Classical and modular approaches to exponential Diophantine equations
II. The Lebesgue–Nagell equation

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Abstract
This is the second in a series of papers where we combine the classical approach to exponential Diophantine equations (linear forms in logarithms, Thue equations, etc.) with a modular approach based on some of the ideas of the proof of Fermat’s Last Theorem. In this paper we use a general and powerful new lower bound for linear forms in three logarithms, together with a combination of classical, elementary and substantially improved modular methods to solve completely the Lebesgue–Nagell equation $x^2 + D = y^n$, $x$, $y$ integers, $n \geq 3$, for $D$ in the range $1 \leq D \leq 100$.

1. Introduction

Arguably, the two most celebrated achievements of the 20th century in the field of Diophantine equations have been Baker’s theory of linear forms in logarithms, and Wiles’ proof of Fermat’s Last Theorem. We call Baker’s approach to Diophantine equations the ‘classical approach’. The proof of Fermat’s Last Theorem is based on what we term the ‘modular approach’. The proponents of the classical approach are too many to mention; the modular approach is still in its infancy, but among the early contributors let us just mention Frey, Serre, Ribet, Darmon, Merel, Kraus, Bennett, Skinner, Ivorra, etc.

The motivation for our series of papers, of which this is the second, is that neither approach (on its own and as it stands at the moment) is powerful enough to resolve unconditionally many of the outstanding exponential Diophantine equations. Our thesis is that one should, where possible, attack exponential Diophantine equations by a combination of classical and modular approaches. The precise aims of this series were formulated in our first paper [BMS] as follows.

(I) To present theoretical improvements to some aspects of the classical approach.

(II) To show how local information obtained through the modular approach can be used to reduce the size of the bounds, both for exponents and for variables, of solutions to exponential Diophantine equations.

(III) To show how local information obtained through the modular approach can be pieced together to provide a proof that there are no missing solutions less than the bounds obtained in (I), (II).

(IV) To solve various famous exponential Diophantine equations.

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In [BMS] we gave a new lower bound for linear forms in three logarithms and used a combination of classical and modular methods to determine all the perfect powers in the Fibonacci and Lucas sequences. In the present paper, we apply a more general and powerful lower bound for linear forms in three logarithms due to Mignotte [Mig], together with a combination of elementary, classical and substantially improved modular methods to study the following exponential Diophantine equation:

\[ x^2 + D = y^n, \quad x, y \text{ integers, } n \geq 3. \]  

(1)

Here, \( D \) denotes a non-zero integer. We have chosen to name this equation the Lebesgue–Nagell equation; the reason for the name Lebesgue–Nagell is given in § 2, together with some historical remarks. However, for now we mention that the equation has previously been solved for 81 values of \( D \) in the range \( 1 \leq D \leq 100 \), using elementary, classical and modular methods; the remaining values are apparently beyond these methods as they stand. We prove the following theorem.

**Theorem 1.** All solutions to (1) with \( D \) in the range

\[ 1 \leq D \leq 100 \]  

(2)

are given in the table in Appendix A. In particular, the only integer solutions \((x, y, n)\) to the generalised Ramanujan–Nagell equation

\[ x^2 + 7 = y^n, \quad n \geq 3, \]

satisfy \(|x| = 1, 3, 5, 11, 181\).

We choose to give a complete proof of Theorem 1, rather than treating the 19 remaining values of \( D \) in the range (2).

It is noted that the solutions for even \( n \) can be deduced quickly, for then \( D \) is expressible as a difference of squares. It is therefore sufficient to solve the equation

\[ x^2 + D = y^p, \quad x, y \text{ integers, } p \geq 3 \text{ is prime}; \]

(3)

the solutions to (1) can then be recovered from the solutions to (3).

We give three modular methods for attacking (3). Two are refinements of known methods and a third that is completely new. Using a computer program based on these modular methods, we can show, for any \( D \) in the above range, that the exponent \( p \) is large (showing that \( p > 10^9 \) is quite practical). Our modular approach also yields the following rather surprising result: either each prime factor of \( y \) divides \( 2D \) or \( y > (\sqrt{p} - 1)^2 \). We are then able to deduce not only that \( p \) is large, but also that \( y \) is large. This information helps us to reduce the size of the upper bound on \( p \) obtained from the lower bound for the linear forms in three logarithms, making the computation much more practical. This idea of using the modular approach to force lower bounds for solutions of Diophantine equations was used previously, for instance by Bennett [Ben04]. Our total computer time for the computations in this paper is roughly 206 days on various workstations (the precise details are given in due course).

Using our approach should make it possible to solve (1) for any \( D \), with \(|D| \) not too large, that is not of the form \( D = -a^2 \pm 1 \); if \( D \) is of this form then (1) has a solution \((x, y) = (a, \pm 1)\) for all odd values of the exponent \( n \), and the modular methods we explain later are not very successful in this situation. To deal with this case requires further considerations, which we leave for another paper. Note, however, that the case \( D = 1 \) turns out to be quite easy and was solved in 1850 by Lebesgue [Leb50]. Furthermore, (1) has been solved for some negative values of \( D \) of the form \( D = -a^2 \pm 1 \), including \( D = -1 \) and \( D = -8 \) (see, for example, [Ivo03, Sik03]). However, proving that the only integer solutions to \( x^2 - 2 = y^n \) with \( n \geq 3 \) satisfy \(|x| = 1 \) remains a challenging open question. For some modest progress on this question, due to the authors, see [Sik].
2. On the history of the Lebesgue–Nagell equation

Equation (1) has a long and glorious history and there are literally hundreds of papers devoted to special cases of this equation. Most of these are concerned with (1) either for special values of \( n \) or special values of \( y \). For example, for \( D = 2 \) and \( n = 3 \), Fermat asserted that he had shown that the only solutions are given by \( x = 5, y = 3 \); a proof was given by Euler [Eul70]. Equation (1) with \( n = 3 \) is the intensively studied Mordell equation (see [GPZ98] for a modern approach).

Another notable special case is the generalized Ramanujan–Nagell equation

\[
x^2 + D = k^n,
\]

where \( D \) and \( k \) are given integers. This is an extension of the Ramanujan–Nagell equation \( x^2 + 7 = 2^n \), proposed by Ramanujan [Ram13] in 1913 and first solved by Nagell [Nag48] in 1948 (see also the collected papers of Nagell [Nag02]). This equation has exactly five solutions with \( x \geq 1 \) (see [Mig84] for a very simple proof) and is, in this respect, singular: indeed, Bugeaud and Shorey [BS01] established that (4) with \( D \) positive and \( k \) a prime number not dividing \( D \) has at most two solutions in positive integers \( x, n \), except for \((D, k) = (7, 2)\). They also listed all of the pairs \((D, k)\) as above for which (4) has exactly two solutions. Much earlier, Apéry, [Ape60a, Ape60b] proved by \( p \)-adic arguments that \( x^2 + D = k^n \), with \( k \) prime, has at most two positive integer solutions except if \((D, k) = (7, 2)\). We direct the reader to [BS01] for further results and references.

Returning to (1), the first result for general \( y, n \) seems to be the proof in 1850 by Lebesgue [Leb50] that there are no non-trivial solutions for \( D = 1 \). The next cases to be solved were \( D = 3, 5 \) by Nagell [Nag48] in 1923. It is for this reason that we call (1) the Lebesgue–Nagell equation. The case with \( D = -1 \) is particularly noteworthy: a solution was sought for many years as a special case of the Catalan conjecture. This case was finally settled by Ko [Ko65] in 1965.

The history of the Lebesgue–Nagell equation is meticulously documented in an important article by Cohn [Coh93b] and so we are saved the trouble of compiling an exhaustive survey. In particular, Cohn refines the earlier elementary approaches of various authors (especially of Ljunggren [Lju63, Lju64]) and completes the solution for 77 values of \( D \) in the range \( 1 \leq D \leq 100 \). The solution for the cases \( D = 74, 86 \) was completed by Mignotte and de Weger [MW96] (indeed, Cohn solved these two equations of type (3) except for \( p = 5 \), in which case difficulties occur as the class numbers of the corresponding imaginary quadratic fields are divisible by 5). Bennett and Skinner [BS04, Proposition 8.5] applied the modular approach to solve the cases \( D = 55 \) and 95. The 19 remaining values

\[
7, 15, 18, 23, 25, 28, 31, 39, 45, 47, 60, 63, 71, 72, 79, 87, 92, 99, 100,
\]

are clearly beyond the scope of Cohn’s elementary method, although Cohn’s method can still give non-trivial information even in these cases and is revisited in \( \S \) 5. Moreover, as far as we can see, the modular method used by Bennett and Skinner (which is what we call Method I) is not capable of handling these values on its own, even though it still gives useful information in most cases.

Cohn [Coh93b], also makes a challenge of proving that the only solutions to the equation

\[
x^2 + 7 = y^n
\]

have \( |x| = 1, 3, 5, 11, 181 \). This challenge was taken up by Lesage [Les98] who proved, by classical arguments, that if \( x > 181 \) then \( 5000 < n < 6.6 \times 10^{15} \) and also by Siksek and Cremona [SC03] who used the modular approach to show that there are no further solutions for \( n \leq 10^8 \) (consequently, \( n \) must be prime); they also suggested that an improvement to lower bounds in linear forms in three logarithms may finally settle the problem. With the benefit of hindsight, we know that they were almost, although not entirely, correct. The substantial improvement to lower bounds in linear forms in three logarithms used here was certainly needed. However, for this lower bound to be
even more effective, a further insight obtained from the modular approach was also needed: namely that \( y \) is large as indicated in the introduction; note further that Lesage proved that \( y > 10^9 \) by classical arguments (linear forms in two 2-adic logarithms from [BL96]) and some heavy computer verification.

### 3. Reduction to Thue equations

Our main methods for attacking (3) are linear forms in logarithms (to bound \( p \)) and the modular approach, although for some small values of \( p \) it is necessary to reduce the equation to a family of Thue equations. The method for reducing (3) to Thue equations is well known. We do, however, feel compelled to give a succinct recipe for this, in order to set up notation that is needed later.

It is appropriate to point out that there are other approaches that could be used to solve (3) for small \( p \). For \( p = 3 \) we can view the problem as that of finding integral points on an elliptic curve, a problem that is aptly dealt with in the literature (see [Sma98, GPZ98]). For \( p \geq 5 \), the equation \( x^2 + D = p^r \) defines a curve of genus \( \geq 2 \); one can sometimes determine all rational points on this curve using the method of Chabauty [CF96], although this would also require computing the Mordell–Weil group of the Jacobian (see [PS97, Sch95a, Sto98, Sto01, Sto02]).

We do not assume in this section that \( D \) is necessarily in the range (2), merely that \( -D \) is not a square. We write (here and throughout the paper)

\[
D = D_1^2D_2, \quad D_1, D_2 \text{ are integers, } D_2 \text{ square-free.}
\]

Let \( \mathcal{L} = \mathbb{Q}(\sqrt{-D_2}) \) and \( \mathcal{O} \) be its ring of integers. Throughout the present paper, we denote the conjugate of an element \( \alpha \) (respectively of an ideal \( \mathfrak{a} \)) by \( \overline{\alpha} \) (respectively by \( \overline{\mathfrak{a}} \)).

Let \( p_1, \ldots, p_r \) be the prime ideals of \( \mathcal{O} \) dividing \( 2D \). Let \( A \) be the set of integral ideals \( \mathfrak{a} \) of \( \mathcal{O} \) such that:

- \( \mathfrak{a} = p_1^{a_1} \cdots p_r^{a_r} \), with \( 0 \leq a_i < p_i \);
- \( \gcd(\mathfrak{a}, \overline{\mathfrak{a}}) \) divides \( 2D_1\sqrt{-D_2} \);
- \( \overline{\mathfrak{a}} \) is a perfect \( p \)th power.

If \((x, y)\) is a solution to (3), then one effortlessly sees that

\[
(x + D_1\sqrt{-D_2})\mathcal{O} = \mathfrak{a}\mathfrak{b}^p
\]

for some \( \mathfrak{a} \in A \) and some integral ideal \( \mathfrak{b} \).

Now let \( b_1, \ldots, b_h \) be integral ideals forming a complete set of representatives for the ideal class group of \( \mathcal{O} \). Thus, \( \mathfrak{b}_i \) is a principal ideal for some \( i \) and so \( \mathfrak{b}_i = \beta_i'\mathcal{O} \) for some \( \beta_i' \in \mathcal{O} \). The fractional ideal \( \mathfrak{a}\mathfrak{b}_i^{-p} \) is easily seen to be also principal. The ideal \( \mathfrak{b} \) is unknown, but the ideals, \( \mathfrak{a}, \mathfrak{b}_1, \ldots, \mathfrak{b}_h \) are known. We may certainly determine which of the fractional ideals \( \mathfrak{a}\mathfrak{b}_i^{-p} \) are principal. Let \( \Gamma' \) be a set containing one generator \( \gamma' \) for every principal ideal of the form \( \mathfrak{a}\mathfrak{b}_i^{-p} \) (\( \mathfrak{a} \in A \) and \( 1 \leq i \leq h \)). It is noted that the elements of \( \Gamma' \) are not necessarily integral, but we know that if \((x, y)\) is a solution to (3) then

\[
(x + D_1\sqrt{-D_2})\mathcal{O} = \gamma'\beta'^p\mathcal{O},
\]

for some \( \gamma' \in \Gamma' \) and some \( \beta' \in \mathcal{O} \). Finally, define \( \Gamma \) as follows:

\[
\Gamma = \begin{cases} 
\Gamma', & \text{if } D_2 > 0, \ D_2 \neq 3, \text{ or if } D_2 = 3 \text{ and } p \neq 3, \\
\bigcup \epsilon^j\Gamma', & \text{if } D_2 = p = 3, \text{ where } \zeta = (1 + \sqrt{-3})/2, \\
\Gamma' \cup \zeta\Gamma' \cup \zeta^{-1}\Gamma', & \text{if } D_2 < 0, \text{ where } \epsilon \text{ runs over } -(p-1)/2, \ldots, (p-1)/2,
\end{cases}
\]

where if \( D_2 < 0 \) (and so \( \mathcal{L} \) is real) we write \( \epsilon \) for the fundamental unit.
We quickly deduce the following.

Proposition 3.1. With notation as above, if \((x, y)\) is a solution to (3) then there exist \(\gamma \in \Gamma\) and \(\beta \in \mathcal{O}\) such that
\[
x + D_1 \sqrt{-D_2} = \gamma \beta^p.
\]
Thus if we let 1, \(\omega\) be an integral basis for \(\mathcal{O}\) then for some \(\gamma \in \Gamma\),
\[
x = \frac{1}{2}(\gamma(U + V\omega)^p + \overline{\gamma}(U + V\overline{\omega})^p)
\]
for some integral solution \((U, V)\) to the Thue equation
\[
\frac{1}{2\sqrt{-D_2}}(\gamma(U + V\omega)^p - \overline{\gamma}(U + V\overline{\omega})^p) = D_1.
\]

3.1 Results I

If \(q\) is a prime we denote by \(v_q : \mathbb{Z} \rightarrow \mathbb{Z}_{\geq 0} \cup \{\infty\}\) the normalized \(q\)-adic valuation.

We now eliminate all cases where it is inconvenient to carry out level-lowering.

Lemma 3.2. Suppose that \(1 \leq D \leq 100\). Suppose that \((x, y, p)\) is a solution to (3) that is missing from our table in Appendix A. Then \(p\) satisfies the following conditions:
\[
\begin{cases}
    p \geq 7, \\
    p \geq v_q(D)+1, & \text{for all primes } q, \\
    p \geq v_2(D)+7, & \text{if } v_2(D) \text{ is even.}
\end{cases}
\] (6)

Proof. It is clear that for any particular \(D\) there are only a handful of primes \(p\) violating any of these conditions. We wrote a \texttt{pari/gp} [BBBCO] program that solved all (3) for \(p\) violating (6): the program first reduces each such equation to a family of Thue equations as in Proposition 3.1 above. These are then solved using the built-in \texttt{pari/gp} function for solving Thue equations (this is an implementation of the method of Bilu and Hanrot [BH96]).

It is perhaps worthwhile to record here two tricks that helped us in this step. First, in writing down the set \(\Gamma\) appearing in Proposition 3.1 we needed a set of integral ideals \(b_1, \ldots, b_h\) representing the ideal class group of the quadratic field \(L\). Both \texttt{pari/gp} and \texttt{MAGMA} [BCP97] have built-in functions that amount to homomorphisms from the ideal class group as an abstract group, to the set of fractional ideals, and these can be used to construct the required set \(b_1, \ldots, b_h\). We have found, however, that we get much simpler Thue equations if we search for the smallest prime ideal representing each non-trivial ideal class, and of course taking \(1\) to represent the trivial ideal class.

To introduce the second trick, we recall that when one is faced with a Thue equation
\[
a_0U^p + a_1U^{p-1}V + \cdots + a_pV^p = b
\]
it is usual to multiply throughout by \(a_0^{p-1}\) and make the substitution \(U' = a_0U\), thus obtaining a monic polynomial on the left-hand side. When \(a_0\) is large, this greatly complicates the equation. The second trick is to first search for a unimodular substitution, which makes the leading coefficient \(a_0\) small.

After optimizing our program, we were able to complete the proof of Lemma 3.2 in about 22 minutes on a 1050 MHz UltraSPARC III computer.

4. Removing common factors

It is desirable when applying the modular approach to (3) to remove the possible common factors of the three terms in the equation. This desire leads to a subdivision of cases according to the possible
common factors, as seen in the following elementary lemma. Here and elsewhere, for a non-zero integer \( a \), the product of the distinct prime divisors of \( a \) is called the radical of \( a \), and denoted by \( \text{rad}(a) \), in particular \( \text{rad}(\pm 1) = 1 \). Furthermore, \( (\cdot) \) stands for the Kronecker symbol.

**Lemma 4.1.** Suppose that \((x, y, p)\) is a solution to (3) such that \( y \neq 0 \) and \( p \) satisfies the condition (6). Then there are integers \( d_1, d_2 \) such that the following conditions are satisfied:

(i) \( d_1 > 0 \);
(ii) \( D = d_1^2 d_2 \);
(iii) \( \gcd(d_1, d_2) = 1 \);
(iv) for all odd primes \( q \mid d_1 \) we have \((\frac{-d_1}{q}) = 1\);
(v) if \( 2 \mid d_1 \) then \( d_2 \equiv 7 \pmod{8} \).

Moreover, there are integers \( s, t \) such that

\[
x = d_1 t, \quad y = \text{rad}(d_1) s,
\]

and

\[
t^2 + d_2 = es^p, \quad \gcd(t, d_2) = 1, \quad s \neq 0,
\]

where

\[
e = \prod_{q \text{ prime } \mid d_1} q^{p-2v_q(d_1)} \quad \text{and} \quad \text{rad}(e) = \text{rad}(d_1).
\]

**Proof.** Suppose that \((x, y, p)\) is a solution to (3) such that \( y \neq 0 \) and condition (6) is satisfied. It is straightforward to see that condition (6) forces \( \gcd(x^2, D) \) to be a square, say \( d_1^2 \) with \( d_1 > 0 \). We can therefore write \( x = d_1 t \) and \( D = d_1^2 d_2 \) for some integers \( t \) and \( d_2 \). Moreover, because

\[
d_1^2 = \gcd(x^2, D) = \gcd(d_1^2 t^2, d_1^2 d_2) = d_1^2 \gcd(t^2, d_2),
\]

we see that \( \gcd(t, d_2) = 1 \). Removing the common factors from \( x^2 + D = y^p \) we obtain \( t^2 + d_2 = es^p \) where \( e \) is given by (8). The integrality of \( e \) follows from the condition (6), and so does the equality of the radicals \( \text{rad}(e) = \text{rad}(d_1) \). Note that (iii) follows from this equality of the radicals and the fact that \( t, d_2 \) are coprime. We have thus proven (i), (ii), (iii) and it is now easy to deduce (iv) and (v). Finally, the condition \( s \neq 0 \) follows from the condition \( y \neq 0 \). \( \square \)

**Definition.** Suppose that \( D \) is a non-zero integer and \((x, y, p)\) is a solution to (3) with \( y \neq 0 \) and \( p \) satisfying (6). Let \( d_1, d_2 \) be as in the above lemma and its proof (thus \( d_1 > 0 \), \( \gcd(x, D) = d_1^2 \) and \( d_2 = D/d_1^2 \)). We call the pair \((d_1, d_2)\) the signature of the solution \((x, y, p)\). We call the pair \((t, s)\) the simplification of \((x, y)\) (or \((t, s, p)\) the simplification of \((x, y, p)\)).

In this terminology, Lemma 4.1 associates with any \( D \) a finite set of possible signatures \((d_1, d_2)\) for the solutions \((x, y, p)\) of (3) satisfying (6) and \( y \neq 0 \). To solve (3) it is sufficient to solve it under the assumption that the solution’s signature is \((d_1, d_2)\) for each possible signature.

**Example 1.** For example, if \( D = 25 \), there are two possible signatures satisfying the conditions of Lemma 4.1; these are \((d_1, d_2) = (1, 25)\) or \((5, 1)\). If \((d_1, d_2) = (1, 25)\), then \( x = t \), \( y = s \) and we must solve the equation

\[
t^2 + 25 = s^p, \quad 5 \mid t,
\]

already solved by Cohn. However, if \((d_1, d_2) = (5, 1)\), then \( x = 5t \), \( y = 5s \), and we must solve the equation

\[
t^2 + 1 = 5^{p-2}s^p,
\]

not solved by Cohn. In either case it is noted that the three terms of the resulting equation are relatively coprime, which is important to apply the modular approach.
5. A simplification of Cohn

We will soon apply our modular machinery to (3) with \(D\) in the range (2). Before doing this it is helpful to introduce a simplification due to Cohn that will drastically reduce the amount of computation needed later. All the arguments presented in this section are found in Cohn’s papers [Coh93b, Coh03]. Cohn, however, assumed that \(D \neq 7 \pmod{8}\); the result below is not subject to this restriction.

**Proposition 5.1.** Let \(D = D_2^2 D_2\) where \(D_2\) is square-free and \(D_2 > 0\). Suppose that \((x, y, p)\) is a solution to (3) with \(p\) satisfying (6) and let \((d_1, d_2)\) be the signature of this solution. Then:

(i) \(d_1 > 1\); or
(ii) \(p \equiv 7 \pmod{8}\) and \(2 | y\); or
(iii) \(p\) divides the class number \(h\) of the quadratic field \(\mathbb{Q}(\sqrt{-D_2})\); or
(iv) \(y = a^2 + D_2b^2\) for some integers \(a\) and \(b\) such that \(b | D_2\), \(b \neq \pm D_1\),

\[\frac{1}{2\sqrt{-D_2}}[(U + b\sqrt{-D_2})^p - (U - b\sqrt{-D_2})^p] = D_1;\]
or

(v) \(D = 1, (x, y) = (0, 1)\); or

(vi) \(D_2 \equiv 3 \pmod{4}\) and \(y = (a^2 + D_2b^2)/4\) for some odd integers \(a\) and \(b\) such that \(b | D_2\)

\[\frac{1}{2\sqrt{-D_2}}[(U + b\sqrt{-D_2})^p - (U - b\sqrt{-D_2})^p] = 2^p D_1.\]

**Proof.** We only give a brief sketch. Suppose that (i), (ii), (iii) are false. Then \(x + D_1 \sqrt{-D_2}\) = \(p\) for some \(\alpha\) in the ring of integers of \(\mathbb{Q}(\sqrt{-D_2})\). There are two possibilities. The first is that \(\alpha = a + b\sqrt{-D_2}\) for some integers \(a\) and \(b\). By equating the imaginary parts we deduce all of (iv) if \(b \neq \pm D_1\). Thus, suppose that \(b = \pm D_1\). Letting \(\beta = a - b\sqrt{-D_2}\) we see that

\[\frac{\alpha^p - \beta^p}{\alpha - \beta} = \pm 1.\]

If \(\alpha/\beta\) is not a root of unity, then the left-hand side is the \(p\)th term of a Lucas sequence (with \(p \geq 7\)) and a deep theorem of Bilu et al. [BHV01] on primitive divisors of Lucas and Lehmer sequences immediately gives a contradiction. Thus \(\alpha/\beta\) is a root of unity, i.e. \(\alpha/\beta = \pm 1, \pm i\), or \((\pm 1 \pm \sqrt{-3})/2\).

Each case turns out to be impossible, except for \(\alpha = -\beta\), which together with \(b = \pm D_1\) implies (v).

The second possibility for \(\alpha\) is that \(\alpha = (a + b\sqrt{-D_2})/2\) with \(a, b\) odd integers (and \(-D_2 \equiv 1 \pmod{4}\)). Now (vi) follows quickly by equating the imaginary parts of \((x + D_1 \sqrt{-D_2}) = \alpha^p.\)

5.1 Results II

**Corollary 5.2.** Suppose that \(D\) belongs to our range (2) and \((x, y, p)\) is a solution to (3) with \(p\) satisfying the condition (6). If the solution \((x, y, p)\) is missing from the table in Appendix A, then either \(D \equiv 7 \pmod{8}\) and \(2 | y\) or \(d_1 > 1\), where \((d_1, d_2)\) is the signature of the solution.

**Proof.** We apply Proposition 5.1. Using a short MAGMA program we listed all solutions arising from possibilities (iv)–(vi) of Proposition 5.1 with \(1 \leq D \leq 100\). The only solutions found in our range are \((x, y, p) = (0, 1, p)\) for \(D = 1\) and \((x, y, p) = (\pm 8, 2, 7)\) for \(D = 64\) and these are certainly in the table in Appendix A.

To prove the corollary we merely have to take care of possibility (iii) of the proposition. For \(1 \leq D \leq 100\) and primes \(p\) satisfying (6), the only case when \(p\) could possibly divide the class number of \(\mathbb{Q}(\sqrt{-D_2})\) is \(p = 7\) and \(D = 71\) (in which case \(h = 7\)). We solved the equation \(x^2 + 71 = y^7\)

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Appendix A.

6.2 Proposition

by interchanging or the equivalent equation for the simplification (3). In [BS04], Bennett and Skinner give an exhaustive recipe for Frey curves and level lowering under the assumption that the three terms in the form \( Ax^n + By^n = Cz^2 \); it is noted that (7) is indeed of this form.

Proof. Clearly \( y \neq 0 \). We produced our list of solutions with \( x^2 \mid D \) using a short MAGMA program. □

6. Level lowering

In this section we apply the modular approach to (7) under suitable, but mild, hypotheses. Ordinarily, one would have to construct a Frey curve or curves associated with our equation, show that the Galois representation is irreducible (under suitable hypotheses) using the results of Mazur and others [Maz78] and modular by the work of Wiles and others [Wil95, TW95, BCDT01], and finally apply Ribet’s level-lowering theorem [Rib90]. Fortunately we are saved much trouble by the excellent paper of Bennett and Skinner [BS04], which does all of this for equations of the form \( Ax^n + By^n = Cz^2 \); it is noted that (7) is indeed of this form.

Let \( D \) be a non-zero integer. We shall apply the modular approach to the Diophantine equation

\[
x^2 + D = y^p, \quad x^2 \nmid D, \quad y \neq 0 \text{ and } p \geq 3 \text{ is prime,}
\]

or the equivalent equation for the simplification \((s, t)\)

\[
l^2 + d_2 = es^p, \quad t \neq \pm 1, \quad \gcd(t, d_2) = 1, \quad s \neq 0,
\]

under the additional assumption that \( p \) satisfies (6). The assumptions made about \( s, t \) in (10) are there to ensure the non-singularity of the Frey curves, and the absence of complex multiplication when we come to apply the modular approach later on. Before going on we note the following lemma, which in effect states that there is no harm in making these additional assumptions for \( D \) in our range (2).

**Lemma 6.1.** There are no solutions to (3) for \( D \) in the range (2) with \( y = 0 \), or \( x^2 \mid D \), except those listed in the table in Appendix A.

**Proof.** Clearly \( y \neq 0 \). We produced our list of solutions with \( x^2 \mid D \) using a short MAGMA program. □

Lemma 4.1 associates with each equation of the form (9) finitely many signatures \((d_1, d_2)\) satisfying conditions (i)–(v) and corresponding (7). Following Bennett and Skinner [BS04], we associate a Frey curve \( E_t \) with any potential solution of (10) according to Tables 1–3.

Tables 1–3 are divided into cases (a)–(l). We know that \( d_1, d_2 \) are coprime and, hence, at most one of them is even. The possibility that \( d_1, d_2 \) are both odd is dealt with in Table 1. In cases (a), (b), a simple modulo 8 argument convinces us that \( t \) is odd. However, for cases (c) and (d), where \( d_1 \) is odd and \( d_2 \equiv 7 \pmod{8} \), the integer \( t \) can be either odd or even and we assign different Frey curves for each possibility. When \( t \) is odd (case (d)) we add the assumption that \( t \equiv 1 \pmod{4} \); this can be achieved by interchanging \( t \) with \(-t\) if necessary.

Table 2 deals with the possibility of even \( d_1 \) and Table 3 deals with the possibility of even \( d_2 \). In both of these cases \( t \) is necessarily odd and the congruence condition on \( t \) can again be achieved by interchanging \( t \) with \(-t\) if necessary.

**Proposition 6.2.** Suppose that \( D, d_1, d_2 \) are non-zero integers that satisfy (i)–(v) of Lemma 4.1. Suppose also that \( p \) is a prime satisfying (6) and let \( e \) be as defined in (8). Suppose that \((t, s)\) is a solution of (10) and satisfies the supplementary condition (if any) on \( t \) in Tables 1–3. Let \( E_t \) and \( L \) be as in these tables and write \( \rho_p(E_t) \) for the Galois representation on the \( p \)-torsion of \( E_t \). Then the representation \( \rho_p(E_t) \) arises from a cuspidal newform of weight 2 and level \( N = L \rad(D) \).

**Proof.** In [BS04], Bennett and Skinner give an exhaustive recipe for Frey curves and level lowering for equations of the form \( Ax^n + By^n = Cz^2 \) under the assumption that the three terms in the
### Classical and modular approaches

**Table 1. Frey curves with \(d_1, d_2\) odd.**

<table>
<thead>
<tr>
<th>Case</th>
<th>Condition on (d_2)</th>
<th>Condition on (t)</th>
<th>Frey curve (E_t)</th>
<th>(L)</th>
</tr>
</thead>
<tbody>
<tr>
<td>(a)</td>
<td>(d_2 \equiv 1) (mod 4)</td>
<td></td>
<td>(Y^2 = X^3 + 2tX^2 - d_2X)</td>
<td>2(^5)</td>
</tr>
<tr>
<td>(b)</td>
<td>(d_2 \equiv 3) (mod 8)</td>
<td></td>
<td>(Y^2 = X^3 + 2tX^2 + (t^2 + d_2)X)</td>
<td>2(^5)</td>
</tr>
<tr>
<td>(c)</td>
<td>(d_2 \equiv 7) (mod 8)</td>
<td>(t) even</td>
<td>(Y^2 = X^3 + 2tX^2 + (t^2 + d_2)X)</td>
<td>2(^5)</td>
</tr>
<tr>
<td>(d)</td>
<td>(d_2 \equiv 7) (mod 8)</td>
<td>(t \equiv 1) (mod 4)</td>
<td>(Y^2 + XY = X^3 + \left(\frac{t - 1}{4}\right)X^2 + \left(\frac{t^2 + d_2}{64}\right)X)</td>
<td>2</td>
</tr>
</tbody>
</table>

**Table 2. Frey curves with \(d_1\) even, \(d_2\) odd.**

<table>
<thead>
<tr>
<th>Case</th>
<th>Conditions on (t, s, p)</th>
<th>Frey curve (E_t)</th>
<th>(L)</th>
</tr>
</thead>
<tbody>
<tr>
<td>(e)</td>
<td>(t \equiv 1) (mod 4)</td>
<td>(Y^2 + XY = X^3 + \left(\frac{t - 1}{4}\right)X^2 + \left(\frac{t^2 + d_2}{64}\right)X)</td>
<td>1</td>
</tr>
</tbody>
</table>

**Table 3. Frey curves with \(d_1\) odd, \(d_2\) even.**

<table>
<thead>
<tr>
<th>Case</th>
<th>Condition on (d_2)</th>
<th>Condition on (t)</th>
<th>Frey curve (E_t)</th>
<th>(L)</th>
</tr>
</thead>
<tbody>
<tr>
<td>(f)</td>
<td>(v_2(d_2) = 1)</td>
<td></td>
<td>(Y^2 = X^3 + 2tX^2 - d_2X)</td>
<td>2(^6)</td>
</tr>
<tr>
<td>(g)</td>
<td>(d_2 \equiv 4) (mod 16)</td>
<td>(t \equiv 1) (mod 4)</td>
<td>(Y^2 = X^3 + tX^2 - \frac{d_2}{4}X)</td>
<td>2</td>
</tr>
<tr>
<td>(h)</td>
<td>(d_2 \equiv 12) (mod 16)</td>
<td>(t \equiv 3) (mod 4)</td>
<td>(Y^2 = X^3 + tX^2 - \frac{d_2}{4}X)</td>
<td>2(^2)</td>
</tr>
<tr>
<td>(i)</td>
<td>(v_2(d_2) = 3)</td>
<td>(t \equiv 1) (mod 4)</td>
<td>(Y^2 = X^3 + tX^2 - \frac{d_2}{4}X)</td>
<td>2(^4)</td>
</tr>
<tr>
<td>(j)</td>
<td>(v_2(d_2) = 4, 5)</td>
<td>(t \equiv 1) (mod 4)</td>
<td>(Y^2 = X^3 + tX^2 - \frac{d_2}{4}X)</td>
<td>2(^2)</td>
</tr>
<tr>
<td>(k)</td>
<td>(v_2(d_2) = 6)</td>
<td>(t \equiv 1) (mod 4)</td>
<td>(Y^2 + XY = X^3 + \left(\frac{t - 1}{4}\right)X^2 - \frac{d_2}{64}X)</td>
<td>2(^{-1})</td>
</tr>
<tr>
<td>(l)</td>
<td>(v_2(d_2) \geq 7)</td>
<td>(t \equiv 1) (mod 4)</td>
<td>(Y^2 + XY = X^3 + \left(\frac{t - 1}{4}\right)X^2 - \frac{d_2}{64}X)</td>
<td>1</td>
</tr>
</tbody>
</table>

equation are coprime. After a little relabeling, their results apply to \((10)\) and the lemma follows from \(\S \S 2\) and 3 of their paper. It is here that we need the assumptions \(t \neq \pm 1\) and \(s \neq 0\) made in \((10)\).

It is convenient to indulge in the following abuse of language.

**Definition.** If \((t, s, p)\) is a solution to \((10)\) and if the representation \(\rho_p(E_t)\) arises from a cuspidal newform \(f\), then we say that solution \((t, s, p)\) arises from the newform \(f\) (via the Frey curve \(E_t\)), or that the newform \(f\) gives rise to the solution \((t, s, p)\). If \((t, s, p)\) is the simplification of \((x, y, p)\) then we say that \((x, y, p)\) arises from the newform \(f\). If the newform \(f\) is rational and so corresponds to an elliptic curve \(E\), then we also say that the solution \((t, s, p)\) (or \((x, y, p)\)) arises from \(E\).

### 6.1 A summary

It may be helpful for the reader to summarize what we have done and where we are going. Given a non-zero integer \(D\) we would like to solve \((9)\). We can certainly write down all solutions with \(y = 0\) or with \(x^2|D\). We can also solve (at least in principle) all cases where \(p\) violates condition \((6)\) by reducing to Thue equations as in \(\S\) 3. We can thus reduce to \((9)\) and assume that \(p\) satisfies condition \((6)\).
Next, we can write down a list of signatures \((d_1, d_2)\) satisfying conditions (i)–(v) of Lemma 4.1. We reduce the solution of (9) to solving (10) for each signature \((d_1, d_2)\). Now we associate with the signature \((d_1, d_2)\) one or more Frey curves \(E_t\) and levels \(N\), so that any solution to (10) arises from some newform \(f\) at level \(N\) via the Frey curve \(E_t\).

Finally (and this is to come) we must show how to solve (10) under the assumption that the solution arises from a newform \(f\) via a Frey curve \(E_t\). If we can do this for each newform \(f\) at the necessary level and Frey curve \(E_t\), then we will have completed the solution of our (3).

As we shall see, the assumption that a solution arises from a particular newform is a very strong one, for it imposes congruence conditions on \(t\) modulo all but finitely many primes \(l\).

### 6.2 Congruences

For an elliptic curve \(E\) and a prime of good reduction \(l\) we write \(\sharp E(\mathbb{F}_l)\) for the number of points on \(E\) over the finite field \(\mathbb{F}_l\), and let \(a_l(E) = l + 1 - \sharp E(\mathbb{F}_l)\).

**Lemma 6.3.** With notation as above, suppose that the Galois representation \(\rho_p(E_t)\) arises from a cuspidal newform with Fourier expansion around infinity

\[
f = q + \sum_{n \geq 2} c_n q^n,
\]

of level \(N\) (given by Proposition 6.2) and defined over a number field \(K/\mathbb{Q}\). Then there is a place \(\mathfrak{P}\) of \(K\) above \(p\) such that for every prime \(l \mid 2pD\) we have

\[
\begin{align*}
a_l(E_t) &\equiv c_l \pmod{\mathfrak{P}}, \quad \text{if } t^2 + d_2 \not\equiv 0 \pmod{l} \quad \text{(or equivalently } l \nmid s), \\
l + 1 &\equiv \pm c_l \pmod{\mathfrak{P}}, \quad \text{if } t^2 + d_2 \equiv 0 \pmod{l} \quad \text{(or equivalently } l | s).
\end{align*}
\]

**Proof.** The lemma is standard (see [Ser87, p. 196], [BS04, p. 7], [Kra98, Proposition 5.4], etc.). The conditions \(l \nmid 2D\) and \(l | s\) together imply that \(l\) is a prime of good reduction for \(E_t\), whereas the conditions \(l \nmid 2D\) and \(l | s\) imply that \(l\) is a prime of multiplicative reduction. \(\square\)

When the newform \(f\) is rational, there is an elliptic curve \(E\) defined over \(\mathbb{Q}\) whose conductor is equal to the level of the newform \(f\) such that \(a_l(E) = c_l\) for all primes of good reduction \(l\). In this case we can be a little more precise than in Lemma 6.3, thanks to a result of Kraus and Oesterlé.

**Lemma 6.4.** With notation as above, suppose that the Galois representation \(\rho_p(E_t)\) arises from a rational cuspidal newform \(f\) corresponding to an elliptic curve \(E/\mathbb{Q}\). Then for all primes \(l \nmid 2D\) we have

\[
\begin{align*}
a_l(E_t) &\equiv a_l(E) \pmod{p}, \quad \text{if } t^2 + d_2 \not\equiv 0 \pmod{l} \quad \text{(or equivalently } l \nmid s), \\
l + 1 &\equiv \pm a_l(E) \pmod{p}, \quad \text{if } t^2 + d_2 \equiv 0 \pmod{l} \quad \text{(or equivalently } l | s).
\end{align*}
\]

**Proof.** This lemma does appear to be a special case of Lemma 6.3; however, we do allow in this lemma the case \(l = p\), which was excluded before. In fact, Lemma 6.3 together with a result of Kraus and Oesterlé [KO92, Proposition 3] implies that the representations \(\rho_p(E_t)\) and \(\rho_p(E)\) are semi-simply isomorphic. In this case the result of Kraus and Oesterlé also tells us that \(a_l(E_t) \equiv a_l(E) \pmod{p}\) if the prime \(l\) is a prime of good reduction for both curves, and \(a_l(E_t)a_l(E) \equiv l + 1 \pmod{p}\) if \(l\) is a prime of good reduction for one of them and a prime of multiplicative reduction for the other. Now, because \(l \nmid 2D\) we see that \(l \nmid N\), the conductor \(N\) of \(E\) (which is also the level of the newform \(f\) as given by Proposition 6.2). If \(l | s\), then \(l\) is a prime of multiplicative reduction for \(E_t\) and then \(a_l(E_t) = \pm 1\). The lemma follows. \(\square\)
7. Eliminating exponents: Method I

We now focus on equations of the form (10) where, as always, $p$ satisfies (6). Proposition 6.2 tells us that if $(t, s, p)$ is a solution to (10), then it arises from a newform of a certain level (or levels) and all of these can be determined. Let us say that these newforms are $f_1, \ldots, f_n$. Then to solve (10) it is sufficient to solve it, for each $i$, under the assumption that the solution arises from the newform $f_i$. We give three methods for attacking (10) under the assumption that the solution arises from a particular newform $f$.

If successful, the first method will prove that (10) has no solutions except possibly for finitely many exponents $p$ and these are determined by the method. This method is actually quite standard. As far as we know the basic idea is originally due to Serre [Ser87, pp. 203–204]. It is also found in many exponents $B$

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8. Eliminating exponents: Method II

The second method is adapted from the ideas of Kraus [Kra98] (see also [SC03]). It can only be applied to one prime (exponent) \( p \) at a time and, if successful, it does show that there are no solutions to (10) for that particular exponent.

Let us briefly explain the idea of this second method. Suppose that \( f \) is a newform with Fourier expansion as in (11) and suppose that \( p \geq 7 \) is a prime. We are interested in solutions to (10) arising from \( f \). Choose a small integer \( n \) so that \( l = np + 1 \) is prime with \( l \mid D \). Suppose that \( (t, s) \) is a solution to (10) arising from \( f \). Working modulo \( l \) we see that \( d_1^2 t^2 + D = y^p \) is either 0 or an \( n \)th root of unity. (Indeed \( (y^p)^n = y^{p-1} \equiv 0 \text{ or } 1 \pmod{l} \).) As \( n \) is small we can list all such \( t \) in \( \mathbb{F}_l \), and compute \( c_l \) and \( a_l(E_t) \) for each \( t \) in our list. We may then find that for no \( t \) in our list are the relations in Lemma 6.3 satisfied. In this case we have a contradiction and we deduce that there are no solutions to (10) arising from \( f \) for the exponent \( p \).

Let us now write this formally. Suppose that \( p \geq 7 \) is a prime and \( n \) an integer such that \( l = np + 1 \) is also prime and \( l \mid D \). Define

\[
\mu_n(\mathbb{F}_l) = \{ \zeta \in \mathbb{F}_l^* : \zeta^n = 1 \} \quad \text{and} \quad A(n, l) = \left\{ \zeta \in \mu_n(\mathbb{F}_l) : \left( \frac{\zeta - D}{l} \right) = 0 \text{ or } 1 \right\}.
\]

For each \( \zeta \in A(n, l) \), let \( \delta_\zeta \) be an integer satisfying

\[
\delta_\zeta^2 \equiv (\zeta - D)/d_1^2 \pmod{l}.
\]

It is convenient to write \( a_l(\zeta) \) for \( a_l(E_{\delta_\zeta}) \). We can now give our sufficient condition for the insolubility of (10) for the given exponent \( p \).

**Proposition 8.1** (Method II). Let \( D, d_1, d_2 \) be a triple of integers satisfying Lemma 4.1(i)–(v), and let \( p \geq 7 \) be a prime satisfying condition (6). Let \( f \) be a newform with Fourier expansion as in (11) defined over a number field \( K \). Suppose that there exists an integer \( n \geq 2 \) satisfying the following conditions.

(a) The integer \( l = np + 1 \) is prime, and \( l \mid D \).

(b) Either \( \left( \frac{-D}{l} \right) = -1 \), or \( p \mid N_K/Q(4 - c_1^2) \).

(c) For all \( \zeta \in A(n, l) \) we have

\[
\begin{cases}
  p \mid N_K/Q(a_l(\zeta) - c_l), & \text{if } l \equiv 1 \pmod{4}, \\
  p \mid N_K/Q(a_l(\zeta)^2 - c_1^2), & \text{if } l \equiv 3 \pmod{4}.
\end{cases}
\]

Then (10) does not have any solutions for the given exponent \( p \) arising from the newform \( f \).

**Proof.** Suppose that the hypotheses of the proposition are satisfied and that \( (t, s) \) is a solution to (10).

First we show that \( t^2 + d_2 \equiv 0 \pmod{l} \). Suppose otherwise. Thus \( t^2 + d_2 \equiv 0 \pmod{l} \) and so \( l \mid s \). In this case \( \left( \frac{-D}{l} \right) = 1 \) and from (b) we know that \( p \) does not divide \( N_K/Q(4 - c_1^2) \). However, by Lemma 6.3 we know that \( \pm c_l \equiv l + 1 \equiv 2 \pmod{\mathfrak{P}} \) for some place \( \mathfrak{P} \) of \( K \) above \( p \) and we obtain a contradiction showing that \( t^2 + d_2 \equiv 0 \pmod{l} \).

From (10) and the definition of \( e \) in (8), we see the existence of some \( \zeta \in A(n, l) \) such that

\[
d_1^2 t^2 + D \equiv \zeta \pmod{l} \quad \text{and} \quad t \equiv \pm \delta_\zeta \pmod{l}.
\]

Replacing \( t \) by \( -t \) in the Frey curve \( E_t \) has the effect of twisting the curve by \( -1 \) (this can be easily verified for each Frey curve in Tables 1–3). Thus \( a_l(\zeta) = a_l(E_t) \) if \( l \equiv 1 \pmod{4} \) and \( a_l(\zeta) = \pm a_l(E_t) \) if \( l \equiv 3 \pmod{4} \). Moreover, by Lemma 6.3, \( a_l(E_t) \equiv c_l \pmod{\mathfrak{P}} \) for some place \( \mathfrak{P} \) of \( K \) above \( p \). This clearly contradicts (c). Hence, there is no solution to (10) arising from \( f \) for the exponent \( p \).
If the newform \( f \) is rational and moreover corresponds to an elliptic curve with 2-torsion, then it is possible to strengthen the conclusion of Proposition 8.1 by slightly strengthening the hypotheses. The following variant is far less costly in computational terms as we explain below.

**Proposition 8.2 (Method II).** Let \( D, d_1, d_2 \) be a triple of integers satisfying Lemma 4.1(i)–(v), and let \( p \) be a prime satisfying condition (6). Let \( f \) be a rational newform corresponding to elliptic curve \( E/\mathbb{Q} \) with 2-torsion. Suppose that there exists an integer \( n \geq 2 \) satisfying the following conditions.

(a) The integer \( l = np + 1 \) is prime, \( l \leq p^2/4 \) and \( l \mid D \).

(b) Either \( \left( \frac{-D}{l} \right) = -1 \), or \( a_l(E)^2 \neq 4 \) (mod \( p \)).

(c) For all \( \zeta \in A(n,l) \) we have

\[
\begin{cases}
  a_l(\zeta) \neq a_l(E), & \text{if } l \equiv 1 \pmod{4}, \\
  a_l(\zeta) \neq \pm a_l(E), & \text{if } l \equiv 3 \pmod{4}.
\end{cases}
\]

Then (10) does not have any solutions for the given exponent \( p \) arising from the newform \( f \).

**Proof.** Comparing this with Proposition 8.1 we see that it is sufficient to show, under the additional assumptions, that if \( a_l(\zeta)^2 \equiv a_l(E)^2 \) (mod \( p \)) then \( a_l(\zeta) = \pm a_l(E) \), and if \( a_l(\zeta) \equiv a_l(E) \) (mod \( p \)) then \( a_l(\zeta) = -a_l(E) \).

Suppose that \( a_l(\zeta)^2 \equiv a_l(E)^2 \) (mod \( p \)) (the other case is similar). Hence, \( a_l(\zeta) \equiv \pm a_l(E) \) (mod \( p \)). Now note that both elliptic curves under consideration here have 2-torsion. Hence, we can write \( a_l(\zeta) = 2b_1 \) and \( a_l(E) = 2b_2 \) for some integers \( b_1 \) and \( b_2 \). Moreover, by the Hasse–Weil bound we know that \( |b_1| \leq \sqrt{7} \).

Thus

\[
b_1 \equiv \pm b_2 \pmod{p} \quad \text{and} \quad |b_1 + b_2|, |b_1 - b_2| \leq 2\sqrt{7} < p
\]

as \( l < p^2/4 \). Thus, \( b_1 = \pm b_2 \) and this completes the proof.

It remains to explain how this improves our computation. To apply Proposition 8.1 for some \( p \) we need to find a prime \( l \) satisfying conditions (a)–(c). The computationally expensive part is to compute \( a_l(E) = c_l \) and \( a_l(\zeta) \) for all \( \zeta \in A(n,l) \). Let us, however, consider the application of Proposition 8.2 rather than Proposition 8.1. The computation proceeds as before by checking conditions (a), (b) first. When we come to (c), we note that what we have to check is that

\[
\begin{cases}
  \sharp E_\zeta(\mathbb{F}_l) \neq l + 1 - a_l(E), & \text{if } l \equiv 1 \pmod{4}, \\
  \sharp E_\zeta(\mathbb{F}_l) \neq l + 1 \pm a_l(E), & \text{if } l \equiv 3 \pmod{4},
\end{cases}
\]

for each \( \zeta \in A(n,q) \). Rather than computing \( a_l(\zeta) \) for each such \( \zeta \), we first pick a random point in \( E_\zeta(\mathbb{F}_l) \) and check whether it is annihilated by \( l + 1 - a_l(E) \) if \( p \equiv 1 \) (mod 4) and either of the integers \( l + 1 \pm a_l(E) \) if \( p \equiv 3 \) (mod 4). Only if this is the case do we need to compute \( a_l(\zeta) \) to test condition (c). In practice, for primes \( p \approx 10^9 \), this brings a 10-fold speed-up in program run time for Method II.

**9. Eliminating exponents: Method III**

Occasionally, Methods I and II fail to establish the non-existence of solutions to an equation of the form (10) for a particular exponent \( p \) even when it does seem that this equation has no solutions. The reasons for this failure are not clear to us. We, shall, however give a third method, rather similar in spirit to Kraus’ method (Method II), but requiring stronger global information furnished by Proposition 3.1.

Suppose that \( D, d_1, d_2 \) are integers satisfying conditions (i)–(v) of Lemma 4.1. Let \( E_l \) be one of the Frey curves associated with (10) and let \( f \) be a newform of the level predicted by Proposition 6.2.
with Fourier expansion as in (11), defined over a number field $K$. Define $T_l(f)$ to be the set of $\tau \in \mathbb{F}_l$ such that:

- $p|N_{K/Q}(a_l(E_\tau) - c_l)$ and $\tau^2 + d_2 \not\equiv 0 \pmod{l}$; or
- $p|N_{K/Q}(l + 1 \pm c_l)$ and $\tau^2 + d_2 \equiv 0 \pmod{l}$.

We suppose that $-D$ is not a square and follow the notation of § 3. Fix a prime $p$ satisfying (6). Suppose that $l$ is a prime satisfying the following conditions.

(a) $l|2D$.

(b) $l = np + 1$ for some integer $n$.

(c) $(\frac{-D}{l}) = 1$, thus $l$ splits in $\mathcal{L} = \mathbb{Q}(\sqrt{-D})$, say $(l) = \mathfrak{l}_1\mathfrak{l}_2$.

(d) Each $\gamma \in \Gamma$ is integral at $l$; what we mean by this is that each $\gamma$ belongs to the intersection of the localizations $\mathcal{O}_{\mathfrak{l}_i} \cap \mathcal{O}_{\mathfrak{l}_2}$.

We denote the two natural reduction maps by $\theta_1, \theta_2 : \mathcal{O}_{\mathfrak{l}_1} \cap \mathcal{O}_{\mathfrak{l}_2} \rightarrow \mathbb{F}_l$. These of course correspond to the two squareroots for $-D_2$ in $\mathbb{F}_l$ and are easy to compute.

Now let $\Gamma_l$ be the set of $\gamma \in \Gamma$ for which there exists $\tau \in T_l(f)$ such that:

- $(d_1\tau + D_1\theta_1(\sqrt{-D_2}))^n \equiv \theta_1(\gamma)^n \pmod{\theta_1(\gamma)}$; and
- $(d_1\tau + D_1\theta_2(\sqrt{-D_2}))^n \equiv \theta_2(\gamma)^n \pmod{\theta_2(\gamma)}$.

**Proposition 9.1 (Method III).** Let $p$ be a prime satisfying condition (6). Let $S$ be a set of primes $l$ satisfying the conditions (a)–(d) above. With notation as above, if the newform $f$ gives rise to a solution $(t, s)$ to (10), then $d_1 t + D_1 \sqrt{-D_2} = \gamma \beta^p$ for some $\beta \in \mathcal{O}$ and some $\gamma \in \bigcap_{l \in S} \Gamma_l$. In particular, if $\bigcap_{l \in S} \Gamma_l$ is empty, then the newform $f$ does not give rise to any solution to (10) for this exponent $p$.

**Proof.** Suppose that $(t, s)$ is a solution to (10) arising from newform $f$ via the Frey curve $E_l$. Clearly $\theta_1(t) = \theta_2(t)$ is the simple reduction of $t$ modulo $l$. Let $\tau = \theta_1(t) = \theta_2(t) \in \mathbb{F}_l$. It follows from Lemma 6.3 that $\tau \in T_l(f)$. Let $(x, y)$ be the solution to (9) corresponding to $(t, s)$. Thus $x = d_1 t$. We know by Proposition 3.1 that

$$d_1 t + D_1 \sqrt{-D_2} = \gamma \beta^p,$$

for some $\gamma \in \Gamma$ and $\beta \in \mathcal{O}$. Applying $\theta_i$ to both sides and taking $n$th powers (where we recall that $l = np + 1$) we obtain

$$(d_1 \tau + D_1 \theta_i(\sqrt{-D_2}))^n \equiv \theta_i(\gamma)^n \theta_i(\beta)^{t-1} \pmod{\theta_i(\beta)} \quad \text{with} \quad \theta_i(\beta)^{t-1} \equiv 0 \text{ or } 1 \pmod{\theta_i(\beta)}.$$

Thus $\gamma \in \Gamma_l$ as defined above. The proposition follows.

**10. Examples**

It is clear that our three modular methods require computations of newforms of a given level. Fortunately the computer algebra suit MAGMA has a package completely devoted to such computations; the theory for these computations is explained by Cremona [Cre96] for rational newforms, and by Stein [Ste05b] in the general case. As an alternative, we could have used Stein’s Modular Forms Database [Ste05a].

**Example 2 (Absence of newforms).** Lemma 4.1 and Proposition 6.2 lead us to associate solutions to (9), where $p$ satisfies (6), with newforms of certain levels. If there are no newforms of the predicted levels, we immediately deduce that there are no solutions to (9). With the help of a MAGMA program we found all $D = 1, 2, \ldots, 100$ where there are no newforms at the predicted levels. We deduce the following result.

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Corollary 10.1. Let $D$ be an integer belonging to the list

$$4, 16, 32, 36, 64.$$ 

Then (9) does not have any solutions with $p$ satisfying condition (6).

This Corollary does not add anything new, as (1) has already been solved by Cohn’s method for $D = 4, 16, 32, 36, 64$ (but see [Ivo03, Sik03, Le02]).

Example 3. Corollary 5.2 solves (3) for all values of $D$ in the range (2) except for 21 values; these are the 19 values listed in (5) plus $D = 55, 95$. As indicated in § 2 the cases $D = 55$ and 95 have been solved by Bennett and Skinner. It is however helpful to look at the case $D = 95$ again as it shows how Methods I and III work together in harmony. There is only one possible signature $(d_1, d_2) = (1, 95)$. Thus, $t = x$, $s = y$ and we need to solve the equation

$$t^2 + 95 = s^p, \quad \text{where } p \geq 7. \quad (12)$$

As $d_1 = 1$, it follows from Corollary 5.2 that $y$ is even and so $t = x$ is odd. Replacing $t$ by $-t$ if necessary, we can assume that $t \equiv 1 \pmod{4}$. Table 1 leads us to associate the solution $(t, s, p)$ with the Frey curve

$$E_t : \quad Y^2 + XY = X^3 + \left(\frac{t-1}{4}\right)X^2 + \left(\frac{t^2 + 95}{64}\right)X.$$ 

From Proposition 6.2, we know that any solution to (12) arises from a newform of level 190. Using MAGMA we find that there are, up to Galois conjugacy, precisely four newforms at level 190. These are

$$f_1 = q - q^2 - q^3 + q^4 - q^5 + q^6 + O(q^8),$$

$$f_2 = q + q^2 - 3q^3 + q^4 + q^5 - 3q^6 - 5q^7 + O(q^8),$$

$$f_3 = q + q^2 + q^3 + q^4 + q^5 + q^6 - q^7 + O(q^8),$$

$$f_4 = q - q^2 + \phi q^3 + q^4 + q^5 - \phi q^6 + \phi q^7 + O(q^8), \quad \text{where } \phi^2 + \phi - 4 = 0.$$

The first three newforms above are rational and so correspond to the three isogeny classes of elliptic curves of conductor 190. It turns out that none of these elliptic curves have non-trivial 2-torsion. By the remarks made after Proposition 7.1 we know that Method I will be successful in eliminating all but finitely many exponents $p$. Indeed we find (in the notation of Proposition 7.1) that $B_3(f_1) = B_3(f_3) = 15$. Thus, we know that no solutions to (12) arise from the newforms $f_1$ or $f_3$, because otherwise, by Proposition 7.1, $p|15$ which contradicts $p \geq 7$. We also find that $B_3(f_4) = 2^4 \times 3$ and $B_7(f_4) = 2^4 \times 7$. Thus no solution arises from $f_4$. But,

$$B_3(f_2) = 3 \times 7, \quad B_7(f_2) = 3^2 \times 5 \times 7, \quad B_{11}(f_2) = 0,$$

$$B_{13}(f_2) = 3 \times 5 \times 7 \times 13, \quad B_{17}(f_2) = 3^2 \times 7 \times 11.$$

We deduce that there are no solutions arising from $f_2$ with exponent $p > 7$. It does, however, seem likely that there is a solution with $p = 7$. Moreover, an attempt to prove that there is no solution with $p = 7$ using Method II fails: we did not find any integer $2 \leq n \leq 100$ satisfying the conditions of Proposition 8.1.

We apply Method III (and follow the notation of § 9). Write

$$\omega = \frac{1 + \sqrt{-95}}{2}.$$ 

Taking $S = \{113, 127, 239, 337, 491\}$ we find that

$$\bigcap_{t \in S} \Gamma_t = \left\{ \frac{-528 - 2\omega}{2187} \right\}.$$
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Thus if we have any solutions at all then, by Proposition 9.1, we know that

\[(t - 1) + 2\omega = \left(\frac{-528 - 2\omega}{2187}\right)(U + V\omega)^7,\]

for some integers \(U, V\). Equating imaginary parts and simplifying we get

\[-U^7 - 1855VU^6 - 5061V^2U^5 + 214165V^3U^4 + 416605V^4U^3 - 2834013V^5U^2 - 2944375V^6U + 2818247V^7 = 2187.\]

Using pari/gp we find that the only solution to this Thue equation is given by \(U = -3, V = 0\). This shows that \((t, s) = (529, 6)\).

The reader will notice that \((t, s) = (-529, 6)\) is also a solution to (12) with \(p = 7\) but it seems to have been ‘missed’ by the method. This is not the case; we are assuming that the sign of \(t\) has been chosen so that \(t \equiv 1 \pmod{4}\). The solution \((t, s) = (-529, 6)\) arises from some other newform (probably at some different level) and via a different Frey curve that we have not determined.

**Example 4.** For our last example we look at the case where \(D = 25\). This, like 18 other cases, must be resolved by a combination of the modular approach and our lower bound for linear forms in three logarithms which is to come. We assume that \(p \geq 7\) and so \(p\) satisfies conditions (6). There are now two possible signatures \((d_1, d_2) = (1, 25)\) and \((5, 1)\) satisfying the conditions of Lemma 4.1. However, by Corollary 5.2, we may suppose that \(d_1 > 1\) and so \(d_1 = 5, d_2 = 1\). We write \(t = x/5, s = y/5\) where we know that \(t, s\) are integral by Lemma 4.1. Equation (10) becomes

\[t^2 + 1 = 5^{\rho - 2}s^\rho, \quad t \neq \pm 1.\]

Following Table 1, we associate with any solution to this equation the Frey curve

\[E_t: \quad Y^2 = X^3 + 2tX^2 - X,\]

and we know by Proposition 6.2 that any solution must arise from a newform of level 160. Using the computer algebra system MAGMA we find that there are, up to Galois conjugacy, three such newforms:

\[f_1 = q - 2q^3 - q^5 - 2q^7 + O(q^8), \]
\[f_2 = q + 2q^3 - q^5 + 2q^7 + O(q^8), \]
\[f_3 = q + 2\sqrt{2}q^3 + q^5 - 2\sqrt{2}q^7 + O(q^8).\]

The first two newforms are rational, corresponding respectively to elliptic curves 160A1 and 160B1 in Cremona’s tables [Cre96]. The third has coefficients in \(K = \mathbb{Q}(\sqrt{5})\) and is straightforward to eliminate using Method I. In the notation of Proposition 7.1 we find that if \(f_3\) does give rise to any solutions \((t, s, p)\) then \(p|B_3(f_3) = 24\). This is impossible as \(p \geq 7\) and so \(f_3\) does not give rise to any solutions.

We were unable to eliminate newforms \(f_1\) and \(f_2\) using Method I. Instead using our implementation of Method II in MAGMA we showed that there are no solutions arising from either form with \(7 \leq p \leq 100\). With our implementation of the improved Method II (Proposition 8.2) in pari/gp we showed that there are no solutions with \(100 \leq p \leq 163762845\); this took roughly 26 hours on a 2.4 GHz Pentium IV PC. The choice of where to stop the computation is of course not arbitrary, but comes out of our bound for the linear form in logarithms. We will later prove that \(p \leq 163762845\) thereby completing the resolution of this case.

### 11. Results III

We applied the methods of the previous sections to solve all (3) with \(D\) is our range (2).
Classical and modular approaches

Table 4. Computational details for Lemma 11.1 and its proof.

<table>
<thead>
<tr>
<th>$D$</th>
<th>$(d_1, d_2)$</th>
<th>$E^a$</th>
<th>$p_0$</th>
<th>Machine$^b$</th>
<th>Time</th>
</tr>
</thead>
<tbody>
<tr>
<td>7</td>
<td>(1, 7)</td>
<td>14A1</td>
<td></td>
<td>S1</td>
<td>12 h, 43 min</td>
</tr>
<tr>
<td>15</td>
<td>(1, 15)</td>
<td>30A1</td>
<td>624 271 465</td>
<td>P1</td>
<td>252 h, 50 min</td>
</tr>
<tr>
<td>18</td>
<td>(3, 2)</td>
<td>384D1, 384A1, 384G1, 384H1</td>
<td>306 111 726</td>
<td>S3</td>
<td>293 h, 14 min</td>
</tr>
<tr>
<td>23</td>
<td>(1, 23)</td>
<td>46A1</td>
<td>855 632 066</td>
<td>S2</td>
<td>477 h, 36 min</td>
</tr>
<tr>
<td>25</td>
<td>(5, 1)</td>
<td>160A1, 160B1</td>
<td>163 762 845</td>
<td>P2</td>
<td>19 h, 58 min</td>
</tr>
<tr>
<td>28</td>
<td>(2, 7)</td>
<td>14A1</td>
<td>852 830 725</td>
<td>P1</td>
<td>193 h, 41 min</td>
</tr>
<tr>
<td>31</td>
<td>(1, 31)</td>
<td>78A1</td>
<td></td>
<td>S3</td>
<td>484 h, 43 min</td>
</tr>
<tr>
<td>45</td>
<td>(3, 5)</td>
<td>480B1, 480F1, 480G1, 480H1</td>
<td>340 749 424</td>
<td>S1</td>
<td>477 h, 36 min</td>
</tr>
<tr>
<td>47</td>
<td>(1, 47)</td>
<td>94A1</td>
<td>1 555 437 629</td>
<td>S3</td>
<td>451 h, 34 min</td>
</tr>
<tr>
<td>60</td>
<td>(2, 15)</td>
<td>30A1</td>
<td>358 541 296</td>
<td>S1</td>
<td>130 h, 30 min</td>
</tr>
<tr>
<td>63</td>
<td>(1, 63)</td>
<td>42A1</td>
<td>292 825 735</td>
<td>S1</td>
<td>99 h, 45 min</td>
</tr>
<tr>
<td>71</td>
<td>(1, 71)</td>
<td>142C1</td>
<td>2 343 468 548</td>
<td>S3</td>
<td>697 h, 26 min</td>
</tr>
<tr>
<td>72</td>
<td>(3, 8)</td>
<td>96A1, 96B1</td>
<td>451 620 034</td>
<td>S1</td>
<td>316 h, 27 min</td>
</tr>
<tr>
<td>79</td>
<td>(1, 79)</td>
<td>158E1</td>
<td>1 544 381 661</td>
<td>S3</td>
<td>448 h, 47 min</td>
</tr>
<tr>
<td>87</td>
<td>(1, 87)</td>
<td>174D1</td>
<td>1 148 842 108</td>
<td>S3</td>
<td>329 h, 45 min</td>
</tr>
<tr>
<td>92</td>
<td>(2, 23)</td>
<td>46A1</td>
<td>996 255 151</td>
<td>S3</td>
<td>285 h, 10 min</td>
</tr>
<tr>
<td>99</td>
<td>(3, 11)</td>
<td>1056B1, 1056F1</td>
<td>593 734 622</td>
<td>P2</td>
<td>138 h, 46 min</td>
</tr>
<tr>
<td>100</td>
<td>(5, 4)</td>
<td>20A1</td>
<td>163 762 845</td>
<td>S1</td>
<td>21 h, 23 min</td>
</tr>
</tbody>
</table>

$^a$We give here the Cremona code for the elliptic curves $E$ as in his book [Cre96] and his online tables: http://www.maths.nott.ac.uk/personal/jec/ftp/data/INDEX.html.

$b$The machines are as follows: P1, 2.2 GHz Intel Pentium PC; P2, 2.4 GHz Intel Pentium PC; S1, Dual processor 750MHz UltraSPARC III; S2, 650 MHz UltraSPARC IIe; S3, UltraSPARCIII with 12 processors of 1050 MHz speed.

Lemma 11.1. Suppose that $1 \leq D \leq 100$ and $p$ is a prime satisfying (6). If $(x, y, p)$ is a solution to (3) that is not included in the table in Appendix A, then $D$ is one of

$$7, 15, 18, 23, 25, 28, 31, 39, 45, 47, 60, 63, 71, 72, 79, 87, 92, 99, 100.$$  (13)

Moreover, $(x, y, p)$ has signature $(d_1, d_2)$ and arises from an elliptic curve $E$ and $p > p_0$ where $E$, $p_0$ and $(d_1, d_2)$ are given by Table 4.

Proof. We wrote a MAGMA program that does the following. For each $D$ in the range (2) we write down the set of possible signatures $(d_1, d_2)$ satisfying the conditions of Lemma 4.1.

For each such pair $(d_1, d_2)$ write down the (one or two) Frey curves given by Tables 1–3, bearing in mind the information given by Corollary 5.2.

For each Frey curve we compute the conductor (given by Proposition 6.2) of the newforms giving rise to possible solutions and then write down all of these newforms.

We attempt to eliminate each newform $f$ using Method I. This involves searching for primes $l \nmid 2D$ such that (in the notation of Proposition 7.1) $B_l(f) \neq 0$. If we are successful and find such primes $l_1, \ldots, l_m$, then by Proposition 7.1 this exponent divides all of the $B_l(f)$, and so divides their greatest common divisor $B$ (say). If $B$ is divisible by any prime $p$ that satisfies condition (6), then we attempt to eliminate this possible $p$ using Method II: this involves searching for an integer $2 \leq n \leq 100$ satisfying conditions (a)–(c) of Proposition 8.1. If one such $n$ is found then we know that there are no solutions for the exponent $p$. Otherwise, we apply Method III (Proposition 9.1) to write down Thue equations leading to possible solutions.
As predicted by the comments made after Proposition 7.1, Method I succeeded with all non-rational newforms and all rational newforms corresponding to elliptic curves with only trivial 2-torsion (it also succeeded with some rational newforms corresponding to elliptic curves with non-trivial 2-torsion). Indeed, we found no solutions arising from non-rational newforms for $D$ in our range $1 \leq D \leq 100$.

We are left only with rational newforms $f$ that correspond to elliptic curves $E$ having some non-trivial 2-torsion. The details of these are documented in Table 4. For primes $p < 100$ satisfying condition (6) we attempt to show that there are no solutions arising from $E$ for the particular exponent $p$ using Method II (as before). If this fails for a particular exponent $p$, then we use Method III to write down the Thue equations leading to the possible solutions.

Our proof that $p > 100$ is now complete except that there are some Thue equations to solve. We had to solve Thue equations of degree 7 for $D = 7, 47, 79$ and 95. These were solved using pari/gp and the solutions are incorporated in the table in Appendix A. We also had to solve a Thue equation of degree 11 for $D = 23$, of degree 17 for $D = 28$ and of degree 13 for $D = 92$. We were unable to (unconditionally) solve these three Thue equations using the built-in functions of pari/gp. The reason is that, in each case, it was impossible for pari/gp to prove that the system of units it had found, although of correct rank, was maximal. We are grateful to Dr. Guillaume Hanrot for sending us his pari program for solving Thue equations without the full unit group. This program, based on [Han00], solved all three equations in a few minutes.

For the next step we implemented our improved Method II (Proposition 8.2) in pari/gp (see the remark after the proof). To complete the task and show that $p > p_0$ for any missing solution we used our pari/gp program to disprove the existence of any missing solution for each prime $100 \leq p \leq p_0$. We ran this pari/gp program on various machines as indicated in Table 4. The total computer time for this step is roughly 206 days.

Remark. The reader may be surprised that some of the computations were done in MAGMA whereas others were carried out in pari/gp. As stated earlier, MAGMA has a package for computing modular forms. This is essential for us and is unavailable in pari/gp.

For showing that $p > p_0$, it is simply not practical to use MAGMA. Here we are using the improved Method II (Proposition 8.2). The main bottleneck in Method II is computing $a_l(E)$ for primes $l$ that can be about $10^{11}$ (recall $l$ is a prime satisfying $l \equiv 1 \pmod{p}$). For this pari/gp uses the theoretically slower Shanks–Mestre method [Coh93a] rather than the theoretically faster Schoof–Elkies–Atkin [Sch95b] method used by MAGMA. However, for primes of the indicated size it seems that pari/gp is about 10 times faster than MAGMA.

The reader may also note that two of the machines we used are multiprocessor machines. The computation for each $D$ could have been speeded up considerably by parallelising. We however decided against this, so as to keep our programs simple and transparent.

12. The ‘modular’ lower bound for $y$

In this section we would like to use the modular approach to prove a lower bound for $y$ with $D$ in the range (2). Before doing this we prove a general result for arbitrary non-zero $D$.

**Proposition 12.1.** Suppose that $D$ is a non-zero integer, and $d_1, d_2$ satisfy Lemma 4.1(i)–(v). Suppose that $(t, s, p)$ is a solution to (10) arising from a rational newform $f$ via a Frey curve $E_t$. Then either $\text{rad}(s)|2d_1$ or $|s| > (\sqrt{p} - 1)^2$.

**Proof.** As the newform is rational we know that the newform $f$ corresponds to an elliptic curve $E/\mathbb{Q}$ whose conductor equals the level of $f$. 48
Suppose that $\text{rad}(s)$ does not divide $2d_1$. As $t$ and $d_2$ are coprime we see that there is some prime $l|s$ so that $l|2D$. By Lemma 6.4 we see that $p$ divides $l + 1 \pm a_l(E)$. It follows from the Hasse–Weil bound that $l + 1 \pm a_l(E) \neq 0$, and so

$$p \leq l + 1 \pm a_l(E) < (\sqrt{l} + 1)^2,$$

again using Hasse–Weil. Thus $l > (\sqrt{p} - 1)^2$. The proposition follows as $l|s$.

Corollary 12.2. Suppose that $D$ is one of the values in (13). If $(x, y, p)$ is a solution to (9) not in the table in Appendix A, then $y > (\sqrt{p} - 1)^2$.

Proof. Suppose that $D$ is in the range (2) and $(x, y, p)$ is some solution to (9) not in the table in Appendix A. From the preceding sections we know that this solution must satisfy condition (6). Moreover by Lemma 4.1, $x = d_1t$ and $y = \text{rad}(d_1)s$, where $(t, s, p)$ satisfy (10) for some $d_1$ and $d_2$ satisfying conditions (i)–(v) of that lemma.

We have determined, for $1 \leq D \leq 100$, all solutions to (10) arising from non-rational newforms (indeed there were none). Thus we may suppose that our putative solution arises from a rational newform. By Proposition 12.1 we see that either $|y| \geq |s| > (\sqrt{p} - 1)^2$ or $\text{rad}(s)|2d_1$. We must prove that the second possibility does not arise.

Suppose that $\text{rad}(s)|2d_1$. From Lemma 4.1 we see that $\text{rad}(y)|2d_1$. We first show that $\text{rad}(y) \neq 2$. For in this case we have reduced to an equation of the form $x^2 + D = 2^m$. For $|D| < 2^{96}$, Beukers [Beu81, Corollary 2] shows that $m \leq 18 + 2\log |D|/\log 2$. A short MAGMA program leads us to all the solutions to this equation for $1 \leq D \leq 100$ and we find that these are already in our table in Appendix A.

Thus, we may suppose that $\text{rad}(y)|2d_1$ and $\text{rad}(y) \neq 2$. An examination of the possible cases reveals the following possibilities

$$D = 18, 45, 72, 99 \text{ and } \text{rad}(y) = 3, \quad D = 25, 100 \text{ and } \text{rad}(y) = 5.$$

On removing the common factors, each case quickly reduces to an equation that has already been solved. For example, we must solve $x^2 + 100 = y^p$ under the assumption that $\text{rad}(y) = 5$ or equivalently the equation $x^2 + 100 = 5^m$. Removing the common factor reduces to the equation $X^2 + 4 = 5^{m-2}$. However, $X^2 + 4 = Y^m$ has already been solved and only has the solutions $(X, Y, n) = (2, 2, 3), (11, 5, 3)$. We quickly see that the only solution to $x^2 + 100 = 5^p$ is $(x, p) = (55, 5)$.

13. The linear form in logarithms

It is useful at this point to recap what we have done so far. We would like to complete the proof of Theorem 1 by showing that our table in Appendix A is not missing any solutions. So let us suppose that our table in Appendix A is missing some solution $(x, y, p)$ to (3) for some value of $D$ in our range (2). We have proved (Lemma 11.1) that $D$ is one of the values in (13). Moreover (again by Lemma 11.1 and by Corollary 12.2), any missing solution $(x, y, p)$ must satisfy

$$p > p_0, \quad y > (\sqrt{p} - 1)^2,$$

with $p_0$ given by Table 4. Our aim is to show that $p \leq p_0$: a contradiction.

From the table of values of $p_0$ we know that

$$|x|, p \geq 10^8$$

and indeed much more, although this inequality is sufficient for much of our later work. In the remainder of this paper we assume that $D$ is one of the remaining values (13), and always write
(as before) \( D = D_1^2D_2 \), where \( D_2 \) is square-free. The triple \((x, y, p)\) will always be a solution to (3) supposedly missing from our table in Appendix A and hence satisfying the above inequalities.

In this section we write down the linear form in logarithms corresponding to (3) and apply a theorem of Matveev to obtain upper bounds for the exponent \( p \). These upper bounds obtained from Matveev’s theorem are not small enough to contradict our lower bounds for \( p \) obtained in Lemma 11.1, but they are needed when we come to apply our bounds for linear forms in three logarithms given in the next section.

**Lemma 13.1.** Let \((d_1, d_2)\) be the signature of our supposedly missing solution \((x, y, p)\) (which we know from Lemma 11.1). Define

\[
d = \begin{cases} 
  d_1, & \text{for } D \not\equiv 7 \pmod{8}, \\
  2d_1, & \text{for } D \equiv 7 \pmod{8}.
\end{cases}
\]

Then \( d \) is a prime power, say \( d = q^c \) for some prime \( q \), where, moreover, \( q \) splits in \( L = \mathbb{Q}(\sqrt{-D_2}) \). Let \( k_0 \) be the smallest positive integer such that the ideal \( \mathfrak{q}^{k_0} \) is principal, say \( \mathfrak{q}^{k_0} = (\alpha_0) \). Also let

\[
k = \begin{cases} 
  k_0, & \text{if } k_0 \text{ is odd}, \\
  k_0/2, & \text{if } k_0 \text{ is even},
\end{cases}
\quad \text{and } \kappa = \begin{cases} 
  2, & \text{if } k_0 \text{ is odd}, \\
  1, & \text{if } k_0 \text{ is even},
\end{cases}
\quad \text{so that } k = \frac{\kappa k_0}{2}.
\]

Then there exists \( \gamma \in \mathcal{L} \) such that

\[
\left( \frac{x - D_1\sqrt{-D_2}}{x + D_1\sqrt{-D_2}} \right)^k = \alpha^\kappa \gamma^p, \quad \text{where } \alpha = \frac{\alpha_0}{\alpha_0}, \quad h(\alpha) = \frac{k_0 \log d}{2}, \quad h(\gamma) = \frac{k \log y}{2}.
\]

**Proof.** We begin with the factorization

\[
(x + D_1\sqrt{-D_2})(x - D_1\sqrt{-D_2}) = y^p.
\]

Our first step is to show that any prime divisor \( q \) of \( y \) splits in \( \mathcal{L} \). Suppose otherwise, then we may write \((q) = \mathfrak{q} \) or \((q) = \mathfrak{q}^2 \) for some prime ideal \( \mathfrak{q} \) satisfying \( \mathfrak{q} = \mathfrak{q} \). If \( p = 2r + 1 \) then clearly \( q^r \) divides both factors on the left-hand side above and so divides \( 2D_1\sqrt{-D_2} \). This is impossible in view of the fact that \( p \) is enormous and \( 1 \leq D \leq 100 \). Thus, we have shown that every prime divisor \( q \) of \( y \) splits in \( \mathcal{L} \). Put

\[
y = \prod_{i \in I} q_i^{a_i} \quad \text{and} \quad (q_i) = q_i\mathfrak{q}_i, \quad q_i \neq \mathfrak{q}_i, \quad i \in I, \quad \text{then} \quad (x + D_1\sqrt{-D_2}) = \prod_{i \in I} (q_i^{b_i}\mathfrak{q}_i),
\]

where we assume (for ease of notation) that \( b_i \geq c_i \) for all \( i \). Thus

\[
(x - D_1\sqrt{-D_2}) = \prod_{i \in I} (q_i^{c_i}\mathfrak{q}_i), \quad \text{with } b_i + c_i = pa_i, \quad \text{for all } i \in I.
\]

Then, clearly,

\[
\mathfrak{n} := \gcd(x + D_1\sqrt{-D_2}, x - D_1\sqrt{-D_2}) = \prod_{i \in I} (q_i^{c_i}) = \prod_{i \in I} (q_i^{c_i}).
\]

This shows that \( \mathfrak{n} = \langle d \rangle \) where \( d \in \mathbb{Z} \). We would like to calculate this \( d \) and verify that its value is in agreement with (16). From the definition of \( \mathfrak{n} \) we see that \( d | 2x \) and \( d | 2D_1 \). However, by our definition of signature, \( \gcd(x^2, D) = d_1^2 \). It follows that \( d^2 | d_1^2 \) and so \( d | 2d_1 \). However, \( d | x \) and \( d | D_1 \). Hence, \( d_1 | \mathfrak{n} \) and so \( d_1 | d \). Thus, \( d = d_1 \) or \( d = d_2d_1 \). We note the following cases.

- If \( D_2 \not\equiv 7 \pmod{8} \) then \( 2 | y \). Thus \( 2 | d \) and so \( d = d_1 \).
- Suppose that \( D_2 \equiv 7 \pmod{8} \). Now from Lemma 4.1 and its proof we know that \( D = d_1^2d_2 \) and \( x = d_1t \) where \( \gcd(t, d_2) = \gcd(d_1, d_2) = 1 \). Clearly \( d_2 = d_2D_2 \) with \( d_3 = D_1/d_1 \) integral.
Suppose first that \( d_1 \) is even. It follows easily that \( t, d_2 \) are odd and
\[
(d) = d = 2d_1 \left( \frac{t + d_3 \sqrt{-D_2}}{2}, \frac{t - d_3 \sqrt{-D_2}}{2} \right).
\]

Hence \( (2d_1) = d \) and so \( d = 2d_1 \).

- The only case left to consider is \( D_2 \equiv 7 \pmod{8} \) and \( d_1 \) is odd. By examining Table 4 we see that \( d_1 = 1 \). Thus, \( 2|y \) by Corollary 5.2. Clearly \( x \) is odd and the same argument as above shows that \( d = 2 = 2d_1 \).

This proves that \( d \) satisfies (16). By looking again at the possible values of \( d_1 \) in Table 4 we see that \( d \) is a prime power in all cases. Let \( j \in I \) such that \( d = \bar{q}_j \). Thus, \( c_i = 0 \) for all \( j \neq i \). Then
\[
(x + D_1 \sqrt{-D_2}) = \bar{q}_j^{c_j} \cdot \prod_{j \neq i} q_j^{p_{ai}},
\]
whence
\[
(x + D_1 \sqrt{-D_2}) = (\bar{q}_j q_j^{-1})^{c_j} \cdot \prod_{j \in I} q_j^{p_{ai}} = (\bar{a}^{-1}) g^p,
\]
where \( a \) and \( g \) are integral ideals with \( a = \bar{q}_j \), \( N(a) = \bar{q}_j \), \( d, N(g) = y \), and \( N \) denotes the norm. Thus, as ideals,
\[
\left( \frac{x - D_1 \sqrt{-D_2}}{x + D_1 \sqrt{-D_2}} \right) = (\bar{a}^{-1})^2 (\bar{g}^{-1})^p.
\]

We define \( k_0, k, \kappa, \alpha_0 \) as in the statement of the lemma. Thus, \( a^{k_0} = (\alpha_0) \) and we have the relation (between ideals)
\[
(x + D_1 \sqrt{-D_2})^k = (\bar{a}/\bar{a})^k \bar{g}^kp = a^{2k}(N(a))^{-k} g^k = (\alpha_0)^{2k}(\alpha_0)^{-k} g^k.
\]
However, \( p \) is an enormous prime certainly not dividing the class number. This shows that \( g^k \) is also principal, \( g^k = (\gamma_0) \), say, where \( \gamma_0 \) is an algebraic integer chosen so that the following equality of elements of \( \mathcal{L} \) holds
\[
(x + D_1 \sqrt{-D_2})^k = \alpha_0^d d^{-k} \gamma_0^p, \quad \text{with } N(\alpha_0) = d^{k_0} \text{ and } N(\gamma_0) = y^k.
\]
Put \( \alpha = \bar{\alpha}_0/\alpha_0 \) and \( \gamma = \pm \gamma_0/\gamma_0 \). The proof of the lemma is complete except for the statements about the heights of \( \alpha, \gamma \). These follow from Lemma 13.2 below.

### Lemma 13.2

Let \( \alpha \) be an algebraic number whose conjugates are all (including \( \alpha \) itself) of modulus equal to 1, then \( h(\alpha) = (\log a)/\deg \alpha \), where \( a \) is the leading coefficient of the minimal polynomial of \( \alpha \). In particular, if \( \alpha = \bar{\alpha}_0/\alpha_0 \) where \( \alpha_0 \) is a non-real quadratic irrationality, then \( h(\alpha) = \frac{1}{2} \log N(\alpha_0) \).

**Proof.** Set \( d = \deg \alpha \). By hypothesis \( \alpha \) is a root of a polynomial of \( \mathbb{Z}[X] \) of the form \( P(X) = aX^d + \cdots \). We have \( h(\alpha) = \frac{1}{2} \log M(P) \), where \( M \) is Mahler’s measure, and the first result easily follows because the roots of \( P \) are of modulus 1. This proves the first assertion. The (easy) proof of the second assertion is omitted. \[\square\]

We now write the linear form in three logarithms. Define
\[
\Lambda = \log \left( \frac{x - D_1 \sqrt{-D_2}}{x + D_1 \sqrt{-D_2}} \right),
\]
where we have taken the principal determination of the logarithm.

### Lemma 13.3

We have
\[
\log |\Lambda| \leq \frac{p}{2} \log y + \log(2.2D_1 \sqrt{D_2}).
\]
Proof. We will rely on the lower bounds (15). Clearly
\[
\left| \frac{x - D_1 \sqrt{-D_2}}{x + D_1 \sqrt{-D_2}} - 1 \right| < 2 \frac{D_1 \sqrt{D_2}}{|x|}.
\]
A standard inequality [Sma98, Lemma B.2] shows that
\[
|\Lambda| < 2.1 \frac{D_1 \sqrt{D_2}}{|x|},
\]
so that
\[
\log |\Lambda| < -\log |x| + \log(2.1D_1 \sqrt{|D_2|}).
\]
Using the fact that \(y^n - x^2 = D\) and a similar argument to that above, we deduce the lemma. \(\square\)

To bound \(p\) we use the theory of linear forms of (at most three) logarithms. We need the special case of three logarithms of a theorem of Matveev.

**Theorem 2** (Matveev). Let \(\lambda_1, \lambda_2, \lambda_3\) be \(\mathbb{Q}\)-linearly independent logarithms of non-zero algebraic numbers and let \(b_1, b_2, b_3\) be rational integers with \(b_1 \neq 0\). Define \(\alpha_j = \exp(\lambda_j)\) for \(j = 1, 2, 3\) and
\[
\Lambda = b_1 \lambda_1 + b_2 \lambda_2 + b_3 \lambda_3.
\]
Let \(D\) be the degree of the field \(\mathbb{Q}(\alpha_1, \alpha_2, \alpha_3)\) over \(\mathbb{Q}\). Put \(\chi = [\mathbb{R}(\alpha_1, \alpha_2, \alpha_3) : \mathbb{R}]\). Let \(A_1, A_2, A_3\) be positive real numbers, which satisfy
\[
A_j \geq \max\{D\text{th}(\alpha_j), |\lambda_j|, 0.16\} \quad (1 \leq j \leq 3).
\]
Assume that \(B \geq \max\{1, \max\{|b_j|/A_j / A_1; 1 \leq j \leq 3\}\}\). Then
\[
\log |\Lambda| > -C_1 D^2 A_1 A_2 A_3 \log(1.5eDB \log(eD)),
\]
where
\[
C_1 = \frac{5 \times 16.5}{6\chi} e^3(7 + 2\chi) \left(\frac{3\chi}{2}\right) (20.2 + \log(3^{5.5}\mathcal{D}^2 \log(eD))).
\]
In particular, for \(D = 2\) and \(\chi = 2\), this gives
\[
\log |\Lambda| > -1.80741 \times 10^{11} A_1 A_2 A_3 \log(13.80736B) \quad (17)
\]
For a proof see [Mat00].

**13.1 A preliminary bound for \(p\)**
It follows from Lemma 13.1 that
\[
k\Lambda = \kappa \log \alpha + p \log \gamma + iq\pi = \kappa \log \alpha + p \log \gamma + r \log(-1), \quad r \in \mathbb{Z},
\]
which appears as a linear form of logarithms. However, a small transformation of this form leads to better estimates. Write
\[
k\Lambda = \kappa \log(\varepsilon_1 \alpha) + p \log(\varepsilon_2 \gamma) + iq\pi, \quad q \in \mathbb{Z},
\]
where \(\varepsilon_1\) and \(\varepsilon_2 = \pm 1\) are chosen so that \(|\log(\varepsilon_1 \alpha)| < \pi/2\) and \(|\log(\varepsilon_2 \gamma)| < \pi/2\), where we take principal values for the logarithms and \(q\) such that \(|\Lambda|\) is minimal.

Remark. Indeed, we can take any roots of unity in \(\mathcal{L}\) for \(\varepsilon_1\) and \(\varepsilon_2\). The only relevant case for our set of outstanding values of \(D\) are \(D = 25, 100\), where \(\mathcal{L} = \mathbb{Q}(\sqrt{-1})\), whence we can realize \(|\log(\varepsilon_1 \alpha)| < \pi/4\) and \(|\log(\varepsilon_2 \gamma)| < \pi/4\), and we write
\[
\Lambda = 2 \log \alpha + p \log \gamma + q \log \zeta, \quad \text{where } \zeta = e^{i\pi/2}.
\]
We now return to the general case. By Lemma 13.3
\[
\log |k\Lambda| \leq -\frac{p}{2} \log y + \log(2.2kD_1\sqrt{D_2}).
\]
Our lower bounds for \(x, y\) and \(p\) imply that \(\log |k\Lambda|\) is very small and it is straightforward to deduce that \(|r| \leq (p + 1)/2\). We can write \(k\Lambda\) in the form
\[
k\Lambda = b_1\lambda_1 + b_2\lambda_2 + b_3\lambda_3
\]
with \(b_1 = \kappa\) (= 1 or 2), \(\alpha_1 = \varepsilon_1\alpha\), \(b_2 = p\), \(\alpha_2 = \varepsilon_2\gamma\), \(b_3 = q\), \(\alpha_3 = -1\) and
\[
h(\alpha_1) = \frac{k}{\kappa} \log d, \quad \lambda_1 = \log \alpha_1, \quad h(\alpha_2) = \frac{k \log y}{2}, \quad |\lambda_2| < \pi/2, \quad h(\alpha_3) = 0, \quad \lambda_3 = i\pi,
\]
extcept for the case \(\mathcal{L} = \mathbb{Q}(\sqrt{-1})\) studied in the previous remark where \(\lambda_3 = i\pi/2\).

Applying Theorem 2, we have \(D = \chi = 2\) and we can take
\[
A_1 = \max\left\{\frac{2k \log d}{\kappa}, \frac{\pi}{2}\right\}, \quad A_2 = \max\left\{k \log y, \frac{\pi}{2}\right\}, \quad A_3 = \pi
\]
and (using some change of notation in Theorem 2) \(B = p + 1\) (this choice of \(B\) is justified by the inequality \(|q| \leq (p + 1)/2\) proved above), and we get
\[
p \leq C_2 k^2 \log(2D_1) \log p.
\]
This implies \(p \leq C_3 k^2 \log(2D_1) \log(k^2 \log(2D_1))\), and thus
\[
p \leq C_4 D_2 \log(2D_1) \log(D_2 \log(2D_1)),
\]
where the constants are easily made explicit.

**Lemma 13.4.** Suppose that \(D\) is one of the remaining values \((13)\) and \((x, y, p)\) is a solution to \((9)\) missing from our table in Appendix A.

- If \(D = 7\) then \(p < 6.81 \times 10^{12}\).
- Otherwise, if \(D\) is square-free, then \(p < 1.448 \times 10^{15}\).
- For other values of \(D\), we have \(p < 3.966 \times 10^{14}\).

**Proof.** This is a simple application of Matveev’s Theorem 2. If \(D = 7\) it is easy to show that the \(\alpha_0\) arising in Lemma 13.1 is (up to conjugation) \((1 + \sqrt{-7})/2\), we know that \(k = 1\); thus \(N(\alpha_0) = 2\) and \(\Im(\log \alpha_0) = 1.2094292028\ldots\). Then we can apply \((17)\) with \(A_1 = \pi/2, A_2 = \log y, \log A_3 = \pi\) and \(B = p + 1\). After a few iterations we get the stated bound on \(p\).

In the application of Theorem 2, we can take, for all the square-free values of \(D\),
\[
A_1 = \begin{cases} 7 \log 2, & \text{if } k_0 \text{ is odd}, \\ 8 \log 2, & \text{if } k_0 \text{ is even}, \end{cases} \quad A_2 = \begin{cases} 7 \log y, & \text{if } k_0 \text{ is odd}, \\ 4 \log y, & \text{if } k_0 \text{ is even}, \end{cases} \quad A_3 = \pi,
\]
so that \(A_1A_2 \leq 49 \log 2 \times \log y\) and then we get \(p < 1.448 \times 10^{15}\).

For all the remaining values of \(D\), we can take
\[
A_1 = \begin{cases} \log 10, & \text{if } h = 1, \\ \pi/2, & \text{if } h = 2, \\ 3 \log 2, & \text{if } h = 3, \end{cases} \quad A_2 = \begin{cases} \log y, & \text{if } h = 1, \\ \log y, & \text{if } h = 2, \\ 3 \log y, & \text{if } h = 3, \end{cases} \quad A_3 = \pi,
\]
so that \(A_1A_2 \leq 9 \log 2 \times \log y\), and we get now \(p < 3.966 \times 10^{14}\).
14. A new estimate on linear forms in three logarithms

14.1 Statement of the result

We shall apply the following theorem.

Theorem 3. We consider three non-zero algebraic numbers $\alpha_1$, $\alpha_2$ and $\alpha_3$, all $\neq 1$, which are either all real or all complex of modulus one. Moreover, we assume that

\[
\text{either } \alpha_1, \alpha_2 \text{ and } \alpha_3 \text{ are multiplicatively independent, or }
\]

\[
\text{two multiplicatively independent, the third a root of unity } \neq 1.
\]

We also consider three non-zero rational integers $b_1$, $b_2$, $b_3$ with $\gcd(b_1,b_2,b_3) = 1$ and the linear form

\[\Lambda = b_1 \log \alpha_1 + b_2 \log \alpha_2 + b_3 \log \alpha_3 \neq 0,\]

where the $\log \alpha_i$ are arbitrary determinations of the logarithm, but which are all real or all purely imaginary. Without loss of generality, we assume that

\[b_2 | \log \alpha_2| = |b_1 \log \alpha_1| + |b_3 \log \alpha_3| \pm |\Lambda|.
\]

Let $K$, $L$, $R$, $R_1$, $R_2$, $R_3$, $S$, $S_1$, $S_2$, $S_3$, $T$, $T_1$, $T_2$, $T_3$ be rational integers that are all $\geq 3$, with

\[L \geq 5, \quad R > R_1 + R_2 + R_3, \quad S > S_1 + S_2 + S_3, \quad T > T_1 + T_2 + T_3.\]

Let $\rho > 2$ be a real number. Assume first that

\[
\left( \frac{KL}{2} + \frac{L}{4} - 1 - \frac{2K}{3L} \right) \log \rho \geq (D+1) \log N + gL(a_1 R + a_2 S + a_3 T) + \mathcal{D}(K-1) \log b - 2 \log(e/2),
\]

where $N = K^2L$, $\mathcal{D} = [\mathbb{Q}(\alpha_1, \alpha_2, \alpha_3) : \mathbb{Q}] / [\mathbb{R}(\alpha_1, \alpha_2, \alpha_3) : \mathbb{R}]$, $e = \exp(1)$,

\[g = \frac{1}{4} - \frac{N}{12RST}, \quad b = (b_2 \eta_0)(b_2 \zeta_0) \left( \prod_{k=1}^{K-1} k! \right)^{-4/K(K-1)},
\]

with

\[\eta_0 = \frac{R-1}{2} + \frac{(S-1)b_1}{2b_2}, \quad \zeta_0 = \frac{T-1}{2} + \frac{(S-1)b_3}{2b_2},
\]

and

\[a_i \geq \rho |\log \alpha_i| - \log |\alpha_i| + 2\mathcal{D}h(\alpha_i), \quad i = 1, 2, 3.
\]

If, for some positive real number $\chi$ and $V := (((R_1 + 1)(S_1 + 1)(T_1 + 1)))^{1/2}$:

(i) $(R_1 + 1)(S_1 + 1)(T_1 + 1) > K \max\{R_1 + S_1 + 1, S_1 + T_1 + 1, R_1 + T_1 + 1, \chi V\}$;

(ii) $\text{Card}\{\alpha_1 \alpha_2 \alpha_3 : 0 \leq r \leq R_1, 0 \leq s \leq S_1, 0 \leq t \leq T_1\} > L;

(iii) $(R_2 + 1)(S_2 + 1)(T_2 + 1) > 2K^2;$

(iv) $\text{Card}\{\alpha_1 \alpha_2 \alpha_3 : 0 \leq r \leq R_2, 0 \leq s \leq S_2, 0 \leq t \leq T_2\} > 2KL$; and

(v) $(R_3 + 1)(S_3 + 1)(T_3 + 1) > 6K^2L$;

then either

\[\Lambda' > \rho^{-KL},\]

where

\[\Lambda' = |\Lambda| \cdot \max\left\{ \frac{LRe^{LR|\Lambda|/(2b_1)}}{2|b_1|}, \frac{LSe^{LS|\Lambda|/(2b_2)}}{2|b_2|}, \frac{LT e^{LT|\Lambda|/(2b_3)}}{2|b_3|} \right\},\]

or at least one of the following conditions (C1), (C2), (C3) hold:

\[|b_1| \leq R_i \quad \text{and} \quad |b_2| \leq S_i \quad \text{and} \quad |b_3| \leq T_i, \quad (i = 1, 2) \quad (C_i)\]
(C3) either there exist two non-zero rational integers $r_0$ and $s_0$ such that
\[ r_0b_2 = s_0b_1, \]
with
\[ |r_0| \leq B_S := \frac{(R_1 + 1)(T_1 + 1)}{\chi V - \max\{R_1, T_1\}} \quad \text{and} \quad |s_0| \leq B_R := \frac{(S_1 + 1)(T_1 + 1)}{\chi V - \max\{S_1, T_1\}}. \]

or there exist rational integers $r_1$, $s_1$, $t_1$ and $t_2$, with $r_1s_1 \neq 0$, such that
\[(t_1b_1 + r_1b_3)s_1 = r_1b_2t_2, \quad \gcd(r_1, t_1) = \gcd(s_1, t_2) = 1,
\]
which also satisfy, for $\delta = \gcd(r_1, s_1)$,
\[ 0 < |r_1s_1|/\delta \leq B_T := \frac{(R_1 + 1)(S_1 + 1)}{\chi V - \max\{R_1, S_1\}}, \quad |s_1t_1|/\delta \leq B_R \quad \text{and} \quad |r_1t_2|/\delta \leq B_S. \]

**Proof.** A detailed proof can be found in [Mig]. It contains many technical improvements when compared with the result proved in [BMS], but the main progress is a zero-lemma due to Laurent [Lau03], which improves [Gou02] and provides an important improvement on the zero-lemma used in our previous paper [BMS].

### 14.2 How to use Theorem 3

To apply the theorem, we consider an integer $L \geq 5$ and real parameters $m > 0$, $\rho > 2$ (then one can define the $a_i$) and we put
\[ K = [mLa_1a_2a_3], \quad \text{with} \quad ma_1a_2a_3 \geq 2. \]

To simplify the presentation, we also assume $m \geq 1$ and $a_1, a_2, a_3 \geq 1$, and put
\[ R_1 = [c_1a_2a_3], \quad S_1 = [c_1a_1a_3], \quad T_1 = [c_1a_1a_2], \]
\[ R_2 = [c_2a_2a_3], \quad S_2 = [c_2a_1a_3], \quad T_2 = [c_2a_1a_2], \]
\[ R_3 = [c_3a_2a_3], \quad S_3 = [c_3a_1a_3], \quad T_3 = [c_3a_1a_2], \]

where the $c_i$ satisfy the conditions (i)–(v) of the theorem.

Clearly, condition (i) is satisfied if
\[ (c_1^2(a_1a_2a_3)^2)^{1/2} \geq \chi ma_1a_2a_3L, \quad c_1^2 \cdot a \geq 2mL, \quad \text{where} \ a = \min\{a_1, a_2, a_3\}. \]

Condition (ii) is true when $2c_1^2a_1a_2a_3 \cdot \min\{a_1, a_2, a_3\} \geq L$; we can take
\[ c_1 = \max\{(\chi mL)^{2/3}, (2mL/a)^{1/2}\}. \]

To satisfy (iii) and (iv) we can take
\[ c_2 = \max\{2^{1/3}(mL)^{2/3}, \sqrt{m/a}L\}. \]

Finally, because of the hypothesis (M), condition (v) holds for
\[ c_3 = (6m^{-2})^{1/3}L. \]

**Remark.** When $\alpha_1$, $\alpha_2$, $\alpha_3$ are multiplicatively independent then it is enough to take $c_1$ and $c_3$ as above and $c_2 = 2^{1/3}(mL)^{2/3}$.

Then we have to verify the condition (19). When this inequality holds, one obtains the lower bound $|\Lambda| > \rho^{-K}L$ and we get
\[ \log |\Lambda| > -KL \log \rho - \log(\max\{R, S, T\} \cdot L), \]
except maybe if at least one of the conditions (C1), (C2) or (C3) holds.
15. Completion of the proof of Theorem 1

Having given our new bounds for linear forms in three logarithms we now use them to complete the proof of Theorem 1. We have indeed shown in Lemma 11.1 that if \((x, y, p)\) is a missing solution then \(p > p_0\) where \(p_0\) is given in Table 4. To complete the proof it is enough to show that \(p \leq p_0\).

In § 13 we wrote down the linear form in logarithms we obtain for each outstanding value of \(D\). We will content ourselves by giving the details of this calculation for \(D = 7\). The other cases are practically identical (but with different constants).

We have seen in Lemma 13.3 that
\[
\Lambda := \log \frac{x - \sqrt{-7}}{x + \sqrt{-7}} \quad \text{satisfies} \quad \log |\Lambda| \leq -\frac{p}{2} \log y + \log(2.2\sqrt{7}).
\]

Writing \(a_0 = (1 + \sqrt{-7})/2\) we saw that the linear form is given by
\[
\Lambda = 2 \log(\varepsilon_1 \bar{\alpha}/\alpha_0) + p \log(\varepsilon_2 \bar{\gamma}/\gamma) + iq\pi, \quad \varepsilon_1, \varepsilon_2 = \pm 1,
\]
for some rational integer \(q\) with \(|q| < p\) and we get
\[
\log |\Lambda| > -KL \log \rho - \log(\max\{R, S, T\} \cdot L),
\]
except maybe if at least one of the conditions (C1), (C2) or (C3) holds.

Now we proceed effectively to the computation of an upper bound for \(p\). The first step is to recall that we have proved in Lemma 13.4, by applying Matveev’s theorem (Theorem 2), that
\[
p < 6.81 \times 10^{12}.
\]

We then apply our Theorem 3 with the initial condition \(p < 6.81 \times 10^{12}\) and with the lower bound \(y \geq 22\); note that we do not yet assume our lower bound (14) obtained through the modular approach. There are two reasons for this.

- The first reason is that we would like to demonstrate how powerful our new lower bound for linear forms in three logarithms is, even without the help of the modular approach.
- The second reason is that when we later make the assumption (14), and apply our lower bound for linear forms in three logarithms, the reader will appreciate the saving brought by the ‘modular lower bound’ for \(y\).

So for now we assume simply that \(y \geq 22\), which holds because \(y\) is even, not a power of 2 and that \(-7\) is a quadratic residue for every odd prime factor of \(y\) (see [Les98]). In a few steps we can prove that
\[
p < 4.2 \times 10^8.
\]

The reader should compare this bound with the bound \(p < 6.81 \times 10^{12}\) obtained by Matveev’s theorem.

We now assume our ‘modular’ lower bound for \(y\) in (14), with \(p > 1.3 \times 10^8\), and then we shall obtain a much better bound for \(p\). We give much more detail.

We have to distinguish two cases
\[
b_1 = 2, \quad \alpha_1 = \varepsilon' \bar{\alpha}/\alpha, \quad b_2 = p, \quad \alpha_2 = \varepsilon \bar{\gamma}/\gamma, \quad b_3 = q, \quad \alpha_3 = -1, \quad (I)
\]
and
\[
b_1 = 2, \quad \alpha_1 = \varepsilon' \bar{\alpha}/\alpha, \quad b_2 = q, \quad \alpha_2 = -1, \quad b_3 = p, \quad \alpha_3 = \varepsilon \bar{\gamma}/\gamma. \quad (II)
\]

Let us first consider case (I). Applying Theorem 3 we get
\[
p < 4.3 \times 10^8
\]
with the choices \( L = 110, \rho = 6, m = 71.602\,265.32, \chi = 0.7 \) and
\[
R_1 = 178\,896, \quad S_1 = 29\,587, \quad T_1 = 47\,734, \quad R_2 = 285\,899, \quad S_2 = 47\,284, \quad T_2 = 76\,285
\]
and \( R_3 = 1\,975\,684, \quad S_3 = 326\,756, \quad T_3 = 527\,164, \)
unless at least one of the conditions (C1), (C2), (C3) holds. For these values, it is clear that, because since we know that \( p > 10^8 \), conditions (C1) and (C2) do not hold.\(^1\) The values of \( B_R \) and \( B_T \) defined in Theorem 3 are equal to
\[
B_R = 127, \quad B_T = 483.
\]
The first case of condition (C3) implies that \( p < B_R = 127 \), a contradiction. Thus, we have to consider the last alternative:
\[
(t_1b_1 + r_1b_3)s_1 = r_1t_2b_2.
\]
Putting \( r_1 = \delta r'_1 \) and \( s_1 = \delta s'_1 \) and simplifying, we get here
\[
(2t_1 + \delta r'_1 q)s'_1 = r'_1 t_2 p.
\]
which shows that \( s'_1 = 1 \) and \( r'_1 = 1 \) or \( 2 \), and gives
\[
(2/r'_1)t_1 + \delta q = t_2 p, \quad \text{with } |t_2| \leq B_T/2.
\]
We write \( \delta \Lambda = \delta b_1 \log \alpha_1 + \delta b_2 \log \alpha_2 + (t_2b_2 - (b_1/r'_1)t_1) \log \alpha_3 \), that is
\[
\delta \Lambda = \frac{2}{r_1} \log(\alpha_1^{r_1} / \alpha_3^{t_1}) + p \log(\alpha_2^{t_2} / \alpha_3^{s_1}) = \text{a linear form in two logarithms}
\]
and we apply \([LM95]\) (with \( L = 10 \) and \( \rho = 16 \)), which now gives \( p < 3 \times 10^8 \).

Thus, we have proved that, in case I,
\[
p < 4.3 \times 10^8.
\]

Concerning case (II), we first note that in the non-degenerate case we obtain \( p < 4.3 \times 10^8 \) as before. Then we note that (C1) or (C2) implies that \( p \leq \max\{T_1, T_2\} \), (the present \( T_i \) play the role of the previous \( S_i \), and both are bounded independently of \( y \).) Now we study condition (C3). For the first alternative \( r_0b_2 = s_0b_1 \), we get \( |q| < B_R \) and we can apply \([LM95]\) to the linear form in two logarithms
\[
\Lambda = (\log(\varepsilon' \alpha / \alpha)^2 + q \log(-1)) + p \log(\varepsilon \bar{\gamma} / \gamma),
\]
which works quite well. Consider now the second alternative, which gives here (we have \( t_2 \neq 0 \); if \( t_2 = 0 \), then \( p < 10^8 \))
\[
(2/r'_1)t_1 + \delta p = t_2 q', \quad \text{with } |s_1| \leq B_S/2 \quad \text{and } q = s'_1 q', \quad r'_1 = 1 \text{ or } 2.
\]
We write now \( t_2 \Lambda = t_2b_1 \log \alpha_1 + t_2b_2 \log \alpha_2 + (s_1b_3 + (b_1/r'_1)s'_1 t_1) \log \alpha_3 \), that is
\[
t_2 \Lambda = \frac{2}{r_1} \log(\alpha_1^{r_1} / \alpha_3^{t_1}) + p \log(\alpha_2^{t_2} / \alpha_3^{s_1})
\]
and we apply \([LM95]\) (again with \( L = 10 \) and \( \rho = 16 \)), which now gives \( p < 2 \times 10^8 \).

Thus, we have proved that, in all cases,
\[
p < 4.3 \times 10^8.
\]

\(^1\)To be more precise we can take the above values for \( S_1, S_2 \) and \( S_3 \) independently of \( y \) but the \( R_i \) and \( T_i \) have to be increased for \( y > 1.3 \times 10^8 \), as can be seen on the definition of the parameters given in the previous section (\( a_1 \) and \( a_2 \) are independent of \( y \) but not \( a_3 \)). Luckily, the larger \( y \) is, the better our resulting estimate for \( p \) will be and thus we can always replace \( y \) by some lower bound for it.
Iterating this process four times we obtain that, in all cases

\[ p < 1.3 \times 10^8, \]

which is indeed better than the upper bound used in the modular computation.

**Remark.** We note that without the modular lower bound for \( y \) we were able to show that \( p < 1.11 \times 10^9 \), but with this modular lower bound we were able to improve this to \( p < 1.81 \times 10^8 \). Whilst it is certainly possible to reach the former target with the methods of this paper, it would have taken about six times as long as it took to reach the latter. From this it is a plausible guess that without the modular lower bound for \( y \) the computational part for the entire proof for all the values of \( 1 \leq D \leq 100 \) might have taken at least 800 days rather than 206 days.

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**Appendix A**

We list all the solutions to (1) in the range \( 1 \leq D \leq 100 \).

| \( D \) | Solutions \( (|x|, |y|, n) \) |
|---|---|
| 1 | \((0, 1, n)\) |
| 2 | \((5, 3, 3)\) |
| 3 | \((2, 2, 3), (11, 5, 3)\) |
| 4 | \((1, 2, 3), (181, 32, 3), (3, 2, 4), (5, 2, 5), (181, 8, 5), (11, 2, 7), (181, 2, 15)\) |
| 5 | \((0, 2, 3)\) |
| 6 | \((4, 3, 3), (58, 15, 3)\) |
| 7 | \((2, 2, 4)\) |
| 8 | \((70, 17, 3)\) |
| 9 | \((7, 4, 3), (1, 2, 4), (7, 2, 6)\) |
| 10 | \((0, 2, 4), (4, 2, 5)\) |
| 11 | \((8, 3, 4)\) |
| 12 | \((3, 3, 3), (15, 3, 5)\) |
| 13 | \((18, 7, 3), (22434, 55, 5)\) |
| 14 | \((14, 6, 3)\) |
| 15 | \((2, 3, 3), (3, 2, 5), (45, 2, 11)\) |
| 16 | \((10, 5, 3)\) |
| 17 | \((1, 3, 3), (207, 35, 3)\) |
| 18 | \((0, 3, 3)\) |
| 19 | \((6, 4, 3), (22, 8, 3), (225, 37, 3), (2, 2, 5), (6, 2, 6), (10, 2, 7), (22, 2, 9), (362, 2, 17)\) |
| 20 | \((23, 3, 3), (3, 2, 5), (45, 2, 11)\) |
| 21 | \((10, 5, 3)\) |
| 22 | \((1, 3, 3), (207, 35, 3)\) |
| 23 | \((0, 3, 3)\) |
| 24 | \((6, 4, 3), (22, 8, 3), (225, 37, 3), (2, 2, 5), (6, 2, 6), (10, 2, 7), (22, 2, 9), (362, 2, 17)\) |
### Classical and modular approaches

| $D$ | Solutions $(|x|, |y|, n)$ |
|-----|--------------------------|
| 31  | (15, 4, 4), (1, 2, 5), (15, 2, 8) |
| 32  | (7, 3, 4), (0, 2, 5), (88, 6, 5) |
| 33  |  |
| 34  |  |
| 35  | (36, 11, 3) |
| 36  |  |
| 37  |  |
| 38  |  |
| 39  | (5, 4, 3), (31, 10, 3), (103, 22, 3), (5, 2, 6) |
| 40  | (52, 14, 3) |
| 41  |  |
| 42  |  |
| 43  |  |
| 44  | (9, 5, 3) |
| 45  | (96, 21, 3), (6, 3, 4) |
| 46  | (13, 6, 3), (41, 12, 3), (500, 63, 3), (14, 3, 5), (9, 2, 7) |
| 47  | (4, 4, 3), (148, 28, 3), (4, 2, 6) |
| 48  | (524, 65, 3), (24, 5, 4) |
| 49  |  |
| 50  |  |
| 51  |  |
| 52  |  |
| 53  | (26, 9, 3), (156, 29, 3), (26, 3, 6) |
| 54  | (17, 7, 3) |
| 55  | (3, 4, 3), (419, 56, 3), (3, 2, 6) |
| 56  | (76, 18, 3), (5, 3, 4) |
| 57  |  |
| 58  |  |
| 59  |  |
| 60  | (2, 4, 3), (1586, 136, 3), (14, 4, 4), (50354, 76, 5), (2, 2, 6), (14, 2, 8) |
| 61  | (8, 5, 3) |
| 62  |  |
| 63  | (1, 4, 3), (13537, 568, 3), (31, 4, 5), (1, 2, 6), (31, 2, 10) |
| 64  | (0, 4, 3), (0, 2, 6), (8, 2, 7) |
| 65  | (4, 3, 4) |
| 66  |  |
| 67  | (110, 23, 3) |
| 68  |  |
| 69  |  |
| 70  |  |
| 71  | (21, 8, 3), (35, 6, 4), (46, 3, 7), (21, 2, 9) |
| 72  | (12, 6, 3), (3, 3, 4) |
| 73  |  |
| 74  | (985, 99, 3), (13, 3, 5) |
| 75  |  |
| 76  | (7, 5, 3), (1015, 101, 3) |
| 77  | (2, 3, 4) |
| 78  |  |
| 79  | (89, 20, 3), (7, 2, 7) |
| 80  | (1, 3, 4) |
| 81  | (46, 13, 3), (0, 3, 4) |
| 82  |  |
| 83  | (140, 27, 3), (140, 3, 9) |


| $D$ | Solutions ($|x|, |y|, n$) |
|-----|---------------------|
| 84  |                     |
| 85  |                     |
| 86  |                     |
| 87  | (16, 7, 3), (13, 4, 4), (13, 2, 8) |
| 88  |                     |
| 89  | (6, 5, 3)          |
| 90  |                     |
| 91  |                     |
| 92  | (6, 2, 7), (90, 2, 13) |
| 93  |                     |
| 94  |                     |
| 95  | (11, 6, 3), (529, 6, 7) |
| 96  | (23, 5, 4)         |
| 97  | (48, 7, 4)         |
| 98  |                     |
| 99  | (12, 3, 5)         |
| 100 | (5, 5, 3), (30, 10, 3), (198, 34, 3), (55, 5, 5) |

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