next we check that the one-to-one correspondence $\Phi : CG(\Delta) \to ICG(\Delta)$ induced by sending a form $Q = ax^2 + bxy + cy^2$ to the ideal $L_Q = (a, b+\sqrt{\Delta})$ respects the group structures defined on $CG(\Delta)$ and $ICG(\Delta)$. Given two classes $[Q_1]$ and $[Q_2]$ in $CG(\Delta)$ we can realize them by concordant forms $[a_1, b, a_2 c]$ and $[a_2, b, a_1 c]$ with $a_1$ and $a_2$ coprime. The product $[Q_1][Q_2]$ in $CG(\Delta)$ is then the class of $[a_1 a_2, b, c]$. The ideals corresponding to these three forms are $L_1 = (a_1, b+\sqrt{\Delta})$, $L_2 = (a_2, b+\sqrt{\Delta})$, and $L_3 = (a_1 a_2, b+\sqrt{\Delta})$. To show that multiplication in $CG(\Delta)$ corresponds under $\Phi$ to multiplication in $ICG(\Delta)$ it will suffice to show that $L_1 L_2 = L_3$. The product $L_1 L_2$ is the ideal $(a_1 a_2, a_1 b+\sqrt{\Delta}, a_2 b+\sqrt{\Delta}, b+\sqrt{\Delta} \cdot b+\sqrt{\Delta})$. This is certainly contained in $(a_1 a_2, b+\sqrt{\Delta})$ since the first generator $a_1 a_2$ is in $L_3$ and the other three generators are multiples of $b+\sqrt{\Delta}$ by elements of $\mathbb{Z}_\Delta$ hence are in $L_3$. On the other hand $L_3$ is contained in $L_1 L_2$ since $a_1 a_2$ is in $L_1 L_2$, and so is $b+\sqrt{\Delta}$ which can be written as a linear combination $ma\frac{b+\sqrt{\Delta}}{2} + na\frac{b+\sqrt{\Delta}}{2}$ for some integers $m$ and $n$, using the fact
that $a_1$ and $a_2$ are coprime so we have $ma_1 + na_2 = 1$ for some integers $m$ and $n$.

The identity element of $CG(\Delta)$ is the class of the principal form $[1,b,c]$ and this is sent by $\Phi$ to the class of the ideal $(1,\frac{b+\sqrt{\Delta}}{2}) = (1)$ which is the identity element of $ICG(\Delta)$. The inverse of an element of $CG(\Delta)$ determined by a form $[a,b,c]$ is the class of the mirror image form $[a,-b,c]$, so under $\Phi$ these correspond to the ideals $(a,\frac{b+\sqrt{\Delta}}{2})$ and $(a,\frac{-b+\sqrt{\Delta}}{2}) = (a,\frac{b-\sqrt{\Delta}}{2})$ and this last ideal is the conjugate of $(a,\frac{b+\sqrt{\Delta}}{2})$ so it gives the inverse of $(a,\frac{b+\sqrt{\Delta}}{2})$ in $ICG(\Delta)$.

Thus the group structures on $CG(\Delta)$ and $ICG(\Delta)$ are really the same, and we can use the notation $CG(\Delta)$ for both without any conflict.

To illustrate this let us consider $CG(\Delta)$ for $\Delta = -104$, so $R_\Delta = \mathbb{Z}[\sqrt{-26}]$. We looked at this example in Section 7.2 and found that $CG(\Delta)$ is a cyclic group of order 6 generated by the form $Q_4 = [5,4,6]$. From the topographs we could see that $Q_4^2$ was either $Q_3 = [3,2,9]$ or $Q_3^{-1} = [3,-2,9]$, but to determine which, we had to find a pair of concordant forms equivalent to $Q_4$ and multiply them together. Now we can use ideals to do the same calculation. The ideal corresponding to $Q_4 = [5,4,6]$ is $(5,2+\sqrt{-26})$ so for $Q_4^2$ the ideal is $(5,2+\sqrt{-26})(5,2+\sqrt{-26}) = (25,10+5\sqrt{-26},-22+4\sqrt{-26})$. The next step is to find a reduced basis for this ideal. As a lattice this ideal is generated by these three elements and their products with $\sqrt{-26}$. Thus we have the matrix

$$
\begin{pmatrix}
25 & 0 & 10 & -130 & -22 & -104 \\
0 & 25 & 5 & 10 & 4 & -22
\end{pmatrix}
$$

Using the three column operations this eventually reduces to $\begin{pmatrix}25 & 7 \\ 0 & 1\end{pmatrix}$ so the ideal is $(25,7+\sqrt{-26})$. The corresponding form is $[25,14,c]$ and we can determine $c$ from the discriminant equation $b^2 - 4ac = -104$ which gives $c = 3$. The form is thus $[25,14,3]$. A portion of the topograph of this form is shown at the right. There is a source vertex surrounded by the three values $3,9,10$ in counterclockwise order. The form $[3,-2,9]$ has exactly this same configuration at its source vertex, so we conclude that $Q_4^2 = Q_3^{-1}$, the same answer we got in Chapter 7.

5. **Unique Factorization of Ideals**

In this section we will be restricting our attention exclusively to discriminants $\Delta$ that are fundamental discriminants, so all forms will be primitive and hence all ideals in $R_\Delta$ will be stable. This means that we will be able to make free use of the formulas $N(LM) = N(L)N(M)$ and $LL = (N(L))$.

Our main goal in this section is to show that all ideals in $R_\Delta$, with the trivial exception of $R_\Delta$ itself, have unique factorizations as products of prime ideals, where an ideal $P$ different from $R_\Delta$ is called **prime** if whenever it is expressed as a product $LM$ of two ideals in $R_\Delta$, either $L$ or $M$ must equal $R_\Delta$, so the factorization becomes
the trivial factorization \( P = R_\Delta P \) that every ideal has. Note that \( R_\Delta \), considered as an ideal in itself, satisfies this condition but we do not call \( R_\Delta \) a prime ideal, just as the number 1 is not considered a prime number.

For an element \( \alpha \) of \( R_\Delta \) we know that \( \alpha \) is prime if its norm \( N(\alpha) \) is a prime in \( \mathbb{Z} \) (though the converse is not always true). The analog for ideals also holds: If the norm \( N(P) \) of an ideal \( P \) is a prime in \( \mathbb{Z} \) then \( P \) is a prime ideal. The argument is similar: Suppose \( P = LM \). Then \( N(P) = N(L)N(M) \). If we assume \( N(P) \) is prime in \( \mathbb{Z} \), then since both \( N(L) \) and \( N(M) \) are positive integers, one of them must equal 1. But the only ideal of norm 1 is \( R_\Delta \) since having norm 1 means that a single untranslated copy of the ideal covers all of \( R_\Delta \), making the ideal equal \( R_\Delta \). Thus we conclude that either \( L \) or \( M \) is \( R_\Delta \) and so \( P \) is a prime ideal.

Let us now work through an example to see how prime factorization of ideals can resolve the problem of nonunique prime factorization of elements. The example will be one we looked at earlier in the chapter, the factorizations

\[
6 = 2 \cdot 3 = (1 + \sqrt{-5}) \cdot (1 - \sqrt{-5})
\]

in \( R_\Delta = \mathbb{Z}[\sqrt{-5}] \) with \( \Delta = -20 \). We saw that 2, 3, and \( 1 \pm \sqrt{-5} \) were prime elements of \( \mathbb{Z}[\sqrt{-5}] \), and the two factorizations do not differ just by units since the only units in \( \mathbb{Z}[\sqrt{-5}] \) are \( \pm 1 \).

The transition from elements of \( R_\Delta \) to ideals in \( R_\Delta \) is achieved by associating to each nonzero element \( \alpha \) in \( R_\Delta \) the principal ideal \( (\alpha) \). Under this transition multiplication of elements corresponds to multiplication of ideals since \( (\alpha \beta) = (\alpha)(\beta) \). Another basic observation is that \( (\alpha) = (\beta) \) if and only if \( \alpha \) and \( \beta \) differ only by a unit. For if \( \beta = \varepsilon \alpha \) for some unit \( \varepsilon \) then \( (\varepsilon) \) contains \( \varepsilon \varepsilon^{-1} = 1 \) so \( (\varepsilon) = R_\Delta \) hence \( (\beta) = (\varepsilon \alpha) = (\varepsilon)(\alpha) = (\alpha) \). Conversely, if \( (\alpha) = (\beta) \) then since \( \beta \) is in \( (\alpha) \) we have \( \beta = \varepsilon \alpha \) for some \( \varepsilon \in R_\Delta \) and similarly we have \( \alpha = \eta \beta \) for some \( \eta \in R_\Delta \). Thus \( \alpha = \eta \beta = \eta \varepsilon \alpha \) hence \( \varepsilon \eta = 1 \) so \( \varepsilon \) and \( \eta \) are units, showing that \( \alpha \) and \( \beta \) differ just by a unit.

There is also a simple relation between the norm \( N((\alpha)) \) of the ideal \( (\alpha) \) and the norm \( N(\alpha) \) of the element \( \alpha \), namely \( N((\alpha)) = |N(\alpha)| \), where the absolute value is needed since norms of ideals are always positive. We can derive this formula from the formula \( LT = (N(L)) \), so in particular we have \( (\alpha)(\overline{\alpha}) = (N((\alpha))) \). The left side of this equation equals \( (\alpha \overline{\alpha}) \) which is \( (N(\alpha)) \). Thus the two principal ideals \( (N((\alpha))) \) and \( (N(\alpha)) \) are equal so the two integers \( N((\alpha)) \) and \( N(\alpha) \) differ only by a unit. If an integer times a unit in \( R_\Delta \) is an integer, that unit must also be an integer. The only integer units are \( \pm 1 \), so we conclude that \( N((\alpha)) = |N(\alpha)| \).

Returning to our example, we now have two different factorizations of the ideal \( (6) \) as a product of two ideals:

\[
(6) = (2)(3) = (1 + \sqrt{-5})(1 - \sqrt{-5})
\]
We would like to determine whether the ideals \(2\), \(3\), and \((1 \pm \sqrt{-5})\) are prime, and if they are not, factor them into prime ideals. One way the ideal \((2)\) might factor is as \(P\overline{P} = (N(P)) = (2)\) if there is an ideal \(P\) in \(\mathbb{Z}[\sqrt{-5}]\) of norm 2. Since 2 is prime, \(P\) would then be a prime ideal, as would \(\overline{P}\), so we would have a prime factorization \((2) = P\overline{P}\).

To find an ideal \(P\) of norm 2 consider a reduced basis \(a, b + c\sqrt{-5}\) for \(P\). The norm of the lattice \(L(a, b + c\sqrt{-5})\) is \(ac\), and we must have \(b < a\) so the only possibilities for the lattice are \(L(2, \sqrt{-5})\), \(L(2, 1 + \sqrt{-5})\), and \(L(1, 2\sqrt{-5})\). The last of these is ruled out since it contains 1 hence the ideal \(P\) would be the whole ring \(\mathbb{Z}[\sqrt{-5}]\). The first lattice \(L(2, \sqrt{-5})\) is not an ideal since it is not taken to itself by multiplication by \(\sqrt{-5}\) since \(-\sqrt{-5}\sqrt{-5} = -5\) is not in the lattice \(L(2, \sqrt{-5})\). The remaining lattice \(L(2, 1 + \sqrt{-5})\) however is an ideal since \(-\sqrt{-5} \cdot 2 = -2 + 2(1 + \sqrt{-5})\) and \(-\sqrt{-5} \cdot (1 + \sqrt{-5}) = -\sqrt{-5} - 5 = -3 \cdot 2 + (1 + \sqrt{-5})\).

Thus we have factored the ideal \((2)\) as \(P\overline{P}\) for the prime ideal \(P = (2, 1 + \sqrt{-5})\). In a similar way one can discover that \((3)\) factors as \(Q\overline{Q}\) for the ideal \(Q = (3, 1 + \sqrt{-5})\) of norm 3, so \(Q\) is also prime. (Later we will find a more efficient way of obtaining these factorizations.) As a result we have a prime factorization of the ideal \((6)\) as \((P\overline{P})(Q\overline{Q})\). These four factors could also be grouped as \((PQ)(\overline{PQ})\). We have \(PQ = (2, 1 + \sqrt{-5})(3, 1 + \sqrt{-5}) = (6, 2 + 2\sqrt{-5}, 3 + 3\sqrt{-5}, -4 + 2\sqrt{-5})\) which simplifies to \((6, 1 + \sqrt{-5})\). In fact 6 is in the ideal \((1 + \sqrt{-5})\) since \((1 + \sqrt{-5})(1 - \sqrt{-5}) = 6\) and we conclude that \(PQ = (1 + \sqrt{-5})\). Taking conjugates we also have \(\overline{PQ} = (1 - \sqrt{-5})\).

The conclusion of all this is that the two factorizations of \((6)\) as \((2)(3)\) and as \((1 + \sqrt{-5})(1 - \sqrt{-5})\) are simply two ways of grouping the four prime factors as \((P\overline{P})(Q\overline{Q})\) and as \((PQ)(\overline{PQ})\). One might wonder whether the grouping \((PQ)(\overline{PQ})\) would give a third factorization of \((6)\) but it happens that \(P = \overline{P}\) so this grouping gives nothing new. In other examples however it does give something new.

It will be helpful to have a criterion for when one ideal \(L\) in \(R_\Delta\) divides another ideal \(M\), meaning that \(M = LK\) for some ideal \(K\). For individual elements of \(R_\Delta\) it is easy to tell when one element divides another since \(\alpha\) divides \(\beta\) exactly when the quotient \(\beta/\alpha\) lies in \(R_\Delta\). For ideals, however, the criterion is rather different:

**Proposition 8.24.** An ideal \(L\) in \(R_\Delta\) divides an ideal \(M\) if and only if \(L\) contains \(M\).

One can remember this as ‘to divide is to contain’ (which sounds like some sort of competitive strategy). At first glance this result may seem a little puzzling since for ordinary numbers the divisors of a number \(n\), apart from \(n\) itself, are smaller than \(n\) while for ideals the divisors are larger, where ‘larger’ for sets means that one set contains the other.

The proposition gives some insight into the choice of the ideals \(P\) and \(Q\) in the preceding example where we factored the ideal \((6)\) in \(\mathbb{Z}[\sqrt{-5}]\) as \((P\overline{P})(Q\overline{Q})\) and as \((PQ)(\overline{PQ})\). Since we want \(P\overline{P} = (2)\) and \(PQ = (1 + \sqrt{-5})\), this means that \(P\) should
divide both \((2)\) and \((1 + \sqrt{-5})\). By the proposition this is the same as saying that \(P\) should contain both \((2)\) and \((1 + \sqrt{-5})\). An obvious ideal with this property is the ideal \((2, 1 + \sqrt{-5})\). Similarly one would be led to try \(Q = (3, 1 + \sqrt{-5})\). Then one could check that these choices for \(P\) and \(Q\) actually work.

Before proving the proposition let us derive a fact which will be used in the proof, a cancellation property of multiplication of ideals: If \(LM_1 = LM_2\) then \(M_1 = M_2\). To see this, first multiply the equation \(LM_1 = LM_2\) by \(\mathcal{L}\) to get \(\mathcal{L}LM_1 = \mathcal{L}LM_2\). Since \(\mathcal{L}L = (n)\) for \(n = N(L)\), a positive integer, we then have \((n)M_1 = (n)M_2\). The ideal \((n)M_1\) is just a rescaling of \(M_1\) by a factor of \(n\) since if we write \(M_1 = (\alpha, \beta)\) then \((n)M_1 = (n\alpha, n\beta)\). Similarly \((n)M_2\) is a rescaling of \(M_2\) by the same factor \(n\). Since \((n)M_1 = (n)M_2\) these rescalings are equal, so after rescaling again by the factor \(1/n\) we conclude that \(M_1 = M_2\).

Now let us prove the proposition.

**Proof:** Suppose first that \(L\) divides \(M\), so \(M = LK\) for some ideal \(K\). A typical element of \(LK\) is a sum \(\sum_i \alpha_i \beta_i\) with \(\alpha_i \in L\) and \(\beta_i \in K\). Since \(L\) is an ideal, each term \(\alpha_i \beta_i\) is then in \(L\) and hence so is their sum. This shows that \(L\) contains \(LK = M\).

For the converse, suppose \(L\) contains \(M\). Then \(L\mathcal{L}\) contains \(M\mathcal{L}\). Since \(L\mathcal{L} = (n)\) for \(n = N(L)\) this says that \((n)\) contains \(M\mathcal{L}\), so every element of \(M\mathcal{L}\) is a multiple of \(n\) by some element of \(R_\Delta\). This means that if we write \(M\mathcal{L} = (\alpha, \beta)\) then we can define an ideal \(K\) by letting \(K = (\alpha/n, \beta/n)\).

Now we have \((n)K = (n)(\alpha/n, \beta/n) = (\alpha, \beta) = M\mathcal{L}\). Multiplying by \(L\) we then have \((n)KL = M\mathcal{L}L = M(\mathcal{N})\). Canceling the factor \((n)\) gives the equation \(KL = M\), which says that \(L\) divides \(M\), finishing the proof of the converse. \(\square\)

When we proved unique prime factorization for \(\mathbb{Z}\) and those rings \(R_\Delta\) which have a Euclidean algorithm, a key step was showing that if a prime \(p\) divides a product \(ab\) then \(p\) must divide either \(a\) or \(b\). Now we prove the corresponding fact for ideals:

**Lemma 8.25.** If a prime ideal \(P\) divides a product \(LM\) of two ideals, then \(P\) must divide either \(L\) or \(M\).

**Proof:** An equivalent statement is that if \(P\) divides \(LM\) but not \(L\), then \(P\) divides \(M\), and this is what we will prove. To do this, consider the set \(P + L\) of all sums \(\alpha + \beta\) of elements \(\alpha \in P\) and \(\beta \in L\). This set \(P + L\) is an ideal since if \(P = (\alpha_1, \alpha_2)\) and \(L = (\beta_1, \beta_2)\) then \(P + L = (\alpha_1, \alpha_2, \beta_1, \beta_2)\). The ideal \(P + L\) is strictly larger than \(P\) since our assumption that \(P\) does not divide \(L\) means that \(P\) does not contain \(L\), so any element of \(L\) not in \(P\) is in \(P + L\) but not \(P\). Thus \(P + L\) contains \(P\), hence divides \(P\), but is not equal to \(P\) so since \(P\) is prime we must have \(P + L = R_\Delta\).

In particular \(P + L\) contains \(1\) so we can write \(1 = \alpha + \beta\) for some \(\alpha \in P\) and \(\beta \in L\). For an arbitrary element \(y \in M\) we then have \(y = \alpha y + \beta y\). The term \(\alpha y\) is in \(P\) since \(\alpha\) is in \(P\) and \(P\) is an ideal. The term \(\beta y\) is in \(LM\) since \(\beta\) is in \(L\) and \(y\)
is in $M$. We assume $P$ divides $LM$ so $P$ contains $LM$ and it follows that $\beta y$ is in $P$. Thus both terms on the right side of the equation $y = \alpha y + \beta y$ are in $P$ so $y$ is in $P$. Since $y$ was an arbitrary element of $M$ this shows that $M$ is contained in $P$, or in other words $P$ divides $M$, which is what we wanted to prove.

Now we can prove our main result:

**Theorem 8.26.** Every ideal in $R_\Delta$ other than $R_\Delta$ itself factors as a product of prime ideals, and this factorization is unique up to the order of the factors.

**Proof:** We first show the existence of a prime factorization for each ideal $L \neq R_\Delta$. If $L$ is prime itself there is nothing to prove, so suppose $L$ is not prime, hence there is a factorization $L = KM$ with neither factor equal to $R_\Delta$. Taking norms, we have $N(L) = N(K)N(M)$. Both $N(K)$ and $N(M)$ are greater than 1 since $R_\Delta$ is the only ideal of norm 1. Hence $N(K) < N(L)$ and $N(M) < N(L)$. By induction on the norm, both $K$ and $M$ have prime factorizations, hence so does $L = KM$. We can start the induction with the case $N(L) = 2$, a prime, hence $L$ is prime. (The case $N(L) = 1$ does not arise since $L \neq R_\Delta$.)

For the uniqueness, suppose an ideal $L$ has prime factorizations $P_1 \cdots P_k$ and $Q_1 \cdots Q_l$. We can assume $k \leq l$ by a notational change if necessary. The prime ideal $P_1$ divides the product $Q_1(Q_2 \cdots Q_l)$ so by the preceding lemma it must divide either $Q_1$ or $Q_2 \cdots Q_l$. In the latter case the same reasoning shows it must divide either $Q_2$ or $Q_3 \cdots Q_l$. Repeating this argument enough times we eventually deduce that $P_1$ must divide some $Q_i$, and after permuting the factors of $Q_1 \cdots Q_l$ we can assume that $P_1$ divides $Q_1$. When one prime ideal divides another prime ideal they must be equal. (Proof: If $P$ divides $Q$ then $Q = PM$ for some $M$, but $Q$ being prime implies either $P = R_\Delta$, which is impossible if $P$ is prime, or $M = R_\Delta$, hence $P = Q$ as claimed.)

Once we have $P_1 = Q_1$ we can cancel this common factor of $P_1 \cdots P_k$ and $Q_1 \cdots Q_l$ to get $P_2 \cdots P_k = Q_2 \cdots Q_l$. Repeating this process often enough we eventually get, after suitably permuting the $Q_i$’s, that $P_1 = Q_1, P_2 = Q_2, \ldots, P_{k-1} = Q_{k-1},$ and $P_k = Q_k \cdots Q_l$. Since $P_k$ is prime, as are the $Q_i$’s, the equation $P_k = Q_k \cdots Q_l$ can have only one term on the right side, so $k = l$ and $P_k = Q_k$. This finishes the proof of the uniqueness of prime factorizations of ideals.

**Corollary 8.27.** If the class number for the fundamental discriminant $\Delta$ is 1, then all elements of $R_\Delta$ other than units and 0 have unique factorizations as products of prime elements, where the uniqueness is up to order and multiplication by units.

**Proof:** If the class number is 1 then all ideals are principal. The result then follows immediately from the theorem and the fact that principal ideals in $R_\Delta$ correspond exactly to nonzero elements of $R_\Delta$ up to multiplication by units, as described earlier in this section.
Proposition 8.28. If $P$ is a prime ideal in $R_\Delta$ then $P$ contains a unique prime positive integer $p$, and either $P = (p)$ with $N(P) = p^2$ or $P\overline{P} = (p)$ with $N(P) = p$.

Proof: We have $P\overline{P} = (N(L))$. Factor the positive integer $N(P)$ as a product $p_1 \cdots p_k$ of primes $p_i$, so $P\overline{P} = (p_1) \cdots (p_k)$. Thus $P$ divides $(p_1) \cdots (p_k)$ and since $P$ is prime it must divide one of the factors, say $(p_i)$. This means $P$ contains $(p_i)$, so in particular $P$ contains $p_i$. Let $p_i = p$. If $P$ contained another prime $q$ it would have to contain 1 since $p$ and $q$ are coprime and hence $mp + nq = 1$ for some integers $m$ and $n$. Thus we would have $P = R_\Delta$ which is impossible since $P$ is a prime ideal. This shows that $p$ is the only prime positive integer in $P$.

Since $P$ divides $(p)$ we have $PQ = (p)$ for some ideal $Q$. Then $N(P)N(Q) = N((p)) = p^2$, which implies that either $N(P) = p$ or $N(P) = p^2$. In the first case we would have $P\overline{P} = (N(P)) = (p)$ and in the second case we would have $N(Q) = 1$ and hence $Q = R_\Delta$ so $P = (p)$.

Our next proposition will give more information on how an ideal $(p)$ factors into prime ideals in $R_\Delta$ when $p$ is an odd prime. The result will be stated in terms of the integer $d$ which is defined to equal $\Delta$ when $\Delta \equiv 1 \mod 4$ and $\Delta/4$ when $\Delta \equiv 0 \mod 4$. Since we are assuming that $\Delta$ is a fundamental discriminant it follows that $d$ is a product of distinct primes. (Otherwise, if $d$ was divisible by 4 then it would be a discriminant and $\Delta = 4d$ would not be fundamental, while if $d$ was divisible by $q^2$ for some odd prime $q$ then we would have $q^2 \equiv 1 \mod 4$ and hence $\Delta \equiv \Delta/q^2 \mod 4$ so $\Delta/q^2$ would be a discriminant, making $\Delta$ again nonfundamental.)

Proposition 8.29. Let $p$ be an odd prime. Then:
(a) If $p$ does not divide $d$ and the Legendre symbol $\left(\frac{d}{p}\right) = +1$ then $(p) = P\overline{P}$ with $P$ a prime ideal in $R_\Delta$ and $P \neq \overline{P}$. Specifically, $P = (p, b + \sqrt{d})$ and $\overline{P} = (p, b - \sqrt{d})$ or vice versa, where $b^2 = d \mod p$.
(b) If $p$ divides $d$ then $(p) = P^2$ with $P$ prime. Specifically, $P = (p, \sqrt{d})$.
(c) If $p$ does not divide $d$ and $\left(\frac{d}{p}\right) = -1$ then $(p)$ is a prime ideal in $R_\Delta$.

Proof: For part (a), since we assume $\left(\frac{d}{p}\right) = +1$ there exists an integer $b$ with $b^2 \equiv d \mod p$. Here $p$ does not divide $b$ since we assume it does not divide $d$. Let $P = (p, b + \sqrt{d})$. Then $P\overline{P} = (p^2, pb + p\sqrt{d}, pb - p\sqrt{d}, b^2 - d) = p(p, b + \sqrt{d}, b - \sqrt{d}, \frac{b^2 - d}{p})$, where the fraction $\frac{b^2 - d}{p}$ is an integer since $b^2 \equiv d \mod p$. The ideal $(p, b + \sqrt{d}, b - \sqrt{d}, \frac{b^2 - d}{p})$ contains the numbers $p$ and $2b$ which are coprime since $p$ does not divide $b$ and $p \neq 2$. Hence this ideal contains 1 and so equals $R_\Delta$. Thus we have $P\overline{P} = (p)$. In particular this says that $N(P) = p$ so $P$ is prime (as is $\overline{P}$).

If $P = \overline{P}$ then $P$ would contain $b + \sqrt{d}$ and $b - \sqrt{d}$ so it would contain $2b$. It also contains $p$ so by the reasoning in the preceding paragraph we would have $P = R_\Delta$, which is impossible for a prime ideal. Thus $P \neq \overline{P}$, which finishes part (a).

For (b), if $p$ divides $d$ we let $P = (p, \sqrt{d})$, so $\overline{P} = (p, -\sqrt{d}) = P$. We have $P^2 =
\[(p, \sqrt{d})^2 = (p^2, p\sqrt{d}, d) = p(p, \sqrt{d}, \frac{d}{p}).\]

The ideal \((p, \sqrt{d}, \frac{d}{p})\) contains the integers \(p\) and \(\frac{d}{p}\) which are coprime since \(d\) is a product of distinct primes, hence this ideal is \(R_\Delta\) and so \(P^2 = (p)\). Notice that this proof of (b) also works when \(p = 2\).

For (c) we will prove the logically equivalent contrapositive statement that if \((p) = P\bar{P}\) then either \(p\) divides \(d\) or \(\left(\frac{d}{p}\right) = +1\). Let \(a, b + c\tau\) be a reduced basis for \(P\). We have \(N(P) = ac = p\) so \(a\) is either 1 or \(p\), but if \(a = 1\) we would have \(P = R_\Delta\) so we must have \(a = p\) and hence \(c = 1\). Thus the second reduced basis element is \(\beta = b + \tau\). We have \(N(\beta) = \beta\bar{\beta}\) and this lies in \(P\) since \(\beta\) does. All the integers in \(P = L(p, b + \tau)\) are multiples of \(p\) so \(p\) divides \(N(\beta)\).

We now split the argument into two cases according to whether \(\Delta\) is 0 or 1 mod 4. If \(\Delta \equiv 0\) mod 4 then \(d = \Delta/4\) and \(\tau = \sqrt{d}\). Since \(p\) divides \(N(\beta) = (b + \sqrt{d}) \cdot (b - \sqrt{d}) = b^2 - d\) we must have \(b^2 \equiv d\) mod \(p\). Thus either \(p\) divides \(d\) or \(\left(\frac{d}{p}\right) = +1\), which is what we wanted to prove. In the opposite case \(\Delta \equiv 1\) mod 4 we have \(\Delta = d\) and \(\tau = \frac{1 + \sqrt{d}}{2}\). Now \(p\) divides \(N(\beta) = (b + \tau) \cdot (b + \bar{\tau}) = b^2 + b(\tau + \bar{\tau}) + \tau\bar{\tau} = b^2 + b + \frac{1 - d}{4}\). Multiplying this last expression by 4 we see that \(p\) also divides \(4b^2 + 4b + 1 - d\), hence \((2b + 1)^2 \equiv d\) mod \(p\). Thus \(d\) is a square mod \(p\) so again either \(p\) divides \(d\) or \(\left(\frac{d}{p}\right) = +1\).

There is an alternative argument for part (c) using quadratic forms. If \((p) = P\bar{P}\) then \(p\) is in \(P\) and the value that the quadratic form \(Q_p\) associated to \(P\) takes on this element of \(P\) is \(Q_p(p) = N(p)/N(P) = p^2/p = p\). Thus \(p\) is representable by a form of discriminant \(\Delta\) so either \(p\) divides \(\Delta\) or \(\left(\frac{\Delta}{p}\right) = +1\), by Proposition 6.5 and Lemma 6.7. Since \(\Delta\) is either \(d\) or \(4d\) and \(p\) is odd, this is equivalent to saying that either \(p\) divides \(d\) or \(\left(\frac{d}{p}\right) = +1\).

In the analog of the preceding proposition for the prime \(p = 2\) it turns out that the answer depends on the value of \(d\) mod 8. Since \(d\) is not divisible by 4, its value mod 8 cannot be 0 or 4. The remaining six possibilities are enumerated below:

**Proposition 8.30.** The factorization of \((2)\) into prime ideals in \(R_\Delta\) is given by:

(a) \((2) = P\bar{P}\) for \((2, \frac{1 + \sqrt{d}}{2})\) if \(d \equiv 1\) mod 8, and in this case \(P \neq \bar{P}\).

(b) \((2) = P^2\) for \((2, \sqrt{d})\) if \(d \equiv 2\) or 6 mod 8.

(c) \((2) = P^2\) for \((2, 1 + \sqrt{d})\) if \(d \equiv 3\) or 7 mod 8.

(d) \((2)\) is prime in \(R_\Delta\) if \(d \equiv 5\) mod 8.

Notice that if \(\Delta\) is odd then \(\Delta = d\) and we are in the cases (a) or (d) above, while if \(\Delta\) is even then \(\Delta = 4d\) and we are in the cases (b) or (c) depending on the parity of \(d\). Thus when \((2)\) factors in \(R_\Delta\), the two factors are the same or different depending on whether 2 divides \(\Delta\) or not. This is exactly the same way that \((p)\) factors when \(p\) is an odd prime, as described in Proposition 8.29.

**Proof:** For (a), note first that \(\Delta\) is odd so \(\frac{1 + \sqrt{d}}{2}\) is an element of \(R_\Delta\). We have \(P\bar{P} = (2, \frac{1 + \sqrt{d}}{2})(2, \frac{1 - \sqrt{d}}{2}) = (4, 2\frac{1 + \sqrt{d}}{2}, 2\frac{1 - \sqrt{d}}{2}, \frac{1 - d}{4}) = 2(2, \frac{1 + \sqrt{d}}{2}, \frac{1 - \sqrt{d}}{2}, \frac{1 - d}{8})\). Here \(\frac{1 - d}{8}\) is an
integer since we assume $d \equiv 1 \pmod{8}$. The ideal $(2, \frac{1+\sqrt{d}}{2}, \frac{1-\sqrt{d}}{2}, \frac{1-d}{8})$ contains $\frac{1+\sqrt{d}}{2} + \frac{1-\sqrt{d}}{2} = 1$ so it is $R_{\Delta}$, hence $P\overline{P} = (2)$. If $P = \overline{P}$ then $P$ would contain both $\frac{1+\sqrt{d}}{2}$ and $\frac{1-\sqrt{d}}{2}$ so it would contain their sum 1 and we would have $P = R_{\Delta}$ which is impossible, so $P \neq \overline{P}$.

In case (b) the number $d$ is even and the proof of part (b) of Proposition 8.29 applies, as noted there.

For (c) we have $d \equiv 3 \pmod{4}$ so we can write $d = 4k + 3$ for some integer $k$. Then $P^2 = (2, 1 + \sqrt{d})^2 = (4, 2 + 2\sqrt{d}, 1 + d + 2\sqrt{d}) = (4, 2 + 2\sqrt{d}, 4k + 4 + 2\sqrt{d}) = 2(2, 1 + \sqrt{d}, 2k + 2 + \sqrt{d})$. The ideal $(2, 1 + \sqrt{d}, 2k + 2 + \sqrt{d})$ contains 2 so it contains all even integers. In particular it contains $2k + 2$ so we can subtract this from the last generator $2k + 2 + \sqrt{d}$ to see that this ideal contains $\sqrt{d}$. Then from the second generator $1 + \sqrt{d}$ we see that the ideal contains 1, making it all of $R_{\Delta}$, and we conclude that $P^2 = (2)$.

For (d) we argue by contradiction, so suppose that (2) is not prime. Then $(2) = P\overline{P}$ for some ideal $P$ with $N(P) = 2$. Since $P$ divides (2) it contains (2) and so it contains 2. On this element 2 in $P$ the associated quadratic form $Q_P$ has the value $N(2)/N(P) = 2^2/2 = 2$. Thus there is a form of discriminant $\Delta$ that represents 2. From Section 6.2 we know that this implies $\Delta \equiv 5 \pmod{8}$. But $\Delta = d$ in part (d) so we have reached the desired contradiction.

Let us conclude this section with some comments on what happens when the discriminant $\Delta$ is not a fundamental discriminant. One might hope that the unique factorization property for ideals still holds at least for stable ideals, the ideals corresponding to primitive forms. However, this is not the case, and here is an example. Take $\Delta = -12$, so $R_{\Delta} = \mathbb{Z}[\sqrt{-3}]$. The class number is 1 in this case so all stable ideals are principal (and recall that principal ideals are always stable). Consider the factorizations $(4) = (2)(2) = (1 + \sqrt{-3})(1 - \sqrt{-3})$. The ideals (2) and $(1 \pm \sqrt{-3})$ are prime since their norms are 4 so any nontrivial factorization as $(\alpha)(\beta)$ would have $N(\alpha) = N(\beta) = 2$ but no elements of $\mathbb{Z}[\sqrt{-3}]$ have norm 2 since $x^2 + 3y^2 = 2$ has no integer solutions. The three ideals (2) and $(1 \pm \sqrt{-3})$ are distinct since the only units in $\mathbb{Z}[\sqrt{-3}]$ are $\pm 1$. Thus we have two different factorizations of (4) into prime ideals when we restrict attention just to stable ideals. If one drops this restriction then unique prime factorization still fails since for the ideal $L = (2, 1 + \sqrt{-3})$ we saw in the discussion following Corollary 8.23 that $L^2 = 2L$, but unique factorization implies the cancellation property so we would then have $L = (2)$, which is false.

One might ask where the proof of unique factorization breaks down for stable ideals in the case of a nonfundamental discriminant. The answer is in the key property in Lemma 8.25 that if a prime ideal $P$ divides a product $LM$ then it must divide one of the factors $L$ or $M$. In the proof of this we considered the ideal $P+L$, but unfortunately this need not be a stable ideal when $P$ and $L$ are stable. For example, in the preceding
paragraph if we take \( P = (2) \), \( L = (1 + \sqrt{-3}) \), and \( M = (1 - \sqrt{-3}) \) then \( P + L \) is the ideal \((2, 1 + \sqrt{-3})\), but this is not stable as we saw after Corollary 8.23. And in fact the ideal \((2)\) does not divide either \((1 + \sqrt{-3})\) or \((1 - \sqrt{-3})\).

6. Applications to Forms

As we have seen, ideals provide an alternative way of constructing the class group \( CG(\Delta) \). One of the main uses of the group structure in \( CG(\Delta) \) in Chapter 7 was in Theorem 7.8 which characterized the primitive forms of discriminant \( \Delta \) representing a given number \( n \) in terms of the forms representing the prime factors of \( n \). When \( \Delta \) is a fundamental discriminant the same characterization can be derived from the unique factorization property of ideals in \( R_\Delta \). This viewpoint provides additional insights into the somewhat subtle answer to the representation problem. Here is a restatement of the result we will now prove using ideals:

**Theorem 8.31.** Let \( \Delta \) be a fundamental discriminant and let \( n > 1 \) be a number represented by at least one form of discriminant \( \Delta \). If the prime factorization of \( n \) is \( n = p_1^{e_1} \cdots p_k^{e_k} \) for distinct primes \( p_i \), with \( e_i = 1 \) for each \( p_i \) dividing \( \Delta \) and \( e_i \geq 1 \) otherwise, then the forms of discriminant \( \Delta \) representing \( n \) are exactly the forms \( Q_1^{e_1} \cdots Q_k^{e_k} \) where \( Q_i \) represents \( p_i \) and the product \( Q_1^{e_1} \cdots Q_k^{e_k} \) is formed in the class group \( CG(\Delta) \).

There are a couple things that come up in the proof that we will explain in advance to avoid complicating the later arguments. The first is the elementary fact that an element \( \alpha \) in \( R_\Delta \) belongs to an ideal \( L \) if and only if the ideal \((\alpha)\) factors as \( (\alpha) = LM \) for some ideal \( M \). This is because \( \alpha \) is an element of \( L \) exactly when the ideal \((\alpha)\) is contained in \( L \), or in other words, \( L \) divides \((\alpha)\), which means \((\alpha) = LM \) for some ideal \( M \).

Next is a reformulation of what it means for a form \( Q_L \) to represent a number \( n \). By definition, \( Q_L(\alpha) = N(\alpha)/N(L) \) for \( \alpha \) in \( L \). Thus if we choose a basis \( \alpha_1, \alpha_2 \) for \( L \) regarded as a lattice and we let \( \alpha = x\alpha_1 + y\alpha_2 \) for integers \( x \) and \( y \), then \( Q_L(x, y) = N(x\alpha_1 + y\alpha_2)/N(L) \). For this to give a representation of \( n \) means that \( x \) and \( y \) are coprime. In terms of \( \alpha \) this is saying that \( \alpha \) is not a multiple \( m\beta \) of any element \( \beta \) of \( L \) with \( m > 1 \). This last condition can be abbreviated to saying just that \( \alpha \) is primitive in \( L \).

We can also define what it means for an ideal \( L \) to be primitive, namely, \( L \) is not divisible by any principal ideal \((m)\) with \( m \) an integer greater than 1, or in other words \( L \) is not a multiple \( mL' \) of any other ideal \( L' \) with \( m > 1 \). We could require \( m \) to be a prime without affecting the definition since if \( L = mL' \) with \( m = pq \) for \( p \) a prime then \( L \) is \( p \) times the ideal \( qL' \). An ideal \( L_Q = L(a, \frac{b+\sqrt{\Delta}}{2}) \) associated to a form \( Q = ax^2 + bxy + cy^2 \) of discriminant \( \Delta \) is always primitive since the element \( \frac{b+\sqrt{\Delta}}{2} \) lies on the first row of \( R_\Delta \) above the \( x \)-axis. By Proposition 8.13 every ideal in
\( R_\Delta \) is equal to \( nL_Q \) for some integer \( n \geq 1 \) and some form \( Q \) of discriminant \( \Delta \), so the primitive ideals are exactly the ideals \( L_Q \).

An equivalent way of formulating the condition for \( L \) to be primitive is to say that the factorization of \( L \) into prime ideals \( L = P_1 \cdots P_k \) satisfies the following two conditions:

(i) No \( P_i \) is a prime ideal \( (p) \) with \( p \) a prime integer. Thus each \( P_i \) has norm a prime rather than the square of a prime.

(ii) There is no pair of factors \( P_i \) and \( P_j \) with \( i \neq j \) such that \( P_i = \overline{P}_j \). In particular if \( P_i = \overline{P}_i \) then \( P_i \) can occur only once in the prime factorization of \( L \).

**Proof of Theorem 8.31:** Suppose that a positive number \( n \) is represented by a form \( Q \). From the correspondence between proper equivalence classes of forms and strict equivalence classes of ideals we may assume \( Q = Q_L \) for some ideal \( L \). Thus \( n = Q_L(\alpha) = N(\alpha)/N(L) \) for some primitive \( \alpha \) in \( L \). Since \( n \) and \( N(L) \) are positive, so is \( N(\alpha) \).

We can reduce to the case that \( \alpha \) is a positive integer by the following argument. We have \( n = N(\alpha)/N(L) = N(\overline{\alpha}\alpha)/N(\overline{\alpha}L) = Q_{\overline{\alpha}L}(\overline{\alpha}\alpha) \). The element \( \overline{\alpha}\alpha \) of \( \overline{\alpha}L \) is primitive in \( \overline{\alpha}L \) since if \( \overline{\alpha}\alpha = q\overline{\alpha}\beta \) for some positive integer \( q \) and some \( \beta \) in \( L \), then \( \alpha = q\beta \) which forces \( q \) to be 1 since \( \alpha \) is primitive in \( L \). The integer \( m = \overline{\alpha}\alpha \) is \( N(\alpha) \) which is positive as noted above. Also, \( m \) is in \( \overline{\alpha}L \) since \( \alpha \) is in \( L \). The ideals \( L \) and \( \overline{\alpha}L \) are strictly equivalent since \( N(\overline{\alpha}) = N(\alpha) > 0 \), so the forms \( Q_{\overline{\alpha}L} \) and \( Q_L \) are properly equivalent. This shows that we may take \( n \) to be represented as \( n = Q_L(m) \) for some primitive positive integer \( m \) in the new \( L \).

Next we reduce to the case that \( L \) is a primitive ideal. If \( L \) is not primitive we can write it as \( L = qL' \) for some integer \( q > 1 \) with \( L' \) primitive. Since \( m \) is in \( L = qL' \) we have \( m = qr \) for some \( r \) in \( L' \), and in fact \( r \) must be an integer since \( r = m/q \) and the only rational numbers in \( R_\Delta \) are integers. Since \( m \) and \( q \) are positive, so is \( r \). Also, \( r \) is primitive in \( L' \) since \( m \) is primitive in \( L \) and we are just rescaling \( m \) and \( L \) by a factor of \( 1/q \) to get \( r \) and \( L' \). The equation \( n = N(m)/N(L) \) can be written as \( n = N(qr)/N(qL') = N(r)/N(L') \) since \( qL' = (q)L' \) and \( N((q)) = N(q) \). This shows that \( n \) is represented as \( Q_{L'}(r) = n \). The form \( Q_{L'} \) is properly equivalent to \( Q_L \) since \( L = qL' \) and \( N(q) > 0 \). The net result of this argument is that we can assume that \( n \) is represented as \( n = Q_L(m) = N(m)/N(L) \) where \( L \) is primitive and \( m \) is a positive integer that is a primitive element of \( L \).

Since \( m \) is in \( L \) we have \( (m) = LM \) for some ideal \( M \). This \( M \) must also be primitive, otherwise if \( M = qM' \) for some ideal \( M' \) and some integer \( q > 1 \), then, arguing as in the preceding paragraph, we would have \( m = qr \) for some positive integer \( r \) in \( M' \) with \( (r) = LM' \). This last equality implies that \( r \) is in \( L \), so \( m \) would not be primitive in \( L \).

Since \( L \) and \( M \) are both primitive, their factorizations into prime ideals satisfy the
earlier conditions (i) and (ii). Then since their product is \((m)\) with \(m\) an integer, we must have \(M = \overline{L}\). Thus \((m) = L\overline{L}\) and so \(m = N(L)\). Now we have \(n = N(m)/N(L) = m^2/m = m\) so \(n = m\) and the representation of \(n\) becomes \(n = Q_L(n)\) with \(L\) primitive and \(n = N(L)\).

Let the factorization of \(L\) into prime ideals be \(L = P_1 \cdots P_k\). Then \(N(P_i)\) is a prime \(p_i\) and \(P_i\) is in \(P_i\) since \(P_i\overline{P}_i = (p_i)\). Also, \(p_i\) is primitive in \(P_i\) since \(p_i\) is prime so if \(p_i\) was not primitive in \(P_i\) then \(P_i\) would contain 1 which is impossible since \(P_i \neq R\). If we denote \(Q_{P_i}\) by \(Q_i\) for simplicity then \(Q_i\) represents \(p_i\) since \(Q_i(p_i) = N(p_i)/N(P_i) = p_i^2/p_i = p_i\).

Since \(n = N(L)\) and \(L = P_1 \cdots P_k\) we have \(n = p_1 \cdots p_k\). The prime factorization \(n = p_1 \cdots p_k\) is unique so the prime ideals \(P_i\) are uniquely determined by \(n\) up to the ambiguity of replacing \(P_i\) by \(\overline{P}_i\). In \(CG(\Delta)\) this amounts to replacing \(Q_i\) by \(Q_i^{-1}\). Keeping in mind the condition (ii), we have now shown that if a form \(Q\) represents \(n\) then in \(CG(\Delta)\) we have \(Q = Q_1^{\pm e_1} \cdots Q_k^{\pm e_k}\) where \(n = p_1^{e_1} \cdots p_k^{e_k}\) is the factorization of \(n\) into powers of distinct primes \(p_i\) and the form \(Q_i\) represents \(p_i\). The condition (ii) implies that \(e_i = 1\) for each \(i\) with \(P_i = \overline{P}_i\), that is, for each \(p_i\) that divides the discriminant \(\Delta\).

Conversely, suppose \(n = p_1^{e_1} \cdots p_k^{e_k}\) is the factorization of \(n\) into powers of distinct primes \(p_i\) with \(e_i = 1\) when \(p_i\) divides \(\Delta\), and suppose the form \(Q_i\) represents \(p_i\). We want to show that \(Q_1^{\pm e_1} \cdots Q_k^{\pm e_k}\) represents \(n\). By the arguments in the first part of the proof applied to \(P_i\) in place of \(n\) there is an ideal \(L_i\) containing \(P_i\) with \(N(L_i) = p_i\), so \(L_i\) is a prime ideal since its norm \(p_i\) is prime. If we set \(L = L_1^{e_1} \cdots L_k^{e_k}\) then \(L\) is primitive since its factorization into prime ideals satisfies conditions (i) and (ii). We have \(n \in L\) since each \(P_i\) is in \(L_i\). Also we have \(N(L) = N(L_1)^{e_1} \cdots N(L_k)^{e_k} = p_1^{e_1} \cdots p_k^{e_k} = n\). Thus \(Q_L(n) = N(n)/N(L) = n^2/n = n\) which means that \(Q_L\) represents \(n\) provided that \(n\) is primitive in \(L\). If \(n\) is not primitive in \(L\) then it factors as \(n = qr\) for some integer \(q > 1\) and some \(r\) in \(L\). By an earlier argument \(r\) must be a positive integer. Since \(r\) is in \(L\), we have \((r) = LM\) for some ideal \(M\). Then \((n) = (qr) = qLM\). We also have \((n) = L\overline{L}\) since \(N(L) = n\). Thus \(qLM = L\overline{L}\) so the cancellation property for ideals implies that \(\overline{L} = qM\). Taking conjugates, this says \(L = q\overline{M}\). This contradicts the fact that \(L\) is primitive. Thus we have shown that \(Q_L\) represents \(n\).

We have \(Q_{L_i}(p_i) = N(p_i)/N(L_i) = p_i^2/p_i = p_i\). Thus both \(Q_i\) and \(Q_{L_i}\) represent the prime \(p_i\) so they must be equivalent, hence in \(CG(\Delta)\) we have \(Q_{L_i} = Q_i^{-1}\). We can choose the sign of the exponent at will since we are free to replace \(L_i\) by \(\overline{L}_i\) in the previous arguments. Then \(Q_1^{\pm e_1} \cdots Q_k^{\pm e_k} = Q_{L_1}^{e_1} \cdots Q_{L_k}^{e_k} = Q_L\) since \(L = L_1^{e_1} \cdots L_k^{e_k}\). Thus \(Q_1^{\pm e_1} \cdots Q_k^{\pm e_k}\) represents \(n\) since \(Q_L\) represents \(n\) \(\square\).

The numbers represented by the principal form \(Q\) of a given discriminant \(\Delta\) are just the norms of primitive elements of \(R\). For other forms there is an analogous
Corollary 8.32. For a fundamental discriminant $\Delta$ the positive numbers represented by a form $Q$ of discriminant $\Delta$ are exactly the norms of primitive ideals strictly equivalent to $L_Q$.

Proof: This is more a corollary of the proof of the preceding theorem than of the theorem itself. Suppose first that the positive number $n$ is represented by $Q$. The first part of the preceding proof produced a primitive ideal $L$ strictly equivalent to $L_Q$ with $n$ represented by $Q_L$ as $n = Q_L(n)$ and with $N(L) = n$. Conversely if $n = N(L)$ for a primitive ideal $L$ strictly equivalent to $L_Q$ then $Q_L(n) = N(n)/N(L) = n^2/n = n$. Let $L$ factor into prime ideals as $L = P_1 \cdots P_k$ with $N(P_i)$ a prime $p_i$, so $n = N(L) = N(P_1) \cdots N(P_k) = p_1 \cdots p_k$. In the next-to-last paragraph of the preceding proof we showed that under these conditions $n$ is primitive in $L$. Thus $n$ is represented by $Q_L$ as $n = Q_L(n)$. The form $Q$ is properly equivalent to $Q_L$ since $L$ is strictly equivalent to $L_Q$, so $Q$ also represents $n$. $\Box$
Chapter 8 Quadratic Fields

Exercises

1. (a) Show that if \( \alpha \) and \( \beta \) are elements of \( \mathbb{Z}[\sqrt{D}] \) such that \( \alpha \) is a unit times \( \beta \), then \( N(\alpha) = \pm N(\beta) \).
(b) Either prove or give a counterexample to the following statement: If \( \alpha \) and \( \beta \) are Gaussian integers with \( N(\alpha) = N(\beta) \) then \( \alpha \) is a unit times \( \beta \).

2. Show that a Gaussian integer \( x + yi \) with both \( x \) and \( y \) odd is divisible by \( 1 + i \) but not by \( (1 + i)^2 \).

3. There are four different ways to write the number \( 1105 = 5 \cdot 13 \cdot 17 \) as a sum of two squares. Find these four ways using the factorization of 1105 into primes in \( \mathbb{Z}[i] \). [Here we are not counting \( 5^2 + 2^2 \) and \( 2^2 + 5^2 \) as different ways of expressing 29 as the sum of two squares. Note that an equation \( n = a^2 + b^2 \) is equivalent to an equation \( n = (a + bi)(a - bi) \).]

4. (a) Find four different units in \( \mathbb{Z}[\sqrt{3}] \) that are positive real numbers, and find four that are negative.
(b) Do the same for \( \mathbb{Z}[\sqrt{11}] \).

5. Make a list of all the Gaussian primes \( x + yi \) with \( -7 \leq x \leq 7 \) and \( -7 \leq y \leq 7 \). (The only actual work here is to figure out the primes \( x + yi \) with \( 0 \leq y \leq x \leq 7 \), then the rest are obtainable from these by symmetry properties.)

6. Factor the following Gaussian integers into primes in \( \mathbb{Z}[i] \): \( 3 + 5i, 8 - i, 10 + i, 5 - 12i, 35i, -35 + 120i, 253 + 204i \).

7. In this problem we consider \( \mathbb{Z}[\sqrt{-2}] \). To simplify notation, let \( \omega = \sqrt{-2} \), so elements of \( \mathbb{Z}[\omega] \) are sums \( x + y\omega \) with \( x, y \in \mathbb{Z} \) and with \( \omega^2 = -2 \). We have \( N(x + y\omega) = x^2 + 2y^2 = (x + y\omega)(x - y\omega) \).
(a) Draw the topograph of \( x^2 + 2y^2 \) including all values less than 70 (by symmetry, it suffices to draw just the upper half of the topograph). Circle the values that are prime (prime in \( \mathbb{Z} \), that is). Also label each region with its \( x/y \) fraction.
(b) Which primes in \( \mathbb{Z} \) factor in \( \mathbb{Z}[\omega] \)?
(c) Using the information in part (a), list all primes in \( \mathbb{Z}[\omega] \) of norm less than 70.
(d) Draw a diagram in the \( xy \)-plane showing all elements \( x + y\omega \) in \( \mathbb{Z}[\omega] \) of norm less than 70 as small dots, with larger dots or squares for the elements that are prime in \( \mathbb{Z}[\omega] \). (There is symmetry, so the primes in the first quadrant determine the primes in the other quadrants.)
(e) Show that the only primes \( x + y\omega \) in \( \mathbb{Z}[\omega] \) with \( x \) even are \( \pm \omega \). (Your diagram in part (d) should give some evidence that this is true.)
(f) Factor \( 4 + \omega \) into primes in \( \mathbb{Z}[\omega] \).

8. (a) According to Proposition 8.11, unique factorization fails in \( \mathbb{Z}[\sqrt{D}] \) when \( D = -3 \) since the number \( D(D - 1) = 12 \) has two distinct prime factorizations in \( \mathbb{Z}[\sqrt{D}] \). On
the other hand, when we enlarge \( \mathbb{Z}[\sqrt{-3}] \) to \( \mathbb{Z}[\omega] \) for \( \omega = \frac{1 + \sqrt{-3}}{2} \) unique factorization is restored. Explain how the two prime factorizations of 12 in \( \mathbb{Z}[\sqrt{-3}] \) give rise to the same prime factorization in \( \mathbb{Z}[\omega] \) (up to units).

(b) Do the same thing for the case \( D = -7 \).

9. Find a recursive formula for a primitive solution of \( x^2 + 7y^2 = 2^k \), showing how a solution for one value of \( k \) gives rise to a solution for the next larger value of \( k \).
Bibliography

— A delightful potpourri showing off many of the wonders of numbers.


— A classical and accessible introduction to number theory.

— Many illuminating pictures, with chapters on topographs and quadratic forms.

— A well-written standard textbook by a master of the subject.

— A pleasing intermingling of algebra and geometry.

— One of the few elementary treatments of binary (two-variable) quadratic forms.

— First published in 1962 under the more fitting title “A Second Course in Number Theory”.

— A scholarly historical study by one of the 20th century greats.

— The next step after studying quadratic curves.

— A masterful expositor writing at the graduate level, in spite of the title.

— Gives the complete answer to the question of which primes can be written in the form $x^2 + ny^2$. Quite a bit of deep mathematics is involved.

And finally two historical references:

— The first book ever written about quadratic forms, presenting the author’s groundbreaking research.

— This article (in German) is where the Farey diagram first appeared.
Glossary of Nonstandard Terminology

In a few instances we have chosen not to use standard terminology for certain concepts, usually because the traditional names seem somewhat awkward in the context of this book, or not as suggestive of the meaning as they could be. Here is a short summary of the main instances where translation may be needed when reading other sources.

**Quadratic Forms.** These are usually divided into three types, but for our purposes it is useful to split one of the three types into two for a total of four types as defined at the beginning of Chapter 5. Here are the traditional names with our equivalents:

- definite = elliptic
- indefinite = hyperbolic or 0-hyperbolic
- semi-definite = parabolic

Besides the convenience of having separate names for hyperbolic and 0-hyperbolic forms, the other motivation for the change is that the ordinary meanings of “definite” and “indefinite” do not seem to convey very well their mathematical meanings.

What we call a symmetry of a quadratic form is more often called an automorph or automorphism of the form, although the latter terms are sometimes reserved just for orientation-preserving symmetries. We call a form having an orientation-reversing symmetry a mirror symmetric form, or a form with mirror symmetry, whereas classically such forms are called ambiguous, a term that has suffered somewhat in the translation from Gauss’s original Latin.

**Primes.** There is uniform agreement about what a prime number is when one is talking about positive integers, namely a number greater than 1 that is divisible only by itself and 1, and for the sake of consistency we use the natural extension of this definition to other sorts of “integers” considered in this book, namely integers in quadratic fields \( \mathbb{Q}(\sqrt{d}) \). Thus we call such integers prime if the only way they factor is with one factor a unit (and they are not units themselves). Over the years it has become more usual to call numbers with this property irreducible rather than prime, using the term prime for numbers with the property that if they divide a product, then they must divide one of the factors. For example in the ring \( \mathbb{Z}[\sqrt{-5}] \) the number 2 is prime according to our definition but not according to the standard definition since 2 divides \( 6 = (1 + \sqrt{-5})(1 - \sqrt{-5}) \) but does not divide either of these two factors.

We make a similar divergence from standard terminology when we define prime ideals in Chapter 8.

**Topographs.** Of much more recent origin is Conway’s notion of the topograph of a quadratic form. Here we do not always follow Conway’s picturesque terminology. What we call a separator line he called a river, and our source vertices and edges are his simple and double wells. He called a region with label 0 a lake but we just call this a 0-region.
Here we enumerate the proper equivalence classes of primitive forms for negative discriminants. The first column gives the discriminant (up to sign), with an asterisk when it is not a fundamental discriminant. The second column gives the class number. In most cases the class group is cyclic so the class number determines the class group in these cases. The exceptions are indicated by an additional entry such as $2 \cdot 2$ to denote the product of two cyclic groups of order 2. The third column gives the various characters for each discriminant. These correspond to the prime divisors of the discriminant, with a few exceptions involving nonfundamental discriminants and the prime 2. The fourth column gives the reduced form for each equivalence class, with $ax^2 + bxy + cy^2$ abbreviated to $[a, b, c]$, followed by signs $+$ and $-$ indicating whether the characters have value $+1$ or $-1$ on each form. The forms in each genus have the same character values, and these are listed consecutively. Forms that lack mirror symmetry have middle coefficients $\pm b$, indicating that the form and its mirror image give distinct elements of the class group.

<p>| $|\Delta|$ | $h_\Delta$ | Char. | Forms       | $|\Delta|$ | $h_\Delta$ | Char. | Forms       |
|-----------|------------|-------|-------------|-----------|------------|-------|-------------|
| 3         | 1          | $x_3$ | $[1, 1, 1] +$ | 31        | 3          | $x_{31}$ | $[1, 1, 8] +$ |
| 4         | 1          | $x_4$ | $[1, 0, 1] +$ | 32        | 2          | $x_4, x_8$ | $[1, 0, 8] +$ |
| 7         | 1          | $x_7$ | $[1, 1, 2] +$ | 34        | 2          | $x_5, x_7$ | $[1, 1, 9] +$ |
| 8         | 1          | $x_8'$| $[1, 0, 2] +$ | 35        | 2          | $x_5, x_7$ | $[1, 1, 9] +$ |
| 11        | 1          | $x_{11}$ | $[1, 1, 3] +$ | 36        | 2          | $x_4, x_3$ | $[1, 0, 9] +$ |
| * 12      | 1          | $x_3$ | $[1, 0, 3] +$ | 39        | 4          | $x_3, x_{13}$ | $[1, 1, 10] +$ |
| 15        | 2          | $x_3, x_5$ | $[1, 1, 4] ++$ | 40        | 2          | $x_8', x_5$ | $[1, 0, 10] +$ |
| * 16      | 1          | $x_4$ | $[1, 0, 4] +$ | 43        | 1          | $x_{43}$ | $[1, 1, 11] +$ |
| 19        | 1          | $x_{19}$ | $[1, 1, 5] +$ | 44        | 3          | $x_4, x_{11}$ | $[1, 0, 11] +$ |
| 20        | 2          | $x_4, x_5$ | $[1, 0, 5] ++$ | 47        | 5          | $x_{47}$ | $[1, 1, 12] +$ |
| 23        | 3          | $x_{23}$ | $[1, 1, 6] +$ | 48        | 2          | $x_4, x_3$ | $[1, 0, 12] +$ |
| 24        | 2          | $x_8, x_3$ | $[1, 0, 6] ++$ |           |            |       |              |
| * 27      | 1          | $x_3$ | $[1, 1, 7] +$ |           |            |       |              |
| * 28      | 1          | $x_7$ | $[1, 0, 7] +$ |           |            |       |              |</p>
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<td>$X_5, X_{19}$</td>
<td>[1, 1, 24] + [4, ±1, 6] ++ [5, 5, 6] ++ [2, ±1, 12] -- [3, 1, 8] --</td>
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<tr>
<td>* 64</td>
<td>2</td>
<td>$X_4, X_8$</td>
<td>[1, 0, 16] ++ [4, 4, 5] +</td>
<td>* 96</td>
<td>4</td>
<td>$2 \cdot 2$</td>
<td>$X_4, X_8, X_3$</td>
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<tr>
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<td>2</td>
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<td>[1, 1, 25] + [5, 1, 5] --</td>
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<td>68</td>
<td>4</td>
<td>$X_4, X_{17}$</td>
<td>[1, 0, 17] ++ [2, 2, 9] ++ [3, ±2, 6] --</td>
<td>* 100</td>
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<td>[1, 0, 25] + [2, 2, 13] --</td>
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<td>71</td>
<td>7</td>
<td>$X_{71}$</td>
<td>[1, 1, 18] + [2, ±1, 9] + [3, ±1, 6] + [4, ±3, 5] +</td>
<td>* 103</td>
<td>5</td>
<td>$X_{103}$</td>
<td>[1, 1, 26] + [2, ±1, 13] + [4, ±3, 7] +</td>
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<td>104</td>
<td>6</td>
<td>$X_8, X_{13}$</td>
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<td>[1, 1, 27] + [3, ±1, 9] +</td>
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<td>3</td>
<td>$X_{19}$</td>
<td>[1, 0, 19] + [4, ±2, 5] +</td>
<td>* 108</td>
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<td>[1, 0, 27] + [4, ±2, 7] +</td>
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<td>[1, 1, 20] + [2, ±1, 10] + [4, ±1, 5] +</td>
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Table 2. Forms of Positive Nonsquare Discriminant.

This table is similar in layout to Table 1. For positive discriminants there is not a unique reduced form within each equivalence class so we have chosen a form which seemed simplest in some less precise sense.

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<th>Δ</th>
<th>$h_\Delta$</th>
<th>Char.</th>
<th>Forms</th>
<th>Δ</th>
<th>$h_\Delta$</th>
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<th>Forms</th>
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<td>[1, 0, −15] +</td>
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</table>
Listed below are the 101 known negative discriminants $\Delta$ for which every primitive form has a mirror-symmetric topograph. This is equivalent to saying that each genus consists of a single equivalence class of forms, or that the class group is either the trivial group or a product of cyclic groups of order 2. The class number $h_{\Delta}$ is then a power of 2 determined by the number of distinct prime divisors of $\Delta$. Asterisks in the table denote nonfundamental discriminants. Among the 101 discriminants there are 65 fundamental discriminants and, coincidentally, 65 even discriminants.

| $|\Delta|$ | $h_{\Delta}$ | $|\Delta|$ | $h_{\Delta}$ | $|\Delta|$ | $h_{\Delta}$ |
|---|---|---|---|---|---|
| 3 | 1 | 120 = $2^3 \cdot 3 \cdot 5$ | 4 | 555 = $3 \cdot 5 \cdot 37$ | 4 |
| $4 = 2^2$ | 1 | 123 = $3 \cdot 41$ | 2 | 595 = $5 \cdot 7 \cdot 17$ | 4 |
| 7 | 1 | 132 = $2^2 \cdot 3 \cdot 11$ | 4 | 627 = $3 \cdot 11 \cdot 19$ | 4 |
| $8 = 2^3$ | 1 | * 147 = $3 \cdot 49$ | 2 | 660 = $2^2 \cdot 3 \cdot 5 \cdot 11$ | 8 |
| 11 | 1 | 148 = $4 \cdot 37$ | 2 | * 672 = $2^5 \cdot 3 \cdot 7$ | 8 |
| * 12 = $2^2 \cdot 3$ | 1 | * 160 = $2^5 \cdot 5$ | 4 | 708 = $2^2 \cdot 3 \cdot 59$ | 4 |
| 15 = $3 \cdot 5$ | 2 | 163 | 1 | 715 = $5 \cdot 11 \cdot 13$ | 4 |
| * 16 = $2^4$ | 1 | 168 = $2^3 \cdot 3 \cdot 7$ | 4 | 760 = $2^3 \cdot 5 \cdot 19$ | 4 |
| 19 | 1 | * 180 = $2^2 \cdot 3^2 \cdot 5$ | 4 | 795 = $3 \cdot 5 \cdot 53$ | 4 |
| 20 = $2^2 \cdot 5$ | 2 | 187 = $11 \cdot 17$ | 2 | 840 = $2^3 \cdot 3 \cdot 5 \cdot 7$ | 8 |
| 24 = $2^3 \cdot 3$ | 2 | * 192 = $2^6 \cdot 3$ | 4 | * 928 = $2^5 \cdot 29$ | 4 |
| * 27 = $3^3$ | 1 | 195 = $3 \cdot 5 \cdot 13$ | 4 | * 960 = $2^6 \cdot 3 \cdot 5$ | 8 |
| * 28 = $2^2 \cdot 7$ | 1 | 228 = $2^2 \cdot 3 \cdot 19$ | 4 | 1012 = $2^5 \cdot 11 \cdot 23$ | 4 |
| * 32 = $2^5$ | 2 | 232 = $2^3 \cdot 29$ | 2 | 1092 = $2^2 \cdot 3 \cdot 7 \cdot 13$ | 8 |
| 35 = $5 \cdot 7$ | 2 | 235 = $5 \cdot 47$ | 2 | * 1120 = $2^5 \cdot 5 \cdot 7$ | 8 |
| * 36 = $2^2 \cdot 3^2$ | 2 | * 240 = $2^4 \cdot 3 \cdot 5$ | 4 | 1155 = $3 \cdot 5 \cdot 7 \cdot 11$ | 8 |
| 40 = $2^3 \cdot 5$ | 2 | 267 = $3 \cdot 89$ | 2 | * 1248 = $2^5 \cdot 3 \cdot 13$ | 8 |
| 43 | 1 | 280 = $2^3 \cdot 5 \cdot 7$ | 4 | 1320 = $2^3 \cdot 3 \cdot 5 \cdot 11$ | 8 |
| * 48 = $2^4 \cdot 3$ | 2 | * 288 = $2^5 \cdot 3^2$ | 4 | 1380 = $2^3 \cdot 3 \cdot 5 \cdot 23$ | 8 |
| 51 = $3 \cdot 17$ | 2 | 312 = $2^3 \cdot 3 \cdot 13$ | 4 | 1428 = $2^2 \cdot 3 \cdot 7 \cdot 17$ | 8 |
| 52 = $2^2 \cdot 13$ | 2 | * 315 = $3^2 \cdot 5 \cdot 7$ | 4 | 1435 = $5 \cdot 7 \cdot 41$ | 4 |
| * 60 = $2^2 \cdot 3 \cdot 5$ | 2 | 340 = $2^2 \cdot 5 \cdot 17$ | 4 | 1540 = $2^2 \cdot 5 \cdot 7 \cdot 11$ | 8 |
| * 64 = $2^6$ | 2 | * 352 = $2^5 \cdot 11$ | 4 | * 1632 = $2^5 \cdot 3 \cdot 17$ | 8 |
| 67 | 1 | 372 = $2^2 \cdot 3 \cdot 31$ | 4 | 1848 = $2^3 \cdot 3 \cdot 7 \cdot 11$ | 8 |
| * 72 = $2^3 \cdot 3^2$ | 2 | 403 = $13 \cdot 31$ | 2 | 1995 = $3 \cdot 5 \cdot 7 \cdot 11$ | 8 |
| * 75 = $3 \cdot 5^2$ | 2 | 408 = $2^3 \cdot 3 \cdot 17$ | 4 | * 2080 = $2^5 \cdot 5 \cdot 13$ | 8 |
| 84 = $2^2 \cdot 3 \cdot 7$ | 4 | 420 = $2^2 \cdot 3 \cdot 5 \cdot 7$ | 8 | 3003 = $3 \cdot 7 \cdot 11 \cdot 13$ | 8 |
| 88 = $2^3 \cdot 11$ | 2 | 427 = $7 \cdot 61$ | 2 | * 3040 = $2^5 \cdot 5 \cdot 19$ | 8 |
| 91 = $7 \cdot 13$ | 2 | 435 = $3 \cdot 5 \cdot 29$ | 4 | 3315 = $3 \cdot 5 \cdot 13 \cdot 17$ | 8 |
| * 96 = $2^5 \cdot 3$ | 4 | * 448 = $2^6 \cdot 7$ | 4 | * 3360 = $2^5 \cdot 3 \cdot 5 \cdot 7$ | 16 |
| * 99 = $3^2 \cdot 11$ | 2 | * 480 = $2^5 \cdot 3 \cdot 5$ | 8 | * 5280 = $2^5 \cdot 3 \cdot 5 \cdot 11$ | 16 |
| * 100 = $2^2 \cdot 5^2$ | 2 | 483 = $3 \cdot 7 \cdot 23$ | 4 | 5460 = $2^2 \cdot 3 \cdot 5 \cdot 7 \cdot 13$ | 16 |
| * 112 = $2^4 \cdot 7$ | 2 | 520 = $2^3 \cdot 5 \cdot 13$ | 4 | * 7392 = $2^5 \cdot 3 \cdot 7 \cdot 11$ | 16 |
| 115 = $5 \cdot 23$ | 2 | 532 = $2^2 \cdot 7 \cdot 19$ | 4 | 283 |
Table 4. Periodic Separator Lines.

The dotted vertical lines are lines of mirror symmetry and the heavy dots along the separator lines are points of rotational skew symmetry.

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