

## Algorithms for diophantine equations

B.M.M. de Weger



## CWI Tracts

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## CWI Tract

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SOLI DEO GLORIA.

Benne de Weger, University of Twente,

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## Chapter 1. Introduction.

### 1.1. Algorithms for diophantine equations.

This monograph deals with certain types of diophantine equations. An equation is a mathematical formula, expressing equality of two expressions that involve one or more unknowns (variables). Solving an equation means finding all solutions, i.e. the values that can be substituted for the unknowns such that the equation becomes a true statement. An equation is called a diophantine equation if the solutions are restricted to be integers in some sense, usually the ordinary rational integers (elements of $\mathbb{Z}$ ) or some subset of that.

Examples of diophantine equations that will be studied in this book are

$$
x^{2}+7=2^{n}
$$

(the Ramanujan-Nagell equation, having only the solutions given by $( \pm x, n)=(1,3),(3,4),(5,5),(11,7),(181,15)$, see Chapter 4);

$$
2^{x}=3^{y}+5^{z}
$$

(a purely exponential equation, having only the solutions $(x, y, z)=(1,0,0)$, $(2,1,0),(3,1,1),(5,3,1),(7,1,3)$, see Chapter 6$)$;

$$
y^{2}=x^{3}-4 \cdot x+1
$$

(an elliptic curve equation, having only 22 solutions, of which the largest are $(x, y)=(1274, \pm 45473)$, see Chapter 8$)$. The three examples mentioned here are only some examples; we will study much wider classes of equations. We also study (in Chapter 5) a diophantine inequality (a formula expressing that one expression is larger than another, where solutions are again restricted to integers). In the following discussion the statements about diophantine equations also hold for this inequality.

What the equations treated in this book have in common is that they can all be solved by the same method. This method consists essentially of three
parts: a transformation step, an application of the Gelfond-Baker theory, and a diophantine approximation step. We explain these steps briefly.

To start with, one transforms the equation into a purely exponential equation or inequality, i.e. a diophantine equation or inequality where the unknowns are all in the exponents, such as in the second example given above. Each type of diophantine equation needs a particular kind of transformation, so that it is difficult to be more specific at this point. In some instances, such as in the second example above, this transformation is easy, if not trivial. In other instances, as in the first example above, it uses some arguments from algebraic number theory, or, as in the third example above, a lot of them.

In general, such a purely exponential equation has the form

$$
\begin{equation*}
\sum_{i=1}^{t} c_{i} \cdot \prod_{j=1}^{s_{i} \alpha_{i j}}{ }_{i j}=c_{0} \cdot \prod_{j=1}^{s_{0} \alpha_{0 j}}{ }_{0 j} \tag{1.1}
\end{equation*}
$$

and a corresponding purely exponential inequality looks like

$$
\begin{equation*}
\left|\sum_{i=1}^{t} c_{i} \cdot \prod_{j=1}^{s_{i}} \alpha_{i j}^{n_{i j}}\right|<\underset{i}{\min }\left|c_{i} \cdot \prod_{j=1}^{s_{i} \alpha_{i j}}\right|^{\delta} \tag{1.2}
\end{equation*}
$$

where $t, s_{i}, c_{i}, \alpha_{i j}, \delta$ are constants with $t, s_{i} \in \mathbb{N}, 0<\delta<1$, and $c_{i}, \alpha_{i j}$ belong to some algebraic extension of $\mathbb{Q}$, and where the $n_{i j}$ are the unknowns in $\mathbb{Z}$. We now suppose that the number of terms $t$ on the left hand side of (1.1) or (1.2) is equal to 2 . This restriction is essential for the second step, in which we use results from the so-called theory of linear forms in logarithms, also known as the Gelfond-Baker theory. (Some special exponential equations of type (1.1) with $t>2$ can also be treated by the Gelfond-Baker method, since they can be reduced to exponential inequalities of type (1.2) with $t=2$, cf. Stroeker and Tijdeman [1982], Alex [1985 ${ }^{\mathrm{a}}$ ], [1985 ${ }^{\mathrm{b}}$ ], Tijdeman and Wang [1988].)

An exponential equation or inequality such as (1.1) or (1.2) with $t=2$ gives rise to a linear form in logarithms

$$
\Lambda=\log \beta_{0}+\sum_{i=1}^{m} n_{i} \cdot \log \beta_{i}
$$

where the $\beta_{i}$ are algebraic constants, and the $n_{i}$ are integral unknowns. Here, the logarithms are real or complex in some instances, or p-adic in
other cases. This relation between equation and linear form in logarithms is such that for a large solution of the equation the linear form is extremely close to zero (in the real or complex sense, or in the p-adic sense). The Gelfond-Baker theory provides effectively computable lower bounds for the absolute values (respectively p-adic values) of such linear forms in logarithms of algebraic numbers. In many cases these bounds have been explicitly computed. Comparing the so-found upper and lower bounds it is possible to obtain explicit upper bounds for the solutions of the exponential diophantine equation or inequality, leading to upper bounds for the solutions of the original equation. This second step, unlike the first (transformation) step, is of a rather general nature.

We remark that many authors have given effectively computable upper bounds for the solutions of a wide variety of diophantine equations, by applying the method sketched above. For a survey, see Shorey and Tijdeman [1986]. Often these authors were satisfied with the knowledge of the existence of such bounds, and they did not actually compute them. If they computed bounds, they did not always determine all the solutions. In this book, solving an equation will always mean: explicitly finding all the solutions.

After the second step, the problem of solving the diophantine equation is reduced to a finite problem, which is treated in the third part of the method. Namely, since we have found explicit upper bounds for the absolute values of the (integral) unknowns, we have to check only finitely many possibilities for the unknowns. However, the word finite does not mean the same as small or trivial. In fact, the constants appearing in the lower bounds that the Gelfond-Baker theory provides for linear forms in logarithms are rather large. Therefore, in practice the upper bounds that can be obtained in this way for the solutions of purely exponential equations can be for instance as large as $10^{40}$. This is far too large to admit simple enumeration of all the possibilities, even with the fastest of computers today.

Proving the existence of an absolute upper bound for the solutions reduces the determination of all the solutions from an infinite task to a finite one. Thus, the application of the Gelfond-Baker theory (the second step) is in a sense infinitely many times as difficult a task than the only finite amount of checking that remains to be done (in the third step). Furthermore, this checking seems to be a technical problem only, not a mathematical one.

Nevertheless, it is the author's opinion that solving this comparatively small technical problem is not only nontrivial, but involves some serious and interesting mathematics. This book hopefully illustrates this opinion.

Notwithstanding the fact that the application of the Gelfond-Baker theory in the second step yields very large upper bounds, it is generally assumed that these upper bounds are far from the actual largest solution. Therefore, it is worthwile to search for methods to reduce these upper bounds to a size that can be more easily handled. Often it is possible to devise such a method using directly certain properties of the original diophantine equation, for example that large solutions must satisfy certain congruences modulo many or large numbers (Grinstead [1978], Brown [1985], Pinch [1988]), or some reciprocity condition (Pethö [1983]). The disadvantage of such methods is that they work only for that particular type of diophantine equation, so that in general for each type of equation a new reduction method must be devised. It would therefore be interesting to have methods for reducing upper bounds for the solutions of inequalities for linear forms in logarithms. They would be useful for solving any type of diophantine problem that leads to such inequalities.

Such methods are searched for in the third step of our method of solving diophantine equations. It is mainly in this third part that new developments can be reported. The arguments we use in the first and second parts are mainly classical, and we apply them to types of equations that have been studied before, and also to new types of equations.

The methods that are needed in the third step are provided by that part of the theory of diophantine approximation that is concerned with studying how close to zero a linear form can be for given values of the variables. Recently important progress has been made in this field, the breakthrough being the invention in 1981 by L. Lovász of the so-called ${ }^{3}$-laticce basis reduction algorithm. We will show how this $L^{3}$-algorithm leads to practically efficient diophantine approximation algorithms, which can be employed for many diophantine equations to show that in a certain interval $\left[\mathrm{X}_{1}, \mathrm{X}_{0}\right.$ ] no solutions exist. Usually $X_{1}$ is of the order of magnitude of $\log X_{0}$. When for $X_{0}$ the theoretical upper bound for the solutions is substituted, a new, and usually much better upper bound $X_{1}$ is found. For many equations the initial upper bound $X_{0}$ is well within reach of practical application of these algorithms, within only a few minutes of computer time. This thus leads
in practice to methods for finding all the solutions of many types of diophantine equations, for which alternative methods have not yet been found or employed with success.

The method outlined above, and used in this book to solve many examples of various diophantine equations, is of an "algorithmic" nature. In a sense it lies between "ad hoc" methods and "theoretical" methods. This we shall explain below. Let a set of diophantine equations with an unspecified parameter in it be given. As an example of such a set, consider the generalized Ramanujan-Nagell equation $x^{2}+D=2^{n}$, where $D$ is a parameter, and $x, n$ are the unknowns.

An ad hoc method is a method for solving the equation for specific values of the parameters only. It may not work at all for other than these particular values. The first example of solving an equation of the type $x^{2}+D=2^{n}$ occurring in the literature is that by Nagell [1948] of $D=7$. The method he used is of an ad hoc nature, since it depends heavily on the special choice of 7 for the parameter D.

A theoretical method is capable of proving results that hold for some large set of values of the parameters. The Gelfond-Baker theory is of a theoretical nature, since it yields upper bounds for the solutions of many equations in terms of their parameters. Other examples are application of the theory of quadratic reciprocity, that shows that $x^{2}+D=2^{n}$ has no solutions at all if $D$ is odd, at least 5 , and not congruent to 7 (mod 8), and application of the theory of hypergeometric functions, which Beukers [1981] used to show that the solutions $(x, n)$ of $x^{2}+D=2^{n}$ satisfy $n<435+10 \cdot{ }^{2} \log |D|$, and if $|D|<2^{96}$ then moreover $n<18+2 \cdot{ }^{2} \log |D|$. Theoretical methods are of ten too general to be able to produce all the solutions of a given equation.

An algorithmic method is a method that is guaranteed to work for any set of values of the parameters, but has to be applied separately to each particular set of parameter values, in order to produce all the solutions. The methods used in this book are mainly of such an algorithmic nature. For the equation $x^{2}+D=2^{n}$ (and actually for a more general equation) we will give an algorithmic method in Chapter 4. In fact, since Beukers' above-mentioned result provides a small upper bound for the solutions, it can be made algorithmic by providing a simple method of enumerating all the solutions
below the upper bound. However, the algorithmic part of this method is trivial, and therefore we still prefer to classify Beukers' method as theoretical. In order to make the Gelfond-Baker theory algorithmic, enumeration of all possibilities is impractical. Therefore more ingenious ways of determining all the solutions below a large upper bound have to be found. We remark that Beukers' method for the more general equation $x^{2}+D=p^{n}$ also has an ad hoc aspect, since it works for some special values of $p$ only. Our method of Chapter 4 does not have this disadvantage.

An ideal towards which one might strive in solving diophantine equations is to devise a computer algorithm, a kind of 'diophantine machine', which only has to be fed with the parameters of the equation, and after a short time gives as output a list of all the solutions. One should have a guarantee (in the strictest mathematical sense of proof) that no solutions are missing.

At first sight the method outlined above, and described in this monograph, seems to be a good candidate to be developed into such a general applicable algorithm. Namely, the second step is of a quite general nature, providing upper bounds for exponential diophantine equations that are explicit in the parameters of the equation. Also the third step, the algorithmic diophantine approximation part, works in principle for any set of values substituted for the parameters. However, the computations have to be performed separately for each particular set of values.

The main difficulties in devising such a diophantine machine' are in the first part of the method outlined above, especially if some algebraic number theory is used. Developments taking place in the theory of algorithmic algebraic number theory on computing fundamental units and on finding factorizations of prime numbers in algebraic extensions, are of importance here. We believe that when suitable algorithms of this kind are available, it will be possible in principle to make such a 'diophantine machine' (but technical difficulties in the third step should not be underestimated). The generality of such an algorithm is restricted by the generality of the first step, the transformation to the linear form in logarithms. In this book we use computer algorithms only if the magnitude of the computational tasks makes this necessary, and keep to "manual" work otherwise. In this way we also try to keep the presentation of the methods lucid.

The reader should be aware of the fact that the computer programs and their
results are part of the proofs of many of our theorems on specific diophantine equations. It is however impossible to publish all details of these programs and computations. The interested reader may obtain the details from the author by request, and is invited to check the computations himself.

The book by Shorey and Tijdeman [1986] gives a good survey of the diophantine equations for which computable upper bounds for the solutions can be found using the Gelfond-Baker method (see also Shorey, van der Poorten, Tijdeman and Schinzel [1977], and Stroeker and Tijdeman [1982]). Some of these equations can be completely solved by the methods described in this book, among which there are purely exponential equations, equations involving binary recurrence sequences, and Thue equations and Thue-Mahler equations. Especially the latter two are of importance in various other parts of number theory. For example, they are the key to solving Mordell equations and various equations arising in algebraic number theory and arithmetic algebraic geometry. The Gelfond-Baker method was used to actually solve a diophantine equation for the first time in the work of Baker and Davenport [1969] in solving the system of diophantine equations

$$
3 \cdot x^{2}-2=y^{2}, \quad 8 \cdot x^{2}-7=z^{2}
$$

Other equations occuring in the literature for which upper bounds for the solutions can be computed, cannot be treated as easily by our algorithmic methods, because the application of the theory of linear forms in logarithms is more complicated for these equations, and moreover the upper bounds are essentially too large. An example of this kind is the Catalan equation $a^{x}-b^{y}=1$ in integers $a, b, x, y, a l l \geqslant 2$. Catalan conjectured in 1844 that this equation has only the solution $(a, b, x, y)=(3,2,2,3)$. Tijdeman [1976] proved that the solutions of the Catalan equation are bounded by a computable number. This number can be taken to be $\exp (\exp (\exp (\exp (730))))$, according to Langevin [1976]. However, we fail to see how the methods that we describe in the forthcoming chapters can be applied for completely solving the Catalan equation, and we believe that Grosswald's remarks on this topic are too optimistic (Grosswald [1984], p. 259, in particular the footnote).

Another diophantine equation, that for centuries has attracted the attention of many mathematicians, is the Fermat equation $x^{n}+y^{n}=z^{n}$ in integers $x$, $y, z, n$, with $n \geqslant 3$ and $x \cdot y \cdot z \neq 0$. It is conjectured to have no solutions. Faltings [1983] proved that for fixed $n$ the number of solutions
is finite. His proof is ineffective. The Gelfond-Baker theory seems not to be strong enough to deal with the Fermat equation in its full generality, not even if $n$ is fixed. For a survey of partial results on the Fermat equation that have been obtained using this theory, see Tijdeman [1985] and Chapter 11 of Shorey and Tijdeman [1986].

We remark that for many diophantine equations recently important progress has been made in determining upper bounds for the number of solutions. See e.g. Evertse [1983], Evertse, Györy, Stewart and Tijdeman [1988] and Schmidt [1988] for a survey. These results are often remarkably sharp, but ineffective, so that they cannot be used for actually finding the solutions.

To conclude this section we give an overview of the contents of this monograph. It is divided into three parts: Chapter 1 is introductory, Chapters 2 and 3 give the necessary preliminaries, and Chapters 4 to 8 deal with various types of diophantine equations.

Sections 1.2 to 1.5 give a short introduction for the non-specialist to respectively the Gelfond-Baker theory, diophantine approximation theory, the algorithmic aspects of diophantine approximation, and the procedure for reducing upper bounds. Chapter 2 contains the preliminary results that we need from algebraic number theory and from the theory of $p$-adic numbers and functions, and quotes in full detail the theorems from the Gelfond-Baker theory which we use. It concludes with some remarks on numerical methods. Chapter 3 gives in detail the algorithms in the field of diophantine approximation theory that we apply in the subsequent chapters. In a sense this chapter is the heart of the book.

Chapters 4 to 8 are each devoted to a certain type of diophantine equation. Let $p_{1}, \ldots, p_{s}$ be a fixed set of distinct primes. Let $S$ be the set of positive integers composed of primes $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{s}}$ only.

Chapter 4 deals with elements of binary recurrence sequences ("generalized Fibonacci sequences") that are in $S$, and gives applications to mixed quadratic-exponential equations, such as the generalized Ramanujan-Nagell equation $x^{2}+k \in S \quad(k$ fixed). The diophantine approximation part of this chapter is interesting for two reasons: the p-adic approximation is very simple, and in the case of the recurrence having negative discriminant, a nice interplay of p-adic and real/complex approximation arguments occurs. The
research for Chapter 4 was done partly in cooperation with A. Pethö from Debrecen. The results have been published in Pethö and de Weger [1986] and de Weger $\left[1986{ }^{\mathrm{b}}\right.$ ].

Chapter 5 deals with the diophantine inequality $0<x-y<y^{\delta}$, where $x, y \in S$, and $\delta \in(0,1)$ is fixed. Chapter 6 deals with $x+y=z$, where $x, y, z \in S$, which can be considered as the p-adic analogue of the inequality of Chapter 5. These two equations are the simplest examples of diophantine equations that can be treated by our method. Since they are already purely exponential equations of the form (1.1) or (1.2) with $t=2$, the first step is trivial: the linear forms in logarithms are directly related to the equations. Therefore they serve as good examples to get a clear understanding of the diophantine approximation part of our method. The results of these chapters have been published in de Weger [1987].

Chapter 7 studies the equation $x+y=z^{2}$, where $x, y \in S$, and $z \in \mathbb{Z}$. This equation is a further generalization of the generalized Ramanujan-Nagell equation, studied in Chapter 4.

In Chapter 8 a procedure is given to solve Thue equations, that works in principle for Thue equations of any degree. It is applied to find all integral points on the elliptic curve $y^{2}=x^{3}-4 \cdot x+1$. We also mention briefly how Thue-Mahler equations can be dealt with. This chapter has been written jointly with $N$. Tzanakis from Iraklion. The results have been published in Tzanakis and de Weger [1989 ${ }^{\text {a }}$, and in de Weger [1989 ${ }^{\text {a }}$ ].

### 1.2. The Gelfond-Baker method.

In Section 1.1 we have explained that before applying the Gelfond-Baker method to some diophantine equation, the equation should be transformed into a purely exponential diophantine equation or inequality with not too many terms (cf. (1.1), (1.2)). In this section we sketch the arguments from the Gelfond-Baker theory that lead to upper bounds for the variables of this exponential equation/inequality.

Let us first treat the case of the inequality (1.2). Since $t=2$ we may assume that it has the form

$$
\left|\alpha_{0} \cdot \prod_{i=1}^{s} \alpha_{i}^{n_{i}}-1\right|<C_{0} \cdot \exp (-\delta \cdot N)
$$

where the $\alpha_{i}$ are fixed algebraic numbers, $N=\max \left|n_{i}\right|$, and $C_{0}, \delta$ are positive constants. In the examples we study, we encounter one of the following two cases: either all $\alpha_{i}$ are real, or $\left|\alpha_{i}\right|=1$ for all $i$. In the real case, if $N$ is large enough, the linear form in logarithms

$$
\Lambda=\log \left|\alpha_{0}\right|+\sum_{i=1}^{S} n_{i} \cdot \log \left|\alpha_{i}\right|
$$

must satisfy

$$
\begin{equation*}
|\Lambda|<C_{0}^{\prime} \cdot \exp (-\delta \cdot N) \tag{1.3}
\end{equation*}
$$

for some $\mathrm{C}_{0}^{\prime}$. In the complex case, the same inequality (1.3) follows for the linear form

$$
\begin{aligned}
& \Lambda=\log \alpha_{0}+\sum_{i=1}^{S} n_{i} \cdot \log \alpha_{i}+k \cdot \log (-1) \\
& \quad=i \cdot\left(\operatorname{Arg} \alpha_{0}+\sum_{i=1}^{S} n_{i} \cdot \operatorname{Arg} \alpha_{i}+k \cdot \pi\right)
\end{aligned}
$$

where the Log and Arg functions take their principal values. Now we can apply one of the many results from the Gelfond-Baker theory, giving an $\operatorname{explicit}$ lower bound for $|\Lambda|$ in terms of $N$, e.g. the following theorem.

THEOREM 1.1. (Baker [1972]). Let $\Lambda$ be as above. There exist computable constants $C_{1}, C_{2}$, depending on the $\alpha_{i}$ only, such that if $\Lambda \neq 0$ then

$$
|\Lambda|>\exp \left(-\left(C_{1}+C_{2} \cdot \log N\right)\right)
$$

We usually know that $\Lambda \neq 0$. Combining (1.3) and Theorem 1.1 we then obtain

$$
N<\frac{C_{1}+\log C_{0}^{\prime}}{\delta}+\frac{C_{2}}{\delta} \cdot \log N
$$

It follows that N is bounded from above.

Next, consider the exponential equation (1.1). By $t=2$ we can write it as

$$
\alpha_{0} \cdot \prod_{i=1}^{s} \alpha_{i}^{n}-1=\beta_{0} \cdot \prod_{j=1}^{r} \beta_{j}^{m}
$$

where the $\alpha_{i}, \beta_{j}$ are fixed algebraic numbers. Let $H_{p}$ be the maximum of the $\left|n_{i}\right|,\left|m_{j}\right|$ where $i, j$ run through the set of indices for which $\alpha_{i}$ resp. $\beta_{j}$ are non-units. Let $H$ be the maximum of the $\left|n_{i}\right|,\left|m_{j}\right|$ where i, $j$ run through the set of all indices. Suppose that $p$ is a rational prime lying above $\beta_{j}$ for some $j$. There are constants $c_{1}, c_{2}$ such that

$$
\operatorname{ord}_{p}\left(\alpha_{0} \cdot \prod_{i=1}^{s} \alpha_{i}^{n_{i}}-1\right) \geqslant c_{1}+c_{2} \cdot m_{j}
$$

Assuming that $\operatorname{ord}_{p}\left(\alpha_{i}\right)=0$ for all $i$, we may write down a $p$-adic linear form in logarithms

$$
\Lambda=\log _{p} \alpha_{0}+\sum_{i=1}^{S} n_{i} \cdot \log _{p} \alpha_{i}
$$

for which, if $m_{j}$ is large enough, it follows that

$$
\begin{equation*}
\operatorname{ord}_{p}(\Lambda) \geqslant c_{1}+c_{2} \cdot m_{j} \tag{1.4}
\end{equation*}
$$

We are now in a position to apply the following result from the p-adic Gelfond-Baker theory. Here, $N=\max \left|n_{i}\right|$.

THEOREM 1.2. (van der Poorten [1977], Yu [1987]). Let $\Lambda$, p be as above. There exist computable constants $\mathrm{C}_{3}, \mathrm{C}_{4}$, depending only on the $\alpha_{i}$ and on $p$, such that if $\Lambda \neq 0$ then

$$
\operatorname{ord}_{p}(\Lambda)<C_{3}+C_{4} \cdot \log N
$$

Applying (1.4) and Theorem 1.2 for all possible $p$ we obtain constants $C_{3}$, $C_{4}$ with

$$
\mathrm{H}_{\mathrm{p}}<\mathrm{C}_{3}^{\prime}+\mathrm{C}_{4}^{\prime} \cdot \log \mathrm{H}
$$

If $H \leqslant C_{5} \cdot H_{p}$ for some constant $C_{5}$, then this immediately yields an upper bound for $H$. If $H>C_{5} \cdot H_{p}$, then it can be shown that there exists a conjugate of the $\alpha_{i}, \beta_{j}$, denoted with a prime sign, for which

$$
\left|\beta_{0}^{\prime} \cdot \prod_{j=1}^{r} \beta_{j}^{m}{ }^{m}\right|<\exp \left(-C_{6} \cdot H\right)
$$

for a constant $\mathrm{C}_{6}$ (cf. the proof of Theorem 1.4, pp. 45-49, of Shorey and Tijdeman [1986]). Now we can apply Theorem 1.1. This yields

$$
\left|\alpha_{0}^{\prime} \cdot \prod_{i=1}^{s} \alpha_{i}^{n}-1\right|>\exp \left(-\left(C_{7}+C_{8} \cdot \log H\right)\right)
$$

It follows that $H$ is bounded from above.

If it happens that none of the $\alpha_{i}, \beta_{j}$ are units, then of course the application of Theorem 1.2 suffices.

We remark that, in order to be able to completely solve a diophantine equation, it is crucial that all constants can be computed explicitly. Therefore we can only use the bounds from the Gelfond-Baker theory that are completely explicit. We give details of such theorems in Section 2.4.

### 1.3. Theoretical diophantine approximation.

In this section we briefly mention some results from diophantine approximation theory, thus giving a background to the next section. We refer to Koksma [1937], Cassels [1957] (Chapters I and III) and to Hardy and Wright [1979] (Chapters XI and XXIII), for further details.

The simplest form of diophantine approximation in the real case is that of approximation of a real number $\vartheta$ by rational numbers $p / q$. It is well known that if $\vartheta$ is irrational, then there exist infinitely many solutions $(p, q) \in \mathbb{Z} \times \mathbb{N}$ with $(p, q)=1$ of the diophantine inequality

$$
\left|\vartheta-\frac{p}{q}\right|<q^{-2} .
$$

All convergents from the continued fraction expansion of $\vartheta$ are such solutions. The convergents are simple to compute for any particular $\vartheta \in \mathbb{R}$.

One way of generalizing this is to study simultaneous approximations to a set of real numbers $\vartheta_{1}, \ldots, \vartheta_{n}$, i.e. rational approximations to $\vartheta_{i}$ all having the same denominator. It is well known that the system of inequalities

$$
\left|\vartheta_{i}-\frac{p_{i}}{q}\right|<q^{-(1+1 / n)} \text { for } i=1, \ldots, n
$$

has infinitely many solutions $\left(p_{1}, \ldots, p_{n}, q\right)$ if at least one of the $\vartheta_{i}$ is irrational. But it is much harder to find solutions of such inequalities than in the case $n=1$. Some multi-dimensional continued fraction algorithms
have been devised (cf. Brentjes [1981] for a survey), but they seem not to have the desired simplicity and generality. We shall see later how we can apply the so-called $L^{3}$-algorithm to this problem.

Another way of generalizing the simplest case of diophantine approximation is to study linear forms, such as

$$
L=\sum_{j=1}^{m} q_{j} \cdot \vartheta{ }_{j},
$$

where $\vartheta_{1}, \ldots, \vartheta_{m}$ are given real numbers, and $q_{1}, \ldots, q_{m}$ are the unknowns in $\mathbb{Z}$. Put $Q=\max _{\mathrm{i}} \mathrm{l} \mid$. A classical theorem guarantees the existence of a solution $\left(p, q_{1}, \ldots, q_{m}\right)$ of the inequality

$$
|L-p|<Q^{-m}
$$

Note that the case $m=1$ becomes our first inequality on dividing by $q=q_{1}$. Also in this case the $L^{3}$-algorithm is very useful, as we shall see below.

We can incorporate the two generalizations above in a further generalization, that of simultaneous approximation of linear forms. Let real numbers $\vartheta_{i j}$ be given for $i=1, \ldots, n, j=1, \ldots, m$. Put

$$
L_{i}=\sum_{j=1}^{m} q_{j} \cdot \vartheta_{i j} \text { for } i=1, \ldots, n
$$

A celebrated theorem of Minkowski states that there exists a solution $\left(p_{1}, \ldots, p_{n}, q_{1}, \ldots, q_{m}\right)$ of the system of inequalities

$$
\left|L_{i}-p_{i}\right|<Q^{-m / n} \text { for } i=1, \ldots, n
$$

As we shall show in Section 1.4, the $\mathrm{L}^{3}$-algorithm may be applied to this general form. We actually compute solutions of systems of inequalities that are slightly weaker in the sense that the right hand side is multiplied by a small constant larger than 1.

We now consider inhomogeneous approximation. This means that for all i there is an inhomogeneous term $\beta_{i}$ in the linear form $L_{i}$, viz.

$$
L_{i}=\beta_{i}+\sum_{j=1}^{m} q_{j} \cdot \vartheta_{i j} \text { for } i=1, \ldots, n
$$

Again, there exists a constant $c$ such that the system

$$
\left|L_{i}-p_{i}\right|<c \cdot Q^{-m / n} \text { for } i=1, \ldots, n
$$

under some independence condition on the $\beta_{i}$ and $\vartheta_{i j}$, has a solution. This is Kronecker's theorem. The simplest case $m=n=1$ comes down to

$$
|q \cdot \vartheta-p+\beta|<c \cdot q^{-1}
$$

The upper bounds given above, that tell us that the order of magnitude of $\left|L_{i}-p_{i}\right|$ can be at least as small as $Q^{-m / n}$, are not only theoretical upper bounds, but they predict the heuristically expected order of magnitude as well. By this we mean that in a generic situation (i.e. when there are no almost-linear relations between the $\vartheta_{i j}$ (and the $\beta_{i}$ ), it is indeed the case that for a given $\mathrm{Q}_{0}$ the minimal $\max _{\mathrm{i}} \mathrm{L}_{\mathrm{i}}-\mathrm{p}_{\mathrm{i}} \mid$, taken over all $\mathrm{Q} \leqslant \mathrm{Q}_{0}$, has the order of magnitude of the upper bound $Q^{-m / n}$.

To conclude this section, we remark that there is a p-adic analogue of this theory of diophantine approximation, founded by Mahler and Lutz. If we replace in the above considerations $\mathbb{R}$ by $\mathbb{Q}_{p}$, the absolute value $|\cdot|$ by the $p$-adic value $I \cdot I_{p}$, and the measure $Q$ for an approximation $\left(p_{1}, \ldots, p_{n}, q_{1}, \ldots, q_{m}\right)$ by any convex norm $\Phi\left(p_{1}, \ldots, p_{n}, q_{1}, \ldots, q_{m}\right)$ on $\mathbb{R}^{n+m}$, then the p-adic analogues of the theorems of Minkowski and Kronecker are essentially analogous to the above mentioned results in the real case. See Koksma [1937] for references to Mahler's work, and Lutz [1951], and for a detailed analysis of the case $n=1, m=2$ see de Weger [1986 ${ }^{\text {a }}$ ].

### 1.4. Computational diophantine approximation.

In this section we give some idea of practically solving the diophantine approximation problems that we encounter in solving diophantine equations. In this section we give no rigorous treatment. We neglect worst cases, and concentrate on how things are expected to work (according to the heuristics of Section 1.3), and appear to work in practice. In the subsequent chapters many examples are given, showing that our methods are indeed useful in practice. Applying the method in practice may be the best way of acquiring the necessary Fingerspitzengefühl for the method.

We shall deal with the following computational diophantine approximation
problem. Let $\vartheta_{i j}, \beta_{i} \in \mathbb{R}$ be given, and let $p_{1}, \ldots, p_{n}, q_{1}, \ldots, q_{m}$ be integral unknowns with $Q=\max _{\mathrm{l}}^{\mathrm{j}} \mathrm{I}$. Let $\mathrm{L}_{\mathrm{i}}$ be as above. Let a positive constant $\mathrm{Q}_{0}$, assumed to be a rather large number, $10^{50}$ say, be given. Find a lower bound for the value of

$$
\max _{i}\left|L_{i}-p_{i}\right|
$$

where $\left(p_{1}, \ldots, p_{n}, q_{1}, \ldots, q_{m}\right)$ runs through the set of values with $Q \leqslant Q_{0}$. From the heuristics outlined in Section 1.3 it follows that one will be satisfied if this lower bound is of the size $Q_{0}^{-m / n}$. For the p-adic case an analogous problem may be formulated.

Related problems in diophantine approximation theory are those of actually finding a good or the best solution of $\max \left|L_{i}-p_{i}\right|<\varepsilon$ for a fixed $\varepsilon>0$. As we shall see, the $L^{3}$-algorithm is a very useful tool for finding good solutions. The problem of finding the best solution however seems to be essentially more difficult. We note that in most of our applications of solving diophantine equations it suffices to have a suitable lower bound for $\max \mid L_{i}-p_{i} l$ for a given $Q_{0}$, while it is unnecessary to know explicitly how sharp this bound is.

The computational tool that we use to solve the afore-mentioned problems is the so-called $L^{3}$-lattice basis reduction algorithm, described in Lenstra, Lenstra and Lovász [1982]. We shall give details of this algorithm in Sections 3.4 and 3.5. Below we briefly indicate how it can be used to solve diophantine approximation problems.

Let $\Gamma$ be a lattice in $\mathbb{R}^{n}$. The $L^{3}$-algorithm accepts as input an arbitrary basis $\underline{b}_{1}, \ldots, \underline{b}_{n}$ of $\Gamma$. As output it gives another basis $\underline{c}_{1}, \ldots, \underline{c}_{n}$ of the same lattice $\Gamma$, that is a so-called reduced basis. The concept reduced means something like nearly orthogonal. From a reduced basis it is possible to compute lower bounds for the following two quantities:
$\rightarrow$ the length of the non-zero lattice point that is nearest to the origin:

$$
\ell(\Gamma)=\min _{\underline{0} \neq \underline{x} \in \Gamma}|\underline{x}|,
$$

(see Lenstra, Lenstra and Lovász [1982], Prop. (1.11), and our Lemma 3.4),
$\rightarrow$ for any given point $y \in \mathbb{R}^{n}$, the distance from $y$ to the nearest lattice point:

$$
\ell(\Gamma, \underline{y})=\min _{\underline{x} \in \Gamma}|\underline{x}-\underline{y}|
$$

(see Babai [1986], and our Lemmas 3.5 and 3.6).

The $L^{3}$-algorithm enjoys the property that these lower bounds are usually near to the actual minimal solutions. In a generic situation, where the lattice is not too distorted, the vectors $c_{i}$ of the reduced basis all have about the same length, which is of the order of magnitude of

$$
\operatorname{det}(\Gamma)^{1 / n}
$$

The value of $\ell(\Gamma)$ as well as the lower bounds computed for it, are about as large as that. If $y$ is not too close to a lattice point, the same holds for $\ell(\Gamma, \underline{y})$. Moreover, the running time of the algorithm is good, both in the theoretical sense (it is polynomial-time in the length of the inputparameters), and in practice (cf. Lenstra [1984], p. 7).

To solve the problem of finding a lower bounds for $\max _{\mathrm{i}}\left|\mathrm{L}_{\mathrm{i}}-\mathrm{p}_{\mathrm{i}}\right|$ as formulated above, we take the lattice $\Gamma$ as follows. Let $C$ be an integer, at least as large as $Q_{0}^{1+m / n}$. The lattice $\Gamma$, of dimension $n+m$, is defined by specifying a basis, namely the column vectors $\underline{b}_{1}, \ldots, \underline{b}_{n+m}$ of the matrix

$$
\mathcal{B}=\left(\begin{array}{ccccccc}
1 & & & & & & \\
\varnothing & & \cdot & & & & \\
& & & & & \\
{\left[\mathrm{C} \cdot \vartheta_{11}\right]} & \ldots & {\left[\mathrm{C} \cdot \vartheta_{1 \mathrm{~m}}\right]} & -\mathrm{C} & & & \\
\vdots & & \vdots & & \varnothing & \cdot & \\
{\left[\mathrm{C} \cdot \vartheta_{\mathrm{n} 1}\right]} & \ldots & {\left[\mathrm{C} \cdot \vartheta_{\mathrm{nm}}\right]} & & & -\mathrm{C}
\end{array}\right) .
$$

(The symbol $\varnothing$ means that all not explicitly given entries in that area are zero). Applying the $L^{3}$-algorithm to this lattice we find a reduced basis, of which the basis vectors will have lengths of about $C^{n /(m+n)}$, which is roughly the size of $\mathrm{Q}_{0}$. Generally speaking, the larger C is, the larger the lengths of the basis vectors of a reduced basis will be (and the larger the lower bounds for $\ell(\Gamma)$ and $\ell(\Gamma, y)$ will be).

Let us first treat the homogeneous case, i.e. $\beta_{i}=0$ for all i. Consider
the lattice point $\underline{x}=\mathcal{B} \cdot\left(\mathrm{q}_{1}, \ldots, \mathrm{q}_{\mathrm{m}}, \mathrm{p}_{1}, \ldots \mathrm{p}_{\mathrm{n}}\right)^{\mathrm{T}}$. It is equal to

$$
\underline{x}=\left(q_{1}, \ldots, q_{m}, \tilde{L}_{1}-C \cdot p_{1}, \ldots, \tilde{L}_{n}-C \cdot p_{n}\right)^{T},
$$

where

$$
\tilde{L}_{i}=\sum_{j=1}^{m} q_{j} \cdot\left[C \cdot \vartheta_{i j}\right] \text { for } i=1, \ldots, n .
$$

From the application of the $L^{3}$-algorithm we find a lower bound for $\ell(\Gamma)$, of size $Q_{0}$. We assume it to be large enough (if this is not the case, we try a somewhat larger value for $C$, and perform the $L^{3}$-algorithm again for the lattice defined for this $C$ ). So we may assume that there is a small constant $c_{1}$ such that

$$
\sum_{i=1}^{n}\left(\tilde{L}_{i}-C \cdot p_{i}\right)^{2} \geqslant \ell(\Gamma)^{2}-m \cdot Q_{0}^{2}>c_{1} \cdot Q_{0}^{2}
$$

We have $\left|\tilde{L}_{i}-C \cdot L_{i}\right| \leqslant m \cdot Q_{0}$, so we may assume that for small constants $c_{2}, c_{3}$

$$
\max _{i}\left|L_{i}-p_{i}\right|>c_{2} \cdot C^{-1} \cdot \max \left|\tilde{L}_{i}-C \cdot p_{i}\right|>c_{3} \cdot Q_{0} / C .
$$

By the choice of $C$ this last bound has the required size.

Next, we study the inhomogeneous case, where not all $\beta_{i}$ are zero. We take the same lattice $\Gamma$ as in the homogeneous case (note that the lattice definition depends only on the $\vartheta_{i j}$ and the C ). Consider the point

$$
y=\left(0, \ldots, 0,-\left[C \cdot \beta_{1}\right], \ldots,-\left[C \cdot \beta_{n}\right]\right)^{T} .
$$

From the reduced basis found by the $L^{3}$-algorithm we have a lower bound for $\ell(\Gamma, \underline{y})$. Assume that it is large enough, and of size $Q_{0}$. We take the same lattice point $\underline{x}=\mathcal{B} \cdot\left(\mathrm{q}_{1}, \ldots, \mathrm{q}_{\mathrm{m}}, \mathrm{p}_{1}, \ldots \mathrm{p}_{\mathrm{n}}\right)^{\mathrm{T}}$ as in the homogeneous case. Then

$$
\underline{x}-\underline{y}=\left(q_{1}, \ldots, q_{m}, \tilde{L}_{1}-C \cdot p_{1}, \ldots, \tilde{L}_{n}-C \cdot p_{n}\right)^{T}
$$

where

$$
\tilde{L}_{i}=\left[C \cdot \beta_{i}\right]+\sum_{j=1}^{m} q_{j} \cdot\left[C \cdot \vartheta_{i j}\right] \text { for } i=1, \ldots, n .
$$

The same reasoning as in the homogeneous case now yields the desired result. Note that if we have performed the $L^{3}$-algorithm once for given $\vartheta_{i j}$, we may use the result to treat the homogeneous case, and many inhomogeneous cases with different $\beta_{i}$ 's as well, as long as the $\vartheta_{i j}$ 's are the same.

The above process describes how to find lower bounds for systems of diophantine inequalities. It will be clear from the above that it is not difficult to find good solutions, i.e. $\left(q_{1}, \ldots, q_{m}, p_{1}, \ldots, p_{n}\right)$ with $Q \leqslant Q_{0}$ and $\max _{i} L_{i}{ }^{-p_{i}} l^{l}$ near to the best possible value. In particular, the basis vectors of a reduced basis are adequate for the homogeneous case, and for the inhomogeneous case the lattice points near to $\mathbb{Y}$ will be such solutions. The lattice points near to $y$ are not difficult to find once a reduced basis is available. Specifically, if $s_{1}, \ldots, s_{n} \in \mathbb{R}$ are the coordinates of $y$ with respect to a reduced basis, then one may take the lattice points with coordinates (with respect to the reduced basis) $t_{i} \in \mathbb{Z}$ that are near to $\mathbf{s}_{i}$ for $i=1, \ldots, n$.

In the definition of the matrix above the expressions $[C \cdot \vartheta j]$ occur. Using these expressions we have constructed a lattice $\Gamma$ that is completely integral, i.e. $\Gamma \subset \mathbb{Z}^{m+n}$. The $L^{3}$-algorithm can be adapted to work exact for those lattices, so that rounding-off errors are avoided (cf. Section 3.5). The "errors" occur only in the difference between the $\tilde{L}_{i}$ and the $C \cdot L_{i}$, and are thus kept under control by choosing the proper constants $c_{1}, c_{2}, c_{3}$. Of course one should take care to have the numerical values of the $\vartheta_{i j}$ and the $\beta_{i}$ correct to sufficient precision. We shall discuss such numerical problems briefly in Section 2.5.

A possible variation of the above diophantine approximation problem is to give weights to the linear forms $\mathrm{L}_{\mathrm{i}}$, i.e. to look for a lower bound for

$$
\max _{i} w_{i} \cdot \mid L_{i}-p_{i} l
$$

where the $w_{i}$ are fixed positive numbers. This situation can be dealt with easily by replacing every $C$ in the $(n+i)$ th row of the matrix by $C \cdot W_{i}$.

Another variation is the problem where not all the variables $q_{j}$ have the same upper bound $Q_{0}$. To illustrate this, assume that $n=1$, and that

$$
L=\sum_{j=1}^{m} q_{j} \cdot \vartheta_{j}
$$

Now suppose that for some $Q_{1}>Q_{2}$ (it will be handy to have $Q_{2} \mid Q_{1}$ ) we are interested in the solutions with

$$
\left|q_{j}\right| \leqslant Q_{1} \text { for } j \leqslant m_{1}, \quad\left|q_{j}\right| \leqslant Q_{2} \text { for } j \geqslant m_{1}+1
$$

Next, let $C$ be of the size of $Q_{1} 1^{+1} \cdot Q_{2}^{m-m} 1$, and take the matrix

Its determinant is of the size of $Q_{1}{ }^{m+1}$. For a lattice point $\left(q_{1}, \ldots, q_{m}, \tilde{L}-C \cdot p\right)^{T}$ we therefore expect that $\max \left(\left|q_{1}\right|, \ldots,\left|q_{m}\right|\right)$, $\left(Q_{1} / Q_{2}\right) \cdot \max \left(\left|q_{m_{1}+1}\right|, \ldots,\left|q_{m}\right|\right)$ and $|\tilde{L}-C \cdot p|$ are all of the size of $Q_{1}$. It follows that $|L-p|$ is of the size of $Q_{1}^{-m_{1}} \cdot Q_{2}^{-\left(m-m_{1}\right)}$, in accordance with the heuristics. This variant is useful when a combination of real and p-adic techniques is used, such as for the Thue-Mahler equation (see Section 8.6).

We conclude this section by giving the analogous method of p-adic diophantine approximation. We assume that the $\vartheta_{i j}, \beta_{i}$ are in $\mathbb{Q}_{p}$, and, moreover, that they are p-adic integers. Let $\mathbb{N}_{0}=\mathbb{N} \cup\{0\}$. For any p-adic integer $\gamma$ and any $\mu \in \mathbb{N}_{0}$ we denote by $\gamma^{(\mu)}$ the unique rational integer such that

$$
\gamma \equiv \gamma^{(\mu)}\left(\bmod p^{\mu}\right), \quad 0 \leqslant \gamma^{(\mu)}<p^{\mu} .
$$

Let $\mu \in \mathbb{N}$ be such that $p^{\mu}$ is roughly the same size as $Q_{0}^{1+m / n}$, and assume that $\mu$ is large enough (it is the analogue of the constant $C$ in the real case above). Take for $\Gamma$ the lattice of which a basis is given by the column vectors of the matrix

$$
\mathcal{B}=\left(\begin{array}{cccccc}
1 & & & & & \\
& \cdot & & & & \varnothing \\
& \cdot & & \\
\vartheta_{11}^{(\mu)} & & & 1 & & \\
11 & \ldots & \vartheta_{1 m}^{(\mu)} & \mathrm{p}^{\mu} & & \\
\vdots & & \vdots & & \cdot & \\
\vartheta_{\mathrm{n} 1}^{(\mu)} & \ldots & & \vartheta_{\mathrm{nm}}^{(\mu)} & & \\
& & & \mathrm{p}^{\mu}
\end{array}\right)
$$

Consider the lattice point

$$
\mathcal{B} \cdot\left(q_{1}, \ldots, q_{m}, z_{1}, \ldots, z_{n}\right)^{T}=\left(q_{1}, \ldots, q_{m}, p_{1}, \ldots, p_{n}\right)^{T}
$$

Then it is obvious that

$$
p_{i}=\sum_{j=1}^{m} q_{j} \cdot \vartheta_{i j}^{(\mu)}+z_{i} \cdot p^{\mu} .
$$

Hence the lattice $\Gamma$ can be described as the set

$$
\begin{aligned}
& \Gamma=\left\langle\left(q_{1}, \ldots, q_{m}, p_{1}, \ldots, p_{n}\right)^{T} \in \mathbb{Z}^{m+n}\right. \\
&\left.\quad \sum_{j=1}^{m} q_{j} \cdot \vartheta_{i j} \equiv p_{i}\left(\bmod p^{\mu}\right) \text { for } i=1, \ldots, n\right\rangle .
\end{aligned}
$$

The $L^{3}$-algorithm provides a lower bound for the length of the nonzero vectors in this set, which is of the same size as $p^{\mu \cdot n /(n+m)}$, and that of $Q_{0}$. This yields the desired result, if $\mu$ is taken large enough.

For the inhomogeneous case, put

$$
y=\left(0, \ldots, 0,-\beta_{1}^{(\mu)}, \ldots,-\beta_{n}^{(\mu)}\right)^{T}
$$

and consider the set

$$
\begin{aligned}
& \Gamma^{*}=\left\langle\left(q_{1}, \ldots, q_{m}, p_{1}, \ldots, p_{n}\right)^{T} \in \mathbb{Z}^{m+n}\right| \\
& \beta_{i}\left.+\sum_{j=1}^{m} q_{j} \cdot \vartheta_{i j} \equiv p_{i}\left(\bmod p^{\mu}\right) \text { for } i=1, \ldots, n\right\rangle .
\end{aligned}
$$

Then $\underline{x} \in \Gamma^{*}$ if and only if $\underline{x}+\underline{y} \in \Gamma$, so $\Gamma^{*}$ is a translated lattice. A lower bound for $\ell(\Gamma, y)$ now yields the desired result.

Again variations are possible, as in the real case, e.g. by replacing on the $(n+i)$ th row the $\mu$ by different $\mu_{i}$. It is even possible in this way to treat more than one prime $p$ at the same time, by replacing on the ( $n+i$ ) th row the $p^{\mu}$ by different $p_{i}^{\mu_{i}}$.

We indicate one more variation for the p-adic case. Suppose we have only one linear form $\Lambda=\sum_{j=1}^{m} q_{j} \cdot \vartheta_{j}$, and one variable $p \in \mathbb{Z}$, and we want to study when $\Lambda$ is congruent to 0 modulo different prime powers $p_{1}^{\mu_{1}}, \ldots, p_{n}^{\mu_{n}}$. Thus we are interested in the set

$$
\begin{array}{r}
\Gamma^{\prime}=\left.\langle |\left[q_{1}, \ldots, q_{m}, p\right]\right|^{T} \in \mathbb{Z}^{m+1} \mid \sum_{j=1}^{m} q_{j} \cdot \vartheta{ }_{j} \equiv p\left(\bmod p_{i}^{\mu_{i}}\right) \\
\text { for } i=1, \ldots, n\rangle
\end{array}
$$

Then we take $\vartheta_{j}^{*} \in \mathbb{Z}$ with

$$
\vartheta_{\mathrm{j}}^{*} \equiv \vartheta_{\mathrm{j}}\left(\bmod p_{\mathrm{i}}^{\mu_{i}}\right) \text { for } \mathrm{i}=1, \ldots, \mathrm{n}, \quad 0 \leqslant \vartheta_{\mathrm{j}}^{*}<\prod_{\mathrm{i}=1}^{\mathrm{n}} \mathrm{p}_{\mathrm{i}}^{\mu_{i}}
$$

for all j . The $\vartheta_{\mathrm{j}}^{*}$ can be computed by the Chinese Remainder Theorem. Now $\Gamma$ ' is the lattice generated by the column vectors of

$$
\left(\begin{array}{ccccc}
1 & & & \\
& \ddots & & \\
& \varnothing & 1 & & \\
& & & \\
\vartheta_{1}^{*} & \ldots & \vartheta_{m}^{*} & \prod_{i=1}^{n} p_{i}^{\mu_{i}}
\end{array}\right)
$$

and we proceed with this lattice as described above.

We conclude this section with three remarks. Firstly, in the case that the dimension of the lattice under consideration is only 2 , the $\mathrm{L}^{3}$-algorithm is essentially the continued fraction algorithm, and so yields nothing new. For the p-adic continued fraction algorithm, see de Weger [1986 ${ }^{2}$ ]. Secondly, the inhomogeneous case of diophantine approximation of one linear form of real numbers can also be treated by what is known as Davenport's lemma, cf. Baker and Davenport [1969] (and its multi-dimensional generalization, cf. Ellison [1971]). We will return to this in Chapter 3, and explain there why we prefer our method.

Finally, one of the nice features of the above method of practical diophantine approximation is that if an extreme solution exists, then in the homogeneous case the lattice (with proper constant $C$ or $\mu$ ) will be distorted. This means that the reduced basis will not be as nice as expected, for example there might be a basis vector in it that is substantially shorter than the other ones. In the inhomogeneous case the existence of an extreme solution means that there is a lattice point extremely near to $y$. The algorithm detects such an extraordinary situation at once, and in most cases the extremal solution is presented explicitly (e.g. in the homogeneous case as one of the vectors of the reduced basis). One can check whether this extremal solution actually satisfies the original equation, and then proceed by replacing in the above reasoning $\ell(\Gamma)$ or $\ell(\Gamma, \underline{y})$ by lower bounds for all vectors in the lattice except the extremal one. These new lower bounds will in general be of the expected size. However, when we solved diophantine equations in practice, we have never met such an extraordinary situation.

### 1.5. The procedure for reducing upper bounds.

We have seen in Section 1.2 how upper bounds for the solutions of the exponential inequalities and equations occuring there can be found. In Section 1.4 we have studied some diophantine approximation theory from a practical point of view. Now these two things come together.

From the application of the Gelfond-Baker theory we are left with the following problem. We have a linear form

$$
\Lambda=\beta+\sum_{j=1}^{m} n_{j} \cdot \vartheta_{j}
$$

where the $\beta$ and $\vartheta_{j}$ are constants (that they are logarithms of algebraic numbers is now of no importance anymore), and the $n_{j}$ are integral unknowns. We know that $\Lambda$ is extremely close to 0 , namely

$$
|\Lambda|<c \cdot \exp (-\delta \cdot N)
$$

where $c, \delta$ are (small) constants, and $N=\max \ln _{j} l$. Finally, we have an explicit upper bound $N_{0}$ for $N$. This $N_{0}$ is very large, $10^{50}$ say.

It will be clear from Section 1.4 that the methods outlined there are of use for solving this problem. For $Q_{0}$ we take $N_{0}$. We have $n=1$. In the real case we expect, by choosing $C$ at least of size $N_{0}^{m+1}$, that

$$
|\Lambda|>c^{\prime} \cdot N_{0}^{-m}
$$

for a small constant $c$ ' . It follows by combining the two inequalities for $|\Lambda|$ that

$$
N<\log \left(c / c^{\prime}\right) / \delta+(m / \delta) \cdot \log N_{0}
$$

So the upper bound $N_{0}$ for $N$ is reduced to an upper bound $N_{1}$ of the size of $\log \mathrm{N}_{0}$, which is a considerable improvement indeed. We now may apply the procedure with $N_{1}$ instead of $N_{0}$, and repeat until no further improvement is obtained. In practice it appears almost always to be the case that in that situation the reduced upper bound is near to the actual largest solution, anyway so small that simple methods of finding all the solutions below that bound suffice.

In the p-adic case an analogous reduction of upper bounds can be reached,
following a similar argument. We have for the linear form $\Lambda$ (cf. (1.4)),

$$
\operatorname{ord}_{p}(\Lambda) \geqslant c_{1}+c_{2} \cdot m_{j},
$$

where $c_{1}, c_{2}$ are small constants, and $m_{j}$ is one of the variables. Moreover, the variables are bounded by a large constant $N_{0}$, that is explicitly known. We take $\mu$ such that $p^{\mu}$ is at least of size $N_{0}^{m+1}$, so that the lower bound for the shortest nonzero vector in $\Gamma$ (or $\Gamma^{*}$ ) is larger than $\sqrt{m} \cdot N_{0}$. Then it follows that the elements of the lattice $\Gamma$ (or of the translated lattice $\Gamma^{*}$ ) cannot be solutions of (1.2). Therefore,

$$
c_{1}+c_{2} \cdot m_{j}<\mu,
$$

so that we find a new upper bound for $m_{j}$, that is of the size of $\mu$, which is about $\log N_{0} / \log p$. We repeat this procedure for all the $m_{j}$, in order to obtain a reduced upper bound for $H_{p}$. If this is not yet sufficient to derive at once a reduced upper bound for $H$, then we can do so by applying a reduction step for real linear forms, where we may take advantage of the fact that for some of the variables a much better upper bound has just been found (cf. the second variation in Section 1.4). Again we repeat the whole procedure as far as possible.

## Chapter 2. Preliminaries.

### 2.1. Algebraic number theory.

In this section we quote results from algebraic number theory that we use throughout the remaining chapters. We refer to Borevich and Shafarevich [1966] or any other textbook on algebraic number theory for full details.

Let $K$ be a finite algebraic extension of $\mathbb{Q}$, of degree $D=[K: \mathbb{Q}]$. There are $D$ embeddings $\sigma: K \rightarrow \mathbb{C}$. Let $\alpha \in K$ be an element of degree $d$, and let $a_{0}>0$ be the leading coefficient of its minimal polynomial over $\mathbb{Z}$. We define the (logarithmic) height $h(\alpha)$ by

$$
h(\alpha)=\frac{1}{\mathrm{D}} \cdot \log \left(\mathrm{a}_{0}^{\mathrm{D} / \mathrm{d}} \cdot \prod_{\sigma}^{\max (1,|\sigma(\alpha)|))}\right.
$$

where the product is taken over all embeddings $\sigma$. Note that this definition does not depend on the field $K$. Hence, if the conjugates of $\alpha$ are $\alpha=\alpha_{1}, \ldots, \alpha_{d}$, then the above definition applied for $K=\mathbb{Q}(\alpha)$ yields

$$
h(\alpha)=\frac{1}{d} \cdot \log \left(a_{0} \cdot \prod_{i=1}^{d} \max \left(1,\left|\alpha_{i}\right|\right)\right)
$$

In particular, if $\alpha \in \mathbb{Q}$, then with $\alpha=p / q$ for $p, q \in \mathbb{Z}$ with ( $p, q$ ) = 1 we have $h(\alpha)=\log \max (|p|,|q|)$, and if $\alpha \in \mathbb{Z}$ then $h(\alpha)=\log |\alpha|$.

Let there be $s$ real and $2 \cdot \mathrm{t}$ non-real embeddings (with $\mathrm{D}=\mathrm{s}+2 \cdot \mathrm{t}$ ). Then Dirichlet's Unit Theorem states that there exists a system of $r=s+t-1$ independent units $\varepsilon_{1}, \ldots, \varepsilon_{r}$, such that the group of units of $K$ is given by

$$
\left\langle\zeta \cdot \varepsilon_{1}{ }^{\mathrm{a}} \cdot \ldots \cdot \varepsilon_{r}{ }^{\mathrm{a}} \mathrm{r} \text { | } \zeta \text { a root of unity, } a_{i} \in \mathbb{Z} \text { for } i=1, \ldots, r\right\rangle
$$

There are only finitely many roots of unity in $K$. Any set of independent units that generate the torsion-free part of the unit group is called a system of fundamental units.

The number $\alpha$ is called an algebraic integer if $a_{0}=1$. Let the norm of an
element $\alpha \in K$ be defined by

$$
N_{K / \mathbb{Q}}(\alpha)=\prod_{\sigma} \sigma(\alpha)=\left(\prod_{i=1}^{d} \alpha_{i}\right)^{\mathrm{D} / \mathrm{d}}
$$

For algebraic integers, $N_{K / \mathbb{Q}}(\alpha) \in \mathbb{Z}$. The units are precisely the elements of norm $\pm 1$. Two elements $\alpha, \beta$ of $K$ are called associates if there is a unit $\varepsilon$ such that $\alpha=\varepsilon \cdot \beta$. Let $(\alpha)$ denote the ideal generated by $\alpha$. Associated elements generate the same ideal, and distinct generators of an ideal are associated. There exist only finitely many non-associated algebraic integers in $K$ with given norm. The ring of algebraic integers is denoted by $\mathcal{O}_{K}$. Let $\alpha_{1}, \ldots, \alpha_{D}$ be elements of $\mathcal{O}_{K}$ that are $\mathbb{Q}-l i n e a r l y$ independent. Then $\mathbb{Z} \cdot \alpha_{1} \times \ldots \times \mathbb{Z} \cdot \alpha_{D}$ is called an order of $K$ if it is a subring of the 'maximal order' $\mathcal{O}_{\mathrm{K}}$.

In $K$ any algebraic integer can be written as a product of irreducible elements. Here an irreducible element (prime element) is an element that has no integral divisors but its own associates. However, this decomposition into primes need not be unique. Ideals can also be decomposed into prime ideals, and this decomposition is unique. A principal ideal is an ideal generated by a single element $\alpha$. Two fractional ideals are called equivalent if their quotient is principal. It is well known that there are only finitely many equivalence classes. Their number is called the class number $h_{K}$. For an ideal $a$ it is always true that $a^{h_{K}}$ is a principal ideal. The norm of the (integral) ideal $a$ is defined by $N_{K / \mathbb{Q}}(a)=\#\left(\mathcal{O}_{K} / a\right)$.

For a prime ideal $p$ there is always a rational prime number $p$ such that $\mathfrak{p}$ is a divisor of ( $p$ ). We say that $p$ lies above $p$. The ramification index $e_{p}$ is the largest power to which $\mathfrak{p}$ divides (p). The residue class degree ${\underset{f}{p}}$ is the integer such that

$$
N_{K / \mathbb{Q}}(\mathfrak{p})=p^{f} \mathfrak{p} .
$$

We denote by $\operatorname{ord}_{\mathfrak{p}}(\mathfrak{a})$ the exact power to which the prime ideal $\mathfrak{p}$ divides the ideal $a$. For fractional ideals $a$ this number can of course be negative. For numbers $\alpha$ we write $\operatorname{ord}_{\mathfrak{p}}(\alpha)$ for $\operatorname{ord}_{\mathfrak{p}}((\alpha))$. Note that

$$
\operatorname{ord}_{p}(\alpha)=\operatorname{ord}_{\mathfrak{p}}(\alpha) / e_{p}
$$

can be defined for all $\alpha \in \mathrm{K}$. We will return to this in Section 2.3, which deals with $p$-adic number theory.

### 2.2. Some auxiliary lemmas.

In this section we give a few simple auxiliary lemmas. The first one enables us to find an upper bound in closed form for some real number $x>1$ that is bounded by a polynomial in $\log x$. See Pethö and de Weger [1986], Lemma 2. 3 .

LEMMA 2.1. Let $\mathrm{a} \geqslant 0, \mathrm{~h} \geqslant 1, \mathrm{~b}>0$, and let $\mathrm{x} \in \mathbb{R}, \mathrm{x}>1$ satisfy $x \leqslant a+b \cdot(\log x)^{h}$.

If $\mathrm{b}>\left(\mathrm{e}^{2} / \mathrm{h}\right)^{\mathrm{h}}$ then

$$
x<2^{h} \cdot\left(a^{1 / h}+b^{1 / h} \cdot \log \left(h^{h} \cdot b\right)\right)^{h},
$$

and if $b \leqslant\left(e^{2} / h\right)^{h}$ then

$$
x \leqslant 2^{h} \cdot\left(a^{1 / h}+2 \cdot e^{2}\right)^{h}
$$

Proof. We may assume that $x$ is the largest solution of

$$
x=a+b \cdot(\log x)^{h}
$$

By $\left(z_{1}+z_{2}\right)^{1 / h} \leqslant z_{1}^{1 / h}+z_{2}^{1 / h}$ we infer

$$
x^{1 / h} \leqslant a^{1 / h}+c \cdot \log \left(x^{1 / h}\right)
$$

where

$$
\begin{aligned}
c= & h \cdot b^{1 / h} \text {. Define } y \text { by } x^{1 / h}=(1+y) \cdot c \cdot \log c \text {. From } \\
& \log c<\log (c \cdot \log c)
\end{aligned}
$$

it follows that

$$
c^{h} \cdot(\log c)^{h}<b \cdot\left(\log \left(c^{h} \cdot(\log c)^{h}\right)\right)^{h}
$$

which implies $x>c^{h} \cdot(\log c)^{h}$. Hence $y>0$. Now,

$$
\begin{aligned}
(1+y) \cdot c \cdot \log c=x^{1 / h} & \leqslant a^{1 / h}+c \cdot \log (1+y)+c \cdot \log c+c \cdot \log \log c \\
& <a^{1 / h}+c \cdot y+c \cdot \log c+c \cdot \log \log c
\end{aligned}
$$

Hence

$$
y \cdot c \cdot(\log c-1)<a^{1 / h}+c \cdot \log \log c .
$$

If $c \geqslant e^{2}$ it follows that

$$
\begin{gathered}
x^{1 / h}=c \cdot \log c+y \cdot c \cdot \log c<c \cdot \log c+\frac{\log c}{\log c-1} \cdot\left(a^{1 / h}+c \cdot \log \log c\right) \\
<2 \cdot\left(a^{1 / h}+c \cdot \log c\right)
\end{gathered}
$$

If $c \leqslant e^{2}$, then note that $x \leqslant a+\left(e^{2} / h\right)^{h} \cdot(\log x)^{h}$. So we may assume $c=e^{2}$ in this case. The result follows.

The next lemmas make explicit that $x$ and $\log (1+x)$ are near if $|x|$ is small in the real and complex case, respectively.

LEMMA 2.2. Let $a \in \mathbb{R}$. If $a<1$ and $|x|<a$ then

$$
|\log (1+x)|<\frac{-\log (1-a)}{a} \cdot|x|
$$

and

$$
|x|<\frac{a}{1-e^{-a}} \cdot\left|e^{x}-1\right| .
$$

Proof. Note that $\log (1+x) / x$ is a strictly positive and strictly decreasing function for $|x|<1$. Hence it is for $|x|<a \quad$ always less than its value at $x=-a$. The same is true for the function $x /\left(e^{x}-1\right)$.

LEMMA 2.3. Let $0<a \leqslant \pi$. If $|x|<a$ then

$$
|x|<\frac{a}{2 \cdot \sin (a / 2)} \cdot\left|e^{i \cdot x}-1\right| .
$$

If $\mathrm{a}<2,\left|\mathrm{e}^{\mathrm{i} \cdot \mathrm{x}}-1\right|<\mathrm{a}$ and $|\mathrm{x}|<\pi$ then

$$
|x|<\frac{2 \cdot \arcsin (a / 2)}{a} \cdot\left|e^{i \cdot x}-1\right|
$$

Proof. Note that $\left|e^{i \cdot x}-1\right|=2 \cdot\left|\sin \left(\frac{1}{2} \cdot x\right)\right|$. and that $2 \cdot \sin \left(\frac{1}{2} \cdot x\right) / x$ is a positive and even function, that decreases on $0 \leqslant x<a$. Hence it takes its minimal value at $x=a$. The first inequality now follows. The second one can be proved in a similar way.

## 2.3. p-adic numbers and functions.

In this section we mention the facts about p-adic numbers and functions that we use. For details we refer to Bachman [1964] and Koblitz [1977], [1980].

We assume that the reader is familiar with the field of $p$-adic numbers $\mathbb{Q}_{p}$ and the $p$-adic valuation ord ${ }_{p}$. Note that the ordinary ord ${ }_{p}$ as defined in $\mathbb{Q}_{p}$ coincides with the definition given in Section 2.1 . We denote by $\Omega_{p}$ the completion of the algebraic closure of $\mathbb{Q}_{p}$, i.e. the field to which all p-adic theory is applied.

Every nonzero number $\alpha \in \mathbb{Q}_{\mathrm{p}}$ has a p -adic expansion

$$
\alpha=\sum_{i=k}^{\infty} u_{i} \cdot p^{i}
$$

where $k=\operatorname{ord}_{p}(\alpha)$ and the $p$-adic digits $u_{i}$ are in $\{0,1, \ldots, p-1\}$, with $u_{k} \neq 0$. The number 0 can be represented in this way by taking $k=0$ and all digits equal to 0 , and $\operatorname{ord}_{p}(0)=\infty$ by definition. If $\operatorname{ord}_{p}(\alpha) \geqslant 0$ then $\alpha$ is called a $p$-adic integer. The set of $p$-adic integers is denoted by $\mathbb{Z}_{\mathrm{p}}$. A p -adic unit is an $\alpha \in \mathbb{Q}_{\mathrm{p}}$ with $\operatorname{ord}_{\mathrm{p}}(\alpha)=0$. For any p -adic integer $\alpha$ and any $\mu \in \mathbb{N}_{0}$ there exists a unique rational integer $\alpha(\mu)=\sum_{i=0}^{\mu-1} u_{i} \cdot p^{i}$ satisfying

$$
\operatorname{ord}_{\mathrm{p}}\left(\alpha-\alpha^{(\mu)}\right) \geqslant \mu, \quad 0 \leqslant \alpha^{(\mu)} \leqslant \mathrm{p}^{\mu}-1
$$

For $\operatorname{ord}_{\mathrm{p}}(\alpha) \geqslant \mathrm{k}$ we also write $\alpha \equiv 0\left(\bmod \mathrm{p}^{k}\right)$. The p -adic norm is defined by

$$
|\alpha|_{p}=p^{-\operatorname{ord}_{p}(\alpha)}
$$

In Section 2.1 we have seen how to define $\operatorname{ord}_{p}$ and ord $_{p}$ on algebraic extensions of $\mathbb{Q}$. For any $\alpha \in \Omega_{p}$ with $\operatorname{ord}_{p}(\alpha)>1 /(p-1)$ we can define the $p$-adic logarithm $\log _{p}(1+\alpha)$ by the Taylor series

$$
\log _{p}(1+\alpha)=\alpha-\alpha^{2} / 2+\alpha^{3} / 3-\ldots
$$

This logarithmic function has the well known properties of a logarithm, such as $\log _{p}\left(\xi_{1} \cdot \xi_{2}\right)=\log _{p}\left(\xi_{1}\right)+\log _{p}\left(\xi_{2}\right)$ for all $\xi_{1}$, $\xi_{2}$ for which it is defined. Further, $\log _{p}(\xi)=0$ if and only if $\xi$ is a root of unity. In $\mathbb{Q}_{p}$ the only roots of unity are the ( $p-1$ ) th roots of unity (if $p$ is odd). Using these properties, this logarithmic function can be extended to all $\xi \in \Omega_{p}$ with $\operatorname{ord}_{p}(\xi)=0$, as follows. By Fermat's theorem for algebraic number fields there is a $k \in \mathbb{N}$ such that $\operatorname{ord}_{p}\left(\xi^{k}-1\right)>1 /(p-1)$. Then

$$
\log _{p}(\xi)=\frac{1}{k} \cdot \log _{p}\left(1+\left(\xi^{k}-1\right)\right)
$$

An equivalent definition is $\log _{p}(\xi)=\log _{p}(\xi / \zeta)$, where $\zeta$ is a root of unity such that $\operatorname{ord}_{p}(\xi-\zeta)>0$. In this way the p-adic logarithm is a well defined function. Note that $\log _{p}(\xi)$ lies in the subfield of $\Omega_{p}$ generated by $\xi$. Finally we note that if $\operatorname{ord}_{p}(\xi)>1 /(p-1)$ then

$$
\operatorname{ord}_{p}(\xi)=\operatorname{ord}_{p}\left(\log _{p}(1+\xi)\right)
$$

### 2.4. Lower bounds for linear forms in logarithms.

In this section we quote in detail the results from the Gelfond-Baker theory that we use. They yield lower bounds for linear forms in logarithms of algebraic numbers. We do not always give the theorems in their full generality, since in this book only linear forms with rational unknowns occur, whereas most Gelfond-Baker theorems are formulated for linear forms with algebraic unknowns. We selected bounds with fully explicit constants, because only such completely explicit results can be used for our purposes.

The first result in this field for a linear form in logarithms with at least three terms is due to Baker [1966], and in the p-adic case to Coates [1969], [1970]. For a survey of this theory, see Baker [1977] and van der Poorten [1977]. We will use more recent, sharper results, due to Waldschmidt [1980] and Yu [1987]. Further improvements of the constants have been reached (see the references after Lemma 2.4 below), but too recently to be taken into account here.

First we deal with real/complex linear forms in logarithms. We quote the result of Waldschmidt [1980].

LEMMA 2.4 (Waldschmidt). Let $K$ be a number field with [K: $\mathbb{Q}]=\mathrm{D}$. Let $\alpha_{1}, \ldots, \alpha_{n} \in K$, and $b_{1}, \ldots, b_{n} \in \mathbb{Z} \quad(n \geqslant 2)$. Let $V_{1}, \ldots, V_{n}$ be positive real numbers satisfying $1 / \mathrm{D} \leqslant \mathrm{V}_{1} \leqslant \ldots \leqslant \mathrm{~V}_{\mathrm{n}}$ and

$$
V_{j} \geqslant \max \left(h\left(\alpha_{j}\right),\left|\log \alpha_{j}\right| / D\right) \text { for } j=1, \ldots, n
$$

where $\log \alpha_{j}$ for $j=1, \ldots, n$ is an arbitrary but fixed determination of the logarithm of $\alpha_{j}$. Let $V_{j}^{+}=\max \left(V_{j}, 1\right)$ for $j=n, n-1$, and put

$$
\Lambda=\sum_{j=1}^{n} b_{j} \cdot \log \alpha_{j}
$$

Put $B=\max _{1 \leqslant i \leqslant n}\left|b_{i}\right|$. If $\Lambda \neq 0$ then

$$
\begin{gathered}
|\Lambda|>\exp \left(-2^{e(n)} \cdot n^{2 \cdot n} \cdot D^{n+2} \cdot V_{1} \cdot \ldots \cdot V_{n} \cdot \log \left(e \cdot D \cdot V_{n-1}^{+}\right)\right. \\
\\
\left.\cdot\left(\log B+\log \left(e \cdot D \cdot V_{n}^{+}\right)\right)\right)
\end{gathered}
$$

where $\mathrm{e}(\mathrm{n})=\min (8 \cdot \mathrm{n}+51,10 \cdot \mathrm{n}+33,9 \cdot \mathrm{n}+39)$. If, moreover, it is known that $\left[\mathbb{Q}\left(\sqrt[V]{ } \alpha_{1}, \ldots, \sqrt{ } \alpha_{n}\right): \mathbb{Q}\right]=2^{n}$, then we can take $e(n)=9 \cdot n+26$ and replace the factor $n^{2 \cdot n}$ in the above bound for $|\Lambda|$ by $n^{n+4}$.

Waldschmidt's main theorem does not give the constant e(n) as detailed as we do, but he does so in his proof, cf. p. 283. We remark that improvements of the above bounds have recently been found by Blass, Glass, Manski, Meronk and Steiner [1988 ${ }^{\text {a }], ~[1988 ~}{ }^{\text {b }], ~ L o x t o n, ~ M i g n o t t e, ~ v a n ~ d e r ~ P o o r t e n ~ a n d ~}$ Waldschmidt [1987], Philippon and Waldschmidt [1988], and Wüstholz [1988]. For the case $n=2$, the sharpest bound has been given by Mignotte and Waldschmidt [1978], improved again by Mignotte and Waldschmidt [1988].

In the p-adic case we quote two results: one due to Schinzel [1967] (Theorem 1) for the case of a linear form in logarithms with two terms, and another for the general case, due to Yu [1987] (Theorem 1, see also Yu [1988]). We note that Yu's bounds improve much upon the results of van der Poorten [1977]. Moreover, van der Poorten's proofs seem to contain some errors. We give Schinzel's result for quadratic fields only.

LEMMA 2.5 (Schinzel). Let p be prime. Let $\Delta$ be a squarefree integer, and let D be the discriminant of $\mathrm{K}=\mathbb{Q}(\sqrt{ } \Delta)$. Let $\xi=\xi \prime / \xi$ and $\chi=\chi \prime / \chi$ ' be elements of K , where $\xi^{\prime}, \xi ", \chi$, $\chi$ " are algebraic integers. Put

$$
L=\log \max \left(|e \cdot D|^{1 / 4},\left\|\xi^{\prime} \cdot \chi^{\prime}\right\|,\left\|\xi^{\prime} \cdot \chi "\right\|,\|\xi " \cdot \chi \prime\|,\left\|\xi^{\prime \prime} \cdot \chi "\right\|\right) \text {, }
$$

where ll $\gamma \|$ denotes the maximal absolute value of the conjugates of $\gamma \in K$. Let $\mathfrak{p}$ be a prime ideal of $K$ with norm $N p=p^{\rho}$. Put $\psi=2 / \rho \cdot \log p$, $\varphi=\operatorname{ord}_{\mathfrak{p}}(\mathrm{p})$. If $\xi$ or $\chi$ is a $\mathfrak{p}$-adic unit and $\xi^{n} \neq \chi^{m}$, then

$$
\operatorname{ord}_{\mathfrak{p}}\left(\xi^{n}-\chi^{m}\right)<10^{6} \cdot \psi^{7} \cdot \varphi^{-2} \cdot L^{4} \cdot p^{4} \cdot \rho+4 \cdot\left(\log \max (|m|,|n|)+\varphi \cdot L \cdot p^{\rho}+2 / L\right)^{3}
$$

LEMMA 2.6 (Yu). Let $\alpha_{1}, \ldots, \alpha_{n}(\mathrm{n} \geqslant 2)$ be nonzero algebraic numbers. Put $\mathrm{L}=\mathbb{Q}\left(\alpha_{1}, \ldots, \alpha_{\mathrm{n}}\right), \mathrm{d}=[\mathrm{L}: \mathbb{Q}]$. Let $\mathrm{b}_{1}, \ldots, \mathrm{~b}_{\mathrm{n}}$ be rational integers. Let $\mathfrak{p}$ be a prime ideal of $L$, lying above the rational prime $p$. Let $e_{p}$ be the ramification index, and $f_{p}$ the residue class degree of $\mathfrak{p}$. Write $L_{\mathfrak{p}}$ for the completion of $L$ with respect to $\operatorname{ord}_{\mathfrak{p}}$ (then for all $\beta \in L_{p}$ we have $\left.\operatorname{ord}_{\mathfrak{p}}(\beta)=e_{p} \cdot \operatorname{ord}_{p}(\beta)\right)$. Let $q$ be a rational prime such that $q \nmid p \cdot\left(p^{f} p_{-1}\right)$.

Let

$$
\begin{aligned}
& V_{j} \geqslant \max \left(h\left(\alpha_{j}\right), f_{p} \cdot(\log p) / d\right) \text { for } j=1, \ldots, n, \\
& \quad \operatorname{such} \text { that } V_{1} \leqslant \ldots \leqslant V_{n-1}, V_{n-1}^{+}=\max \left(1, V_{n-1}\right), \\
& B_{0} \geqslant \min _{1 \leqslant j \leqslant n, b_{j} \neq 0}\left|b_{j}\right|, B_{n} \geqslant\left|b_{n}\right|, B^{\prime} \geqslant \max _{1 \leqslant j \leqslant n-1}\left|b_{j}\right|, \\
& B \geqslant \max \left(\left|b_{1}\right|, \ldots,\left|b_{n}\right|, 2\right), \\
& W \geqslant \max \left(\log \left(1+\frac{3}{4 \cdot n} \cdot B, \log B_{0}, f_{p} \cdot(\log p) / d\right) .\right.
\end{aligned}
$$

Suppose that $\operatorname{ord}_{\mathfrak{p}}\left(\alpha_{j}\right)=0$ for $j=1, \ldots, n$, that

$$
\begin{equation*}
\left[L\left(\alpha_{1}^{1 / q}, \ldots, \alpha_{n}^{1 / q}\right): L\right]=q^{n} \tag{2.1}
\end{equation*}
$$

that $\operatorname{ord}_{p}\left(b_{n}\right) \leqslant \operatorname{ord}_{p}\left(b_{j}\right)$ for $j=1, \ldots, n$, and $\alpha_{1} b_{1} \ldots \cdot \alpha_{n}{ }_{n} \neq 1$. Then

$$
\begin{gathered}
\operatorname{ord}_{\mathfrak{p}}\left(\alpha_{1}^{b_{1}} \cdot \ldots \cdot \alpha_{n}^{b}-1\right)<C_{1}(p, n) \cdot a_{1}^{n} \cdot n^{n+5 / 2} \cdot q^{2 \cdot n} \cdot(q-1) \cdot \log ^{2}(n \cdot q) \cdot \\
\quad\left(p^{f} \mathfrak{p}_{-1}\right) \cdot\left|\left[2+\frac{1}{p-1}\right]\right|^{n} \cdot\left|\left[f_{p} \cdot(\log p) / d\right]\right|^{-(n+2)} \cdot V_{1} \cdot \ldots \cdot V_{n} \\
\quad \cdot\left(\frac{W}{6 \cdot n}+\log (4 \cdot d)\right) \cdot\left(\log \left(4 \cdot d \cdot V_{n-1}^{+}\right)+f_{p} \cdot(\log p) / 8 \cdot n\right)
\end{gathered}
$$

where

$$
\mathrm{a}_{1}=56 \cdot \mathrm{e} / 15 \text { if } \mathrm{n} \leqslant 7, \quad \mathrm{a}_{1}=8 \cdot \mathrm{e} / 3 \text { if } \mathrm{n} \geqslant 8 \text {, }
$$

and $C_{1}(\mathrm{p}, \mathrm{n})$ is given by the table on the next page, with for $\mathrm{p} \geqslant 5$

$$
C_{1}(p, n)=C_{1}^{\prime}(p, n) \cdot\left|\left[2+\frac{1}{p-1}\right]\right|^{2}
$$

| $n$ | 2 | 3 | 4 | 5 | 6 | 7 | $\geqslant 8$ |
| :---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| $C_{1}(2, n)$ | 768523 | 476217 | 373024 | 318871 | 284931 | 261379 | 2770008 |
| $C_{1}(3, n)$ | 167881 | 104028 | 81486 | 69657 | 62243 | 57098 | 116055 |
| $C_{1}^{\prime}(p, n)$ | 87055 | 53944 | 42255 | 36121 | 32276 | 24584 | 311077 |

Remark. Yu [1989] gives a result in which 'independence condition' (2.1) has been removed, with more or less the same constants. This result will be easier to apply if $d \geqslant 1$.

### 2.5. Numerical methods.

In solving diophantine equations using computational methods from diophantine approximation theory, as we will do in Chapters 4 to 8 , it is necessary to have logarithms (real, complex or p-adic) of algebraic numbers available to a large enough precision (maybe several hundreds of digits). We will not go deeply into the problems of computing such approximations, but make only a few remarks on it in this section.

To start with, the precision with which most computers (mainframes as well as personal computers) work, is insufficient for our purposes. Usually at most double precision ( 52 bits, equivalent to 15 decimal digits), or at best quadruple precision (112 bits, equivalent to 33 decimal digits) is standard available. This is not sufficient for our purposes, not only because we may require larger precision, but also because we want to have the rounding off errors under control, to be sure that no solution of a diophantine equation is missed by unexpected consequences of rounding off errors.

Packages for computations with arbitrary precision are available and very useful, e.g. the MP package of R.P. Brent (cf. Brent [1978]). It is not difficult, as we did, to write one's own package for simple manipulations on multi-precision numbers, such as addition, multiplication and division (cf. Knuth [1981] for efficient algorithms). To the author's knowledge, no such packages are available publicly for manipulations on p-adic numbers, but the programs are similar to those for real numbers, and thus relatively easy (though maybe laborious) to write yourself.

Computing roots of polynomials with integral coefficients can be done by

Newton's method, both in the real and the p-adic case. One should make sure that the result obtained is correct to the desired precision, not (only) by substituting the found approximation of the root into the polynomial and checking that the result is 0 within the desired precision, but (also) by theoretical error estimates for the Newton method, or by using 'interval arithmetic' (see below).

Computing logarithms can be done by the Newton method too. However, we found it easier to use the Taylor series

$$
\log (1+x)=x-x^{2} / 2+x^{3} / 3-\ldots,
$$

or the more rapidly converging series

$$
\log \frac{1+x}{1-x}=2 \cdot\left(x+x^{3} / 3+x^{5} / 5+\ldots\right)
$$

For $|x|$ very small this method works fast, whereas for larger $|x|$ the following idea works well. Compute approximations to the desired precision of $\log 1.1, \log 1.0001$, $\log 1.00000001$, say, and store them. Now compute $\mathrm{x}_{1} \in[1,1.1)$ and $\mathrm{k}_{1} \in \mathbb{N}_{0}$ such that

$$
x=x_{1} \cdot 1.1^{k_{1}}
$$

which is a matter of a few divisions of a multi-precision number with a rational number with small numerator and denominator (11 and 10) only, that can be done fast. Next, compute $x_{2} \in\left[1,1.0001\right.$ ) and $k_{2} \in \mathbb{N}_{0}$ such that

$$
x_{1}=x_{2} \cdot 1.0001^{k_{2}}
$$

and $x_{3} \in[1,1.00000001)$ and $k_{3} \in \mathbb{N}_{0}$ such that

$$
x_{2}=x_{3} \cdot 1.00000001^{\mathrm{k}_{3}}
$$

Then compute $\log x_{3}$ by the Taylor series, which converges very fast, and compute $\log x$ by

$$
\log x=\log x_{3}+k_{3} \cdot \log 1.00000001+k_{2} \cdot \log 1.0001+k_{1} \cdot \log 1.1 .
$$

When computing all this, one should take care of having the rounding off errors at each addition/multiplication under control. This can e.g. be done by using 'interval arithmetic', i.e. doing all computations twice with a few more digits than actually needed, rounding off in different directions at
each step. Then a sufficiently small interval is found in which the exact number lies (with mathematical certainty).

Computation of $\arctan \mathrm{x}$ is done by the Taylor series

$$
\arctan x=x-x^{3} / 3+x^{5} / 5-\ldots
$$

The number $\pi=3.14159 .$. can be computed rapidly by this series for the arctan function, by the identity

$$
\pi=16 \cdot \arctan 1 / 5-4 \cdot \arctan 1 / 239 .
$$

Doing p-adic arithmetic has the advantage above real arithmetic that rounding off errors do not tend to become larger, as long as one is not dividing by a number with positive $p$-adic order. If ord $(x)>0$ then $\log _{p}(1+x)$ can be computed by the Taylor series

$$
\log _{p}(1+x)=x-x^{2} / 2+x^{3} / 3+\ldots
$$

and also it may be useful to compute it by

$$
\log _{p} \frac{1+x}{1-x}=2 \cdot\left(x+x^{3} / 3+x^{5} / 5+\ldots\right)
$$

If $x \not \equiv 0(\bmod p)$ and $x \not \equiv 1(\bmod p)$ then $\log _{p} x$ can be computed, since there exists a $k \in \mathbb{N}$ such that $x^{k} \equiv 1(\bmod p)$, and then

$$
\log _{p} x=\frac{1}{k} \cdot \log _{p}\left(1+\left(x^{k}-1\right)\right)
$$

and the above given Taylor series can be used to compute $\log _{p} x$. Note that in computing the above mentioned Taylor series there will be factors $p$ in the denominators of the terms. Hence, to find the first $\mu$ p-adic digits of $\log _{p}(1+x)$, it is not enough to compute only the first $\mu / \operatorname{ord}_{p}(x)$ terms of the Taylor series, but the first $k$ terms must be taken into account, where k is the smallest integer satisfying

```
k\cdotord
```

For rapid convergence of Taylor series it is desirable to apply them only for numbers $x$ with large p-adic order. For example,

$$
\log _{3} 4=3-3^{2} / 2+3^{3} / 3-\ldots
$$

converges not as fast as

$$
\log _{3} 4=\frac{1}{3} \cdot \log _{3} 64=\frac{1}{3} \cdot\left(7 \cdot 3^{2}-7^{2} \cdot 3^{4} / 2+7^{3} \cdot 3^{6} / 3-\ldots\right)
$$

or as

$$
\log _{3} 4=\log _{3} \frac{1+3 / 5}{1-3 / 5}=2 \cdot\left(3 / 5+3^{3} / 3 \cdot 5^{3}+3^{5} / 5 \cdot 5^{5}+\ldots\right)
$$

or as

$$
\begin{aligned}
\log _{3} 4= & \frac{1}{3} \cdot \log _{3} \frac{1+7 \cdot 3^{2} / 65}{1-7 \cdot 3^{2} / 65}=\frac{2}{3} \cdot\left(7 \cdot 3^{2} / 65+7^{3} \cdot 3^{6} / 3 \cdot 65^{3}\right. \\
& \left.+7^{5} \cdot 3^{10} / 5 \cdot 65^{5}+\ldots\right)
\end{aligned}
$$

The above considerations are sufficient for efficiently performing exact computations with the $\mathrm{L}^{3}$-algorithm, as we present it in Section 3.5. We also use the simple continued fraction algorithm in some instances. This we do as follows. Suppose we want to compute the continued fraction expansion of a real number $\vartheta$, that we have approximated by rational numbers $\vartheta_{1}, \vartheta_{2}$ such that

$$
\vartheta_{1}<\vartheta<\vartheta_{2}<\vartheta_{1}+\varepsilon
$$

for some small $\varepsilon$. We can compute the continued fraction expansions of $\vartheta_{1}$ and $\vartheta_{2}$ exactly. As far as they coincide, they coincide also with the continued fraction expansion of $\vartheta$. If the continued fraction expansion of $\vartheta$ is needed so far that the $k$ th convergent with denominator $q_{k}>X_{0}$ be known exactly, for a given (large) constant $X_{0}$, then $\varepsilon$ should be at least as small as $X_{0}^{-2}$.

Most of the computer calculations done for the research on which this book reports were performed on an IBM 3083 computer at the Centraal Rekeninstituut of the University of Leiden, using the Fortran-77 language. Whenever we give computation times, actual CPU-time on this machine is meant. Also some computations were done at a VAX $11 / 750$ computer at the Rekencentrum of the University of Twente.

## Chapter 3. Algorithms for diophantine approximation.

### 3.1. Introduction.

In this section we give details of the computational methods we use to reduce upper bounds for the solutions of diophantine equations. Our starting point will always be a linear form $\Lambda$ that is close to 0 (in the real or padic sense, with the word "close" defined explicitly in terms of an inequality involving the unknowns), together with a large but explicitly known upper bound for the absolute values of the coefficients of $\Lambda$. Our aim is to reduce the upper bound by showing that there are no solutions between the new and the old upper bound.

Let $\vartheta_{1}, \ldots, \vartheta_{n}, \beta$ be given numbers, in $\mathbb{R}$, or in $\Omega_{p}$, for a fixed prime p. Let $x_{1}, \ldots, x_{n}$ be unknowns in $\mathbb{Z}$. Put

$$
\Lambda=\beta+\sum_{i=1}^{n} x_{i} \cdot \vartheta_{i}
$$

We classify such linear forms according to three criteria:
$\rightarrow$ homogeneous if $\beta=0$, inhomogeneous if $\beta \neq 0$;
$\rightarrow$ one-dimensional if $n=2$, multi-dimensional if $n \geqslant 3$;
$\rightarrow$ real if $\vartheta_{i} \in \mathbb{R}$ for all i, p-adic if $\vartheta_{i} \in \Omega_{p}$ for all i.

The reason that the case $n=2$ is called one-dimensional is that in the homogeneous case the linear form

$$
\Lambda=\mathrm{x}_{1} \cdot \vartheta_{1}+\mathrm{x}_{2} \cdot \vartheta_{2}
$$

leads to studying the simple, one-dimensional continued fraction expansion of $-\vartheta_{1} / \vartheta_{2}$. The inhomogeneous case with $n=1$, viz.

$$
\Lambda=\beta+x \cdot \vartheta
$$

is not of any interest in the real case, but it is of interest in the p-adic case. We call this the zero-dimensional case.

In the p-adic case we require that the quotients $\vartheta_{i} / \vartheta_{j}$ and $\beta / \vartheta_{j}$ are in $\mathbb{Q}_{p}$ itself, whereas the numbers $\vartheta_{i}, \beta$ are allowed to be in some larger subfield of $\Omega_{p}$.

Let $c, \delta$ be positive constants. Put $X=\max _{\mathrm{I}}^{\mathrm{l}} \mathrm{I}$. Let $\mathrm{X}_{0}$ be a (large) positive constant. In the real case we shall always assume that

$$
\begin{align*}
& |\Lambda|<c \cdot \exp (-\delta \cdot X)  \tag{3.1}\\
& X \leqslant X_{0} \tag{3.2}
\end{align*}
$$

Let $c_{1}, c_{2}$ be real constants, with $c_{2}>0$. In the p-adic case we shall assume that $\mathrm{x}_{\mathrm{j}}>0$ for some index $\mathrm{j} \in\{1, \ldots, \mathrm{n}\}$, and

$$
\begin{align*}
& \operatorname{ord}_{p}(\Lambda) \geqslant c_{1}+c_{2} \cdot x_{j}  \tag{3.3}\\
& X \leqslant X_{0} \tag{3.4}
\end{align*}
$$

Our aim is to find a constant $X_{1}$, of the size of $\log X_{0}$, such that in the real case (3.2) can be replaced by $X \leqslant X_{1}$, and in the $p$-adic case the bound $\mathrm{X}_{\mathrm{j}} \leqslant \mathrm{X}_{0}$ (a consequence of (3.4)) can be improved to $\mathrm{X}_{\mathrm{j}} \leqslant \mathrm{X}_{1}$.

In the forthcoming sections we will treat all cases, according to the classification given above. We insert Sections 3.4, 3.5 on the $\mathrm{L}^{3}$-algorithm, which will be our main computational tool, Section 3.6 on finding short vectors in lattices, and Section 3.13 on certain sublattices that are useful for our applications.

### 3.2. Homogeneous one-dimensional approximation in the real case: continued fractions.

We first study the case

$$
\Lambda=x_{1} \cdot \vartheta_{1}+x_{2} \cdot \vartheta_{2}
$$

Put $\vartheta=-\vartheta_{1} / \vartheta_{2}$. We assume that $\vartheta$ is irrational. Let the continued fraction expansion of $\vartheta$ be given by

$$
\vartheta=\left[a_{0}, a_{1}, a_{2}, \ldots .\right],
$$

and let the convergents $p_{n} / q_{n}$ for $n=0,1,2$, .. be defined by

$$
\begin{cases}p_{-1}=1, & p_{0}=a_{0}, \\ p_{n+1}=a_{n+1} \cdot p_{n}+p_{n-1} \\ q_{-1}=0, & q_{0}=1,\end{cases}
$$

It is well known that the convergents satisfy the inequalities

$$
\begin{equation*}
\frac{1}{\left(a_{n+1}+2\right) \cdot q_{n}^{2}}<\left|\vartheta-\frac{p_{n}}{q_{n}}\right|<\frac{1}{a_{n+1} \cdot q_{n}^{2}} \tag{3.5}
\end{equation*}
$$

and that if $p / q$ satisfies the inequality

$$
\begin{equation*}
\left|\vartheta-\frac{p}{q}\right|<\frac{1}{2 \cdot q^{2}} \tag{3.6}
\end{equation*}
$$

then p/q must be one of the convergents (cf. Hardy and Wright [1979], Theorems 163, 171 and 184).

We may assume without loss of generality that $\left|\vartheta_{1}\right|<\left|\vartheta_{2}\right|$, that $x_{1}>0$, and that $\left(x_{1}, x_{2}\right)=1$. From (3.1) it follows that there exists a number $X^{*}$ such that $X \geqslant X^{*}$ implies $X=x_{1}$ and (3.6) for $(p, q)=\left(-x_{2}, x_{1}\right)$. We now have the following criteria.

LEMMA 3.1. (i). If (3.1) and (3.2) hold for $\mathrm{x}_{1}, \mathrm{x}_{2}$ with $\mathrm{X} \geqslant \mathrm{X}^{*}$, then $\left(-x_{2}, x_{1}\right)=\left(p_{k}, q_{k}\right)$ for an index $k$ that satisfies

$$
\begin{equation*}
k \leqslant-1+\log \left(\sqrt{5} \cdot X_{0}+1\right) / \log \left(\frac{1}{2}(1+\sqrt{5})\right) \tag{3.7}
\end{equation*}
$$

Moreover, the partial quotient $\mathrm{a}_{\mathrm{k}+1}$ satisfies

$$
\begin{equation*}
a_{k+1}>-2+\left|\vartheta_{2}\right| \cdot c^{-1} \cdot \exp \left(\delta \cdot q_{k}\right) / q_{k} \tag{3.8}
\end{equation*}
$$

(ii). If for some k with $\mathrm{q}_{\mathrm{k}} \geqslant \mathrm{X}^{*}$

$$
\begin{equation*}
a_{k+1}>\left|\vartheta_{2}\right| \cdot c^{-1} \cdot \exp \left(\delta \cdot q_{k}\right) / q_{k} \tag{3.9}
\end{equation*}
$$

then (3.1) holds for $\left(-x_{2}, x_{1}\right)=\left(p_{k}, q_{k}\right)$.

Proof. (i). By $X \geqslant X^{*}$ and (3.6) it follows that $\left(-x_{2}, x_{1}\right)=\left(p_{k}, q_{k}\right)$ for an index $k$. Since $q_{k}$ is at least the ( $k+1$ ) th Fibonacci number, (3.7) follows from $q_{k}=x_{1}=X \leqslant X_{0}$. To prove (3.8), apply (3.1) and the first inequality of (3.5).
(ii). Combine (3.9) with the second inequality of (3.5).

We may apply Lemma 3.1(i) directly, or as follows.

LEMMA 3.2. Let

$$
A=\max \left(a_{k+1}\right)
$$

where the maximum is taken over all indices $k$ satisfying (3.7). If (3.1) and (3.2) hold for $\mathrm{x}_{1}, \mathrm{x}_{2}$ with $\mathrm{X} \geqslant \mathrm{X}_{1}$, then

$$
X<\frac{1}{\delta} \cdot \log \left(c \cdot(A+2) /\left|\vartheta_{2}\right|\right)+\frac{1}{\delta} \cdot \log X .
$$

Remark. From Lemma 3.2 an upper bound for $X$ follows. We can apply Lemma 2. 1 here, but Lemma 2.1 is sharp for large $b$ only.

Proof. (3.1) and (3.5) yield

$$
\left(a_{n+1}+2\right) \cdot q_{n}^{2}>q_{n} \cdot\left|\vartheta_{2}\right| /|\Lambda|>q_{n} \cdot\left|\vartheta_{2}\right| \cdot c^{-1} \cdot \exp (\delta \cdot X)
$$

The result follows by applying Lemma 3.1(i).

In practice it does not often occur that $A$ is large. Therefore this lemma is useful indeed.

Summarizing, this case comes down to computing the continued fraction of a real number to a certain precision, and establishing that it has no extremely large partial quotients. This idea has been applied in practice by Ellison [1971 ${ }^{\text {b }], ~ b y ~ C i j s o u w, ~ K o r l a a r ~ a n d ~ T i j d e m a n ~(a p p e n d i x ~ t o ~ S t r o e k e r ~ a n d ~ T i j d e m a n ~}$ [1982]), and by Hunt and van der Poorten (unpublished) for solving diophantine equations, by Steiner [1977] in connection with the Syracuse (" $3 \cdot N+1 "$ ) problem, and by Cherubini and Walliser [1987] (using a small home computer only) for determining all imaginary quadratic number fields with class number 1. We shall use it in Chapters 4 and 5.

### 3.3. Inhomogeneous one-dimensional approximation in the real case: the Davenport lemma.

The next case is when $\Lambda$ has the form

$$
\Lambda=\beta+x_{1} \cdot \vartheta_{1}+x_{2} \cdot \vartheta_{2},
$$

where $\beta \neq 0$. We then may use the so-called Davenport lemma, which was introduced by Baker and Davenport [1969]. It is, like the homogeneous case, based on the continued fraction algorithm.

Put again $\vartheta=-\vartheta_{1} / \vartheta_{2}$, and put $\psi=\beta / \vartheta_{2}$. Then we have

$$
\frac{\Lambda}{\vartheta_{2}}=\psi-\mathrm{x}_{1} \cdot \vartheta+\mathrm{x}_{2} .
$$

Let $p / q$ be a convergent of $\vartheta$ with $q>X_{0}$. We have the following result.

LEMMA 3.3. (Davenport). Suppose that, in the above notation,

$$
\begin{equation*}
\|q \cdot \psi\|>2 \cdot X_{0} / q \tag{3.10}
\end{equation*}
$$

(by II•ll we denote the distance to the nearest integer). Then the solutions of (3.1), (3.2) satisfy

$$
\begin{equation*}
X<\frac{1}{\delta} \cdot \log \left(q^{2} \cdot c /\left|\vartheta_{2}\right| \cdot X_{0}\right) \tag{3.11}
\end{equation*}
$$

Proof. From (3.5) and (3.10) we infer

$$
2 \cdot \mathrm{X}_{0} / \mathrm{q}<\left\|q \cdot\left(\psi-\mathrm{x}_{1} \cdot \vartheta+\mathrm{x}_{2}\right)+\mathrm{x}_{1} \cdot(\mathrm{q} \cdot \vartheta-\mathrm{p})\right\|<\mathrm{q} \cdot\left|\Lambda / \vartheta_{2}\right|+\left|\mathrm{x}_{1}\right| / \mathrm{q} .
$$

By (3.1), (3.2), and

$$
X_{0}<q^{2} \cdot c \cdot\left|\vartheta_{2}^{-1}\right| \cdot \exp (-\delta \cdot X)
$$

this leads to (3.11).

If (3.10) is not true for the first convergent with denominator $>X_{0}$, then one should try some further convergents. If $q$ is not essentially larger than $X_{0}$, then (3.11) yields a reduced upper bound for $X$ of size $\log X_{0}$, as desired. If no $q$ of the size of $X_{0}$ can be found that also satisfies (3.10) (a situation which is very unlikely to occur, as experiments show), then not all is lost, since then only very few exceptional possible solutions have to be checked. See Baker and Davenport [1969] for details.

Summarizing, we see that in this case the essential idea is that an extremely large solution of (3.1) and (3.2) leads to a large range of convergents $\mathrm{p} / \mathrm{q}$ of $\vartheta$ for which the values of $\|q \cdot \psi\|$ are all extremely small. In practice it appears to be the case that $q \cdot \psi$ is always far enough from the nearest
integer (the values of $\|q \cdot \psi\|$ seem to be distributed randomly over the interval $[0,0.5]$ ). This method has been used in practice by Baker and Davenport [1969] as we already mentioned, by Ellison, Ellison, Pesek, Stahl and Stall [1972], by Steiner [1986], and by Gáa [1988]. We shall use it in Chapter 4. Note that the method that we develop in Section 3.8 for the multi-dimensional inhomogeneous case, can be used in the one-dimensional case as well, as has been demonstrated in de Weger $\left[1989^{\mathrm{b}}\right.$ ].

### 3.4. The $\mathrm{L}^{3}$-lattice basis reduction algorithm, theory.

To deal with linear forms with $n \geqslant 3$, a straightforward generalization of the case $\mathrm{n}=2$ would be to study multi-dimensional continued fractions. For a good survey of this field, see Brentjes [1981]. However, the available algorithms in this field seem not to have the desired efficiency and generality. Fortunately, since 1981 there is a useful alternative, which in a sense is also a generalization of the one-dimensional continued fraction algorithm.

In 1981, L. Lovász invented an algorithm, that has since then become known as the $\mathrm{L}^{3}$-algorithm. It has been published in Lenstra, Lenstra and Lovász [1982], Fig. 1, p. 521. Throughout this and the next section we refer to this paper as " $\Psi \not \pm "$. The algorithm computes from an arbitrary basis of a lattice in $\mathbb{R}^{n}$ another basis of this lattice, a so-called reduced basis, which has certain nice properties (its vectors are nearly orthogonal).

The algorithm has many important applications in a variety of mathematical fields, such as the factorization of polynomials ( $\mathscr{L} \mathscr{L}$, Lenstra [1984]), public-key cryptography (Lagarias and Odlyzko [1985]), and the disproof of the Mertens Conjecture (Odlyzko and te Riele [1985]). Of interest to us are its applications to diophantine approximation, which already had been noticed in $\mathscr{L} \mathscr{L}, \quad \mathrm{p} .525$. The algorithm has a very good theoretical complexity (polynomial-time in the length of the input parameters), and performs also very well in practical computations.

Let $\Gamma \subset \mathbb{R}^{\mathrm{n}}$ be a lattice, that is given by the basis $\underline{b}_{1}, \ldots, \underline{b}_{\mathrm{n}}$. We introduce the concept of a reduced basis of $\Gamma$, according to $\mathscr{L} \mathscr{L}$, p. 516. The vectors $\underline{b}_{i}^{*}$ (for $i=1, \ldots, n$ ) and the real numbers $\mu_{i, j}$ (for $1 \leqslant j<i \leqslant n)$ are inductively defined by

$$
\underline{b}_{i}^{*}=\underline{b}_{i}-\sum_{j=1}^{i-1} \mu_{i, j} \cdot \underline{b}_{j}^{*}, \quad \mu_{i, j}=\left(\underline{b}_{i}, \underline{b}_{j}^{*}\right) /\left(\underline{b}_{j}^{*}, \underline{b}_{j}^{*}\right) .
$$

Then $\underline{b}_{1}^{*}$, ..., $\underline{b}_{n}^{*}$ is an orthogonal basis of $\mathbb{R}^{n}$. We call the lattice basis $\underline{b}_{1}, \ldots, \underline{b}_{n}$ of $\Gamma$ reduced if

$$
\begin{aligned}
& \left|\mu_{i, j}\right| \leqslant \frac{1}{2} \text { for } 1 \leqslant j<i \leqslant n \\
& \left|\underline{b}_{i}^{*}+\mu_{i, i-1} \cdot \underline{b}_{i-1}^{*}\right|^{2} \geqslant \frac{3}{4} \cdot\left|\underline{b}_{i-1}^{*}\right|^{2} \text { for } 1<i \leqslant n
\end{aligned}
$$

Hence a reduced basis is nearly orthogonal. For a reduced basis $\underline{b}_{1}$, ..., $\underline{b}_{n}$ we have, by $\mathscr{L} \mathscr{L}$ (1.7),

$$
\begin{equation*}
\left|\underline{b}_{i}^{*}\right| \geqslant 2^{-(n-1) / 2} \cdot\left|\underline{b}_{1}\right| \text { for } i=1, \ldots, n . \tag{3.12}
\end{equation*}
$$

We remark that a lattice may have more than one reduced basis, and that the ordering of the basis vectors is not arbitrary. The $L^{3}$-algorithm accepts as input any basis $\underline{b}_{1}, \ldots, \underline{b}_{n}$ of $\Gamma$, and it computes a reduced basis $\underline{c}_{1}, \ldots, \underline{c}_{n}$ of that lattice. The properties of reduced bases that are of most interest to us are the following. Let $y \in \mathbb{R}^{n}$ be a given point, that is not a lattice point. We denote by $\ell(\Gamma)$ the length of the shortest non-zero vector in the lattice, viz.

$$
\ell(\Gamma)=\min _{\underline{0} \neq \underline{x} \in \Gamma}|\underline{x}|,
$$

and by $\ell(\Gamma, \underline{y})$ the distance from $y$ to the nearest lattice point, viz.

$$
\ell(\Gamma, \underline{y})=\min _{\underline{x} \in \Gamma}|\underline{x}-\underline{y}| .
$$

From a reduced basis lower bounds for both $\ell(\Gamma)$ and $\ell(\Gamma, \underline{y})$ can be computed, according to the following results. Lemma 3.4 is Proposition (1.11) from $\mathscr{L} \mathscr{L}$. We recall its proof here, to show the similarity of the proofs of Lemma's 3.4 and 3.5.

LEMMA 3.4. (Lenstra, Lenstra and Lovasz [1982]). Let $\underline{c}_{1}, \ldots, \underline{c}_{n}$ be a reduced basis of the lattice $\Gamma$. Then

$$
\ell(\Gamma) \geqslant 2^{-(n-1) / 2} \cdot\left|\underline{c}_{1}\right|
$$

Proof. Let $\underline{0} \neq \underline{x} \in \Gamma$ be the lattice point with minimal length $|\underline{x}|=\ell(\Gamma)$. Write

$$
\underline{x}=\sum_{i=1}^{n} r_{i} \cdot \underline{c}_{i}=\sum_{i=1}^{n} r_{i}^{*} \cdot \underline{b}_{i}^{*},
$$

with $r_{i} \in \mathbb{Z}, r_{i}^{*} \in \mathbb{R}$. Let $i_{0}$ be the largest index such that $r_{i_{0}} \neq 0$. Then, since $\underline{c}_{1}, \ldots, \underline{c}_{i}$ span the same linear space as $\underline{b}_{1}^{*}, \ldots, \underline{b}_{i}^{*}$ for all $i$, and $\underline{b}_{i+1}^{*}$ is the projection of $\underline{c}_{i+1}$ on the orthogonal complement of this linear space, it follows that $r_{i_{0}}=r_{i_{0}}^{*}$. Hence, by (3.12),

$$
\begin{gathered}
\ell(\Gamma)^{2}=|\underline{x}|^{2}=\sum_{i=1}^{i_{0}} r_{i}^{* 2} \cdot\left|\underline{b}_{i}^{*}\right|^{2} \geqslant r_{i_{0}}^{* 2} \cdot\left|\underline{b}_{i}^{*}\right|^{2}=r_{i_{0}}^{2} \cdot\left|\underline{b}_{i_{0}}^{*}\right|^{2} \\
\geqslant\left|\underline{b}_{i_{0}}^{*}\right|^{2} \geqslant 2^{-(n-1)} \cdot\left|\underline{c}_{1}\right|^{2}
\end{gathered}
$$

LEMMA 3.5. Let $\underline{c}_{1}, \ldots, \underline{c}_{n}$ be a reduced basis of the lattice $\Gamma$, and let $\mathrm{y}=\sum_{\mathrm{i}=1}^{\mathrm{n}} \mathrm{s}_{\mathrm{i}} \cdot{ }^{\mathrm{C}_{\mathrm{i}}}$ for $\mathrm{s}_{1}, \ldots, \mathrm{~s}_{\mathrm{n}} \in \mathbb{R}$, with not all $\mathrm{s}_{\mathrm{i}}$ in $\mathbb{Z}$. Let $\mathrm{i}_{0}$ be the largest index such that $\mathrm{s}_{\mathrm{i}_{0}} \notin \mathbb{Z}$. Then

$$
\ell(\Gamma, \underline{y}) \geqslant 2^{-(n-1) / 2} \cdot\left\|s_{i_{0}}\right\| \cdot\left|\underline{c}_{1}\right| .
$$

Proof. Let $\underline{x} \in \Gamma$ be the lattice point nearest to $\underline{y}$. So $|\underline{x}-\underline{y}|=\ell(\Gamma, \underline{y})$. Write

$$
\underline{x}=\sum_{i=1}^{n} r_{i} \cdot \underline{c}_{i}=\sum_{i=1}^{n} r_{i}^{*} \cdot \underline{b}_{i}^{*}, \quad \underline{y}=\sum_{i=1}^{n} s_{i} \cdot \underline{c}_{i}=\sum_{i=1}^{n} s_{i}^{*} \cdot \underline{b}_{i}^{*}
$$

with $r_{i} \in \mathbb{Z}, r_{i}^{*}, s_{i}, s_{i}^{*} \in \mathbb{R}$. Let $i_{1}$ be the largest index such that $r_{i_{1}} \neq s_{i_{1}}$. Then, reasoning as in the proof of Lemma 3.4, we find

$$
r_{i_{1}}-s_{i_{1}}=r_{i_{1}}^{*}-s_{i_{1}}^{*} .
$$

Using (3.12) it follows that

$$
\ell(\Gamma, \underline{y})^{2} \geqslant\left(r_{i_{1}}-s_{i_{1}}\right)^{2} \cdot\left|\underline{b}_{i_{1}}^{*}\right|^{2} \geqslant\left(r_{i_{1}}-s_{i_{1}}\right)^{2} \cdot 2^{-(n-1)} \cdot\left|\underline{c}_{1}\right|^{2} .
$$

Obviously, $i_{1} \geqslant i_{0}$. If $i_{1}=i_{0}$ the result follows at once. If $i_{1}>i_{0}$ then $s_{i_{1}} \in \mathbb{Z}, s_{i_{1}} \neq r_{i_{1}}$, hence $\left|r_{i_{1}}{ }^{-s_{i_{1}}}\right| \geqslant 1$, and the result follows. a The above lemma is rather weak in the extraordinary situation that ${ }_{s_{i}}$ is
extremely close to an integer. If one of the other $\mathrm{s}_{\mathrm{i}}$ is not close to an integer, we can apply the following variant.

LEMMA 3.6. Let $\underline{c}_{1}, \ldots, c_{n}$ be a reduced basis of the lattice $\Gamma$, and let $y=\sum_{i=1}^{n} s_{i} \cdot \underline{c}_{i}$ for $s_{1}, \ldots, s_{n} \in \mathbb{R}$, with not all $s_{i}$ in $\mathbb{Z}$. Suppose that there is an index $\mathrm{i}_{0}$ and constants $\delta_{1}, 0<\delta_{2} \leqslant \frac{1}{2}$ such that

$$
\begin{aligned}
& \left\|\mathrm{s}_{\mathrm{i}}\right\| \leqslant \delta_{1} \text { for } \mathrm{i}=\mathrm{i}_{0}+1, \ldots, n, \\
& \mathrm{II}_{\mathrm{i}_{0}} \| \geqslant \delta_{2} .
\end{aligned}
$$

Then

$$
\ell(\Gamma, \underline{y}) \geqslant 2^{-(n-1) / 2} \cdot \delta_{2} \cdot\left|\underline{c}_{1}\right|-\left(n-i_{0}\right) \cdot \delta_{1} \cdot \max _{i>i_{0}}\left|\underline{c}_{i}\right|
$$

Proof. With notation as in the proof of Lemma 3.5, let $t_{i}$ be the integer nearest to $s_{i}$, for $i \geqslant i_{0}+1$, and $t_{i}=s_{i}$ for $i \leqslant i_{0}$. Put

$$
\underline{z}=\sum_{i=1}^{n} t_{i} \cdot \underline{c}_{i}=\sum_{i=1}^{n} t_{i}^{*} \cdot \underline{b}_{i}^{*}
$$

with $t_{i}^{*} \in \mathbb{R}$. Let $i_{1}$ be the largest index such that $r_{i_{1}} \neq t_{i_{1}}$. Then

$$
r_{i_{1}}-t_{i_{1}}=r_{i_{1}}^{*}-t_{i_{1}}^{*}
$$

We have

$$
\ell(\Gamma, \underline{y})=|\underline{x}-\underline{y}| \geqslant|\underline{x}-\underline{z}|-|\underline{z}-\underline{y}| .
$$

Now,

$$
|\underline{z}-\underline{y}| \leqslant \sum_{i=i_{0}+1}^{n}\left|s_{i}-t{ }_{i}\right| \cdot\left|\underline{c}_{i}\right| \leqslant\left(n-i_{0}\right) \cdot \delta_{1} \cdot \max _{i>i_{0}}\left|\underline{c}_{i}\right|
$$

and, using (3.12),

$$
\begin{array}{r}
|\underline{x}-\underline{z}|^{2}=\sum_{i=1}^{n}\left(r_{i}^{*}-t_{i}^{*}\right)^{2} \cdot\left|\underline{b}_{i}^{*}\right|^{2} \geqslant\left(r_{i_{1}^{*}}^{*}-t_{i_{1}}^{*}\right)^{2} \cdot\left|\underline{b}_{i}^{*}\right|^{2} \\
\geqslant\left(r_{i_{1}}-t \dot{i}_{1}\right)^{2} \cdot 2^{-(n-1)} \cdot\left|\underline{c}_{1}\right|^{2}
\end{array}
$$

Obviously, $i_{1} \geqslant i_{0}$. If $i_{1}=i_{0}$ the result follows. If $i_{1}>i_{0}$ then
$\mathrm{t}_{\mathrm{i}_{1}} \in \mathbb{Z}, \mathrm{t}_{\mathrm{i}_{1}} \neq \mathrm{r}_{\mathrm{i}_{1}}$, hence $\left|\mathrm{r}_{\mathrm{i}_{1}}-\mathrm{t}_{\mathrm{i}_{1}}\right| \geqslant 1>\delta_{2}$, and the result follows.

Remark. Babai [1986] showed that the $L^{3}$-algorithm can be used to find a lattice point $\underline{x}$ with $|\underline{x}-\underline{y}| \leqslant c \cdot \ell(\Gamma, \underline{y})$ for a constant $c$ depending on the dimension of the lattice only. This result can also be used instead of Lemma 3.5 or 3.6 .

### 3.5. The $\mathrm{L}^{3}$-lattice basis reduction algorithm, practice.

Below (in Fig. 1) we describe the variant of the $\mathrm{L}^{3}$-algorithm that we use in this monograph to solve diophantine equations. This variant has been designed to work with integers only, so that rounding-off errors are avoided completely. In the algorithm as stated in $\mathscr{L} \mathscr{L}$, Fig. 1, p. 521, non-integral rational numbers may occur, even if the input parameters are all integers.

Let $\Gamma \subset \mathbb{Z}^{n}$ be a lattice with basis vectors $\underline{b}_{1}, \ldots, \underline{b}_{n}$. Define $\underline{b}_{i}^{*}, \mu_{i j}$, $\mathrm{d}_{\mathrm{i}}$ as in $\mathscr{L} \mathscr{L}(1.2)$, (1.3), (1.24), respectively. The $\mathrm{d}_{\mathrm{i}}$ can be used as denominators for all numbers that appear in the original algorithm ( $\mathscr{L} \mathscr{L}$, p. 523). Thus, put for all relevant indices i, j

$$
\begin{align*}
& \underline{c}_{i}=d_{i-1} \cdot \underline{b}_{i}^{*}  \tag{3.13}\\
& \lambda_{i, j}=d_{j} \cdot \mu_{i, j}
\end{align*}
$$

They are integral, by $\mathscr{L L} \mathscr{L}$ (1.28), (1.29). Notice that, with $B_{i}=\left|\underline{b}_{i}^{*}\right|^{2}$,

$$
\begin{equation*}
d_{i}=d_{i-1} \cdot B_{i} \tag{3.14}
\end{equation*}
$$

We can now rewrite the algorithm in terms of $\underline{c}_{i}, d_{i}, \lambda_{i}, j$ in stead of $\underline{b}_{i}^{*}$, $B_{i}, \mu_{i, j}$, thus eliminating all non-integral rationals. We give this variant of the $L^{3}$-algorithm in Fig. 1. All the lines in this variant are evident from applying (3.13) and (3.14) to the corresponding lines in the original algorithm, except the lines (A), (B) and (C), which will be explained below.

We added a few lines to the algorithm, in order to compute the matrix of the transformation from the initial to the reduced basis. Let $\mathcal{B}$ be the matrix with column vectors $\underline{b}_{1}, \ldots, \underline{b}_{n}$, the initial basis of the lattice $\Gamma$, which is the input for the algorithm. We say: $\mathcal{B}$ is the matrix associated to the basis $\underline{b}_{1}, \ldots, \underline{b}_{n}$. Let $\mathscr{C}$ be the matrix associated to the reduced

$$
\begin{aligned}
& \mathrm{d}_{0}:=1 \text {; } \\
& \underline{c}_{\mathrm{i}}:=\underline{b}_{\mathrm{i}} \text {; } \\
& \lambda_{i, j}:=\left(\underline{b}_{i}, \underline{c}_{j}\right) ; \\
& \text { (A) } \quad \quad_{i}:=\left(d_{j} \cdot \underline{c}_{i}-\lambda_{i, j} \cdot \subseteq_{j}\right) / d_{j-1} \\
& d_{i}:=\left(\underline{c}_{i}, \underline{c}_{i}\right) / d_{i-1} \\
& \text { k := } 2 \text {; } \\
& \text { (1) perform (*) for } \ell=k-1 \text {; } \\
& \text { if } 4 \cdot d_{k-2} \cdot d_{k}<3 \cdot d_{k-1}^{2}-4 \cdot \lambda_{k, k-1}^{2} \text { go to (2); } \\
& \text { perform (*) for } \ell=k-2, \ldots, 1 \text {; } \\
& \text { if } \mathrm{k}=\mathrm{n} \text { terminate ; } \\
& \mathrm{k}:=\mathrm{k}+1 \text {; go to (1) ; } \\
& \text { (2) }\binom{\underline{b}_{k-1}}{\underline{b}_{k}}:=\binom{\underline{b}_{k}}{\underline{b}_{k-1}} \text {; } \\
& \binom{\underline{u}_{k-1}}{\underline{u}_{k}}:=\binom{\underline{u}_{k}}{\underline{u}_{k-1}} ; \quad\binom{\underline{v}_{\mathrm{k}-1}^{, T}}{\underline{v}_{k}^{, T}}:=\binom{\underline{v}_{k}^{, T}}{\underline{v}_{k}^{, T}} ; \\
& \binom{\lambda_{k-1, j}}{\lambda_{k, j}}:=\binom{\lambda_{k, j}}{\lambda_{k-1, j}} \text { for } j=1, \ldots, k-2 \text {; } \\
& \text { (B) } \quad\binom{\lambda_{i, k-1}}{\lambda_{i, k}}:=\left(\lambda_{i, k-1} \cdot\binom{\lambda_{k, k-1}}{d_{k}}+\lambda_{i, k} \cdot\binom{d_{k-2}}{-\lambda_{k, k-1}}\right) / d_{k-1} \\
& \text { for } i=k+1, \ldots, n \text {; } \\
& \text { (C) } \quad d_{k-1}:=\left(d_{k-2} \cdot d_{k}+\lambda_{k, k-1}^{2}\right) / d_{k-1} \text {; } \\
& \text { if } k>2 \text { then } k:=k-1 \text {; } \\
& \text { go to (1) ; } \\
& \text { (*) if } 2 \cdot\left|\lambda_{\mathrm{k}, \ell}\right|>\mathrm{d}_{\ell} \text { then } \\
& \left\{\begin{array}{l}
\mathrm{r}:=\text { integer nearest to } \lambda_{\mathrm{k}, \ell} / \mathrm{d}_{\ell} ; \\
\underline{b}_{\mathrm{k}}:=\underline{b}_{k}-\mathrm{r} \cdot \underline{b}_{\ell} ; \quad \underline{u}_{\mathrm{k}}:=\underline{u}_{\mathrm{k}}-\mathrm{r} \cdot \underline{u}_{\ell} ; \underline{\mathrm{v}}_{\ell}^{, T}:=\underline{v}_{\ell}^{, T}+\mathrm{r} \cdot \underline{v}_{\mathrm{k}}^{, \mathrm{T}} ; \\
\lambda_{\mathrm{k}, \mathrm{j}}:=\lambda_{\mathrm{k}, \mathrm{j}}-\mathrm{r} \cdot \lambda_{\ell, j} \text { for } \mathrm{j}=1, \ldots, \ell-1 ; \\
\lambda_{\mathrm{k}, \ell}:=\lambda_{\mathrm{k}, \ell}-\mathrm{r} \cdot \mathrm{~d}_{\ell} .
\end{array}\right.
\end{aligned}
$$

Figure 1. Variant of the $\mathrm{L}^{3}$-algorithm.
basis $\underline{c}_{1}$, ..., $\underline{c}_{n}$, which the algorithm delivers as output. Then we define this transformation matrix $V$ by

$$
\mathscr{G}=\mathcal{B} \cdot V
$$

More generally, let $u$ be the matrix of a transformation from some $\mathcal{B}_{0}$ to $\mathcal{B}$, so $\mathcal{B}=\mathscr{B}_{0} \cdot U$. Denote the column vectors of $U$ by $\underline{u}_{1}, \ldots, \underline{u}_{n}$, and the row vectors of $U^{-1}$ by $\underline{v}_{1}^{, T}, \ldots, \underline{v}_{n}^{, T}$. We feed the algorithm with $U$ and $U^{-1}$ as well. All manipulations in the algorithm done on the $\underline{b}_{i}$ are also done on the $\underline{u}_{i}$, and the $\underline{v}_{\mathbf{i}}, T$ are adjusted accordingly. This does not affect the computation time seriously. The algorithm now gives as output matrices $\mathscr{C}, U^{\prime}$ and $U^{,-1}$, such that $\mathscr{C}$ is associated to a reduced basis, $\mathscr{C}=\mathscr{B} \cdot \mathcal{V}$, and $\mathcal{U}^{\prime}=\mathcal{U} \cdot \mathcal{V}$. Note that $\mathcal{V}$ is not computed explicitly, unless $U=\mathcal{G}$ (the unit matrix), in which case $U^{\prime}=V$. It follows that

$$
\mathscr{C}=\mathcal{B} \cdot u^{-1} \cdot U^{\prime}=\mathcal{B}_{0} \cdot U^{\prime}
$$

so $U$, is the matrix of the transformation from $\mathcal{B}_{0}$ to $\mathscr{C}^{\text {G }}$. Note that if $\mathcal{B}_{0}^{-1}$ is known, then it is not much extra effort to compute $\mathscr{G}^{-1}$ as well.

We now explain why lines (A), (B) and (C) are correct.
(A): From $\mathscr{L} \mathscr{L} \mathscr{L}$ (1.2) it follows that

$$
\underline{c}_{i}=d_{i-1} \cdot \underline{b}_{i}-\sum_{k=1}^{i-1} \frac{d_{i-1}}{d_{k-1} \cdot d_{k}} \cdot \lambda_{i, k} \cdot \underline{c}_{k} .
$$

Define for $j=0,1, \ldots, i-1$

$$
\underline{c}_{i}(j)=d_{j} \cdot \underline{b}_{i}-\sum_{k=1}^{j} \frac{d_{j}}{d_{k-1} \cdot d_{k}} \cdot \lambda_{i, k} \cdot \underline{c}_{k} .
$$

Then $\underline{c}_{\mathbf{i}}(0)=\underline{b}_{\mathbf{i}}$, and $\underline{c}_{\mathbf{i}}(\mathbf{i}-1)=\underline{c}_{\mathbf{i}}$. The $\underline{c}_{\mathbf{i}}(\mathbf{j})$ is exactly the vector computed in (A) at the $j$ th step, since

$$
\begin{aligned}
& \frac{d_{j} \cdot \underline{c}_{i}(j-1)-\lambda_{i, j} \cdot \underline{c}_{j}}{d_{j-1}} \\
& \quad=d_{j} \cdot \underline{b}_{i}-\sum_{k=1}^{j-1} \frac{d_{j}}{d_{k-1} \cdot d_{k}} \cdot \lambda_{i, k} \cdot \underline{c}_{k}-\frac{d_{j}}{d_{j-1} \cdot d_{j}} \cdot \lambda_{i, j} \cdot \underline{c}_{j}=\underline{c}_{i}(j) .
\end{aligned}
$$

This explains the recursive formula in line (A). It remains to show that the occurring vectors $\underline{c}_{i}(j)$ are integral. This follows from

$$
d_{j} \cdot \sum_{k=1}^{j} \frac{1}{d_{k-1} \cdot d_{k}} \cdot \lambda_{i, k} \cdot \underline{c}_{k}=d_{j} \cdot \sum_{k=1}^{j} \mu_{i, k} \cdot \underline{b}_{k}^{*},
$$

which is integral by $\mathscr{L L} \mathscr{L}$ p. $523, \ell .11$.
(B), (C): Notice that the third and fourth line, starting from label (2), in the original algorithm, are independent of the first, second and fifth line. Thus a permutation of these lines is allowed. We rewrite the first, second and fifth line as follows (where we indicate variables that have been changed with a prime sign):

$$
\begin{align*}
& \mathrm{B}_{\mathrm{k}-1}^{\prime}:=\mathrm{B}_{\mathrm{k}}+\mu_{\mathrm{k}, \mathrm{k}-1}^{2} \cdot \mathrm{~B}_{\mathrm{k}-1}  \tag{3.15}\\
& \mathrm{~B}_{\mathrm{k}}^{\prime}:=\mathrm{B}_{\mathrm{k}-1} \cdot \mathrm{~B}_{\mathrm{k}} / \mathrm{B}_{\mathrm{k}-1}^{\prime} ;  \tag{3.16}\\
& \mu_{\mathrm{k}, \mathrm{k}-1}^{\prime}:=\mu_{\mathrm{k}, \mathrm{k}-1} \cdot \mathrm{~B}_{\mathrm{k}-1} / \mathrm{B}_{\mathrm{k}-1}^{\prime}  \tag{3.17}\\
& \mu_{\mathrm{i}, \mathrm{k}-1}^{\prime}:=\mu_{\mathrm{k}, \mathrm{k}-1}^{\prime} \cdot \mu_{\mathrm{i}, \mathrm{k}-1}+\left(1-\mu_{\mathrm{k}, \mathrm{k}-1} \cdot \mu_{\mathrm{k}, \mathrm{k}-1}^{\prime}\right) \cdot \mu_{\mathrm{i}, \mathrm{k}}  \tag{3.18}\\
& \mu_{\mathrm{i}, \mathrm{k}}^{\prime}:=\mu_{\mathrm{i}, \mathrm{k}-1}-\mu_{\mathrm{k}, \mathrm{k}-1} \cdot \mu_{\mathrm{i}, \mathrm{k}} \tag{3.19}
\end{align*}
$$

where (3.18) and (3.19) hold for $i=k+1$, ..., $n$. The $d_{i}$ remain unchanged for $i=0,1, \ldots, k-2$, and by (3.16) also for $i=k$. Now, (3.15) is equivalent to

$$
\begin{equation*}
\frac{d_{k-1}^{\prime}}{d_{k-2}}=\frac{d_{k}}{d_{k-1}}+\frac{\lambda_{k, k-1}^{2}}{d_{k-1}^{2}} \cdot \frac{d_{k-1}}{d_{k-2}} \tag{3.20}
\end{equation*}
$$

which explains (C). From (3.17) we find

$$
\frac{\lambda_{\mathrm{k}, \mathrm{k}-1}^{\prime}}{\mathrm{d}_{\mathrm{k}-1}^{\prime}}=\frac{\lambda_{\mathrm{k}, \mathrm{k}-1}}{\mathrm{~d}_{\mathrm{k}-1}} \cdot \frac{\mathrm{~d}_{\mathrm{k}-1}}{\mathrm{~d}_{\mathrm{k}-2}} \cdot \frac{\mathrm{~d}_{\mathrm{k}-2}^{\prime}}{\mathrm{d}_{\mathrm{k}-1}^{\prime}}
$$

hence $\lambda_{k, k-1}$ remains unchanged. From (3.18) we obtain

$$
\frac{\lambda_{i, k-1}^{\prime}}{d_{k-1}^{\prime}}=\frac{\lambda_{k, k-1}}{d_{k-1}^{\prime}} \cdot \frac{\lambda_{i, k-1}}{d_{k-1}}+\left(1-\frac{\lambda_{k, k-1}}{d_{k-1}} \cdot \frac{\lambda_{k, k-1}}{d_{k-1}^{\prime}}\right) \cdot \frac{\lambda_{i, k}}{d_{k}}
$$

whence, by multiplying by $\mathrm{d}_{\mathrm{k}-1} \cdot \mathrm{~d}_{\mathrm{k}-1}$ and using (3.20),

$$
\begin{gathered}
d_{k-1} \cdot \lambda_{i, k-1}^{\prime}=\lambda_{k, k-1} \cdot \lambda_{i, k-1}+\left(d_{k-1} \cdot d_{k-1}^{\prime}-\lambda_{k, k-1}^{2}\right) \cdot \frac{\lambda_{i, k}}{d_{k}} \\
=\lambda_{k, k-1} \cdot \lambda_{i, k-1}+d_{k-2} \cdot \lambda_{i, k}
\end{gathered}
$$

Finally, from (3.19) we see

$$
\frac{\lambda_{i, k}^{\prime}}{d_{k}}=\frac{\lambda_{i, k-1}}{d_{k-1}}-\frac{\lambda_{k, k-1}}{d_{k-1}} \cdot \frac{\lambda_{i, k}}{d_{k}},
$$

and (B) follows.

In our applications we often have a lattice $\Gamma$, of which a basis is given such that the associated matrix, $\mathcal{A}$ say, has the special form

$$
\mathscr{A}=\left(\begin{array}{cccc}
1 & & & \\
& \ddots & \varnothing & \\
& \varnothing & \cdot & 1 \\
& & 1 & \\
\Theta_{1} & \cdots & \Theta_{\mathrm{n}-1} & \Theta_{\mathrm{n}}
\end{array}\right)
$$

where the $\Theta_{i}$ are large integers, that may have several hundreds of decimal digits. We can compute a reduced basis of this lattice directly, using the matrix $\mathcal{A}$ itself as input for the $L^{3}$-algorithm. But it may save time and space to split up the computation into several steps with increasing accuracy, as follows.

Let $k$ be a natural number (the number of steps), and let $\ell$ be a natural number such that the $\Theta_{i}$ have about $k \cdot \ell$ (decimal) digits. For $\mathrm{i}=1, \ldots, \mathrm{n}$ and $\mathrm{j}=1, \ldots, \mathrm{k}$ put

$$
\Theta_{i}^{(j)}=\left[\Theta_{i} / 10^{\ell \cdot(k-j)}\right]
$$

and define $\Psi_{i}^{(j)}$ by

$$
\Theta_{i}^{(j+1)}=10^{\ell} \cdot \Theta_{i}^{(j)}+\Psi_{i}^{(j)} .
$$

Thus, the $\Psi_{i}^{(j)}$ are blocks of $\ell$ consecutive digits of $\Theta_{i}$. Define for the relevant $j$ the $n \times n$ matrices

$$
\begin{aligned}
& \mathcal{E}=\left(\begin{array}{llll}
1 & & & \\
& \ddots & & \varnothing \\
& & \ddots & 1 \\
\\
& & & \\
& & & \\
\ell
\end{array}\right) .
\end{aligned}
$$

Then it follows at once that

$$
A_{j+1}=\mathcal{E} \cdot A_{j}+\mathcal{D}_{j}
$$

Notice that $\mathcal{A}_{\mathrm{k}}=\mathcal{A}$, since $\Theta_{\mathrm{i}}^{(\mathrm{k})}=\Theta_{\mathrm{i}}$. Put $u_{0}=\mathcal{g}, \quad \mathcal{B}_{1}=\mathcal{A}_{1}$. For some $j \geqslant 1$ let $\mathcal{B}_{j}$ and $\mathcal{U}_{j-1}$ be known matrices. Then we apply the $\mathrm{L}^{3}$-algorithm to $\mathcal{B}=\mathcal{B}_{j}, \quad u=u_{j-1}$, and $u^{-1}$. We thus find matrices $\mathscr{G}_{j}, u_{j}$, and $u_{j}{ }^{-1}$ such that

$$
\mathscr{C}_{j}=\mathcal{B}_{j} \cdot U_{j-1}^{-1} \cdot U_{j}
$$

Now put

$$
\mathcal{B}_{j+1}=\mathscr{E}_{j} \cdot \mathscr{C}_{j}+\mathcal{D}_{j} \cdot U_{j}
$$

By induction $\mathcal{B}_{\mathbf{j}}, \mathscr{C}_{\mathbf{j}}$ and $\mathcal{U}_{\mathbf{j}}$ are defined for $\mathrm{j}=1, \ldots, \mathrm{k}$. Note that

$$
\mathcal{B}_{j+1} \cdot u_{j}^{-1}=\mathcal{E} \cdot \mathcal{B G}_{j} \cdot u_{j-1}^{-1}+\mathcal{D}_{j}
$$

so the $\mathcal{B}_{j} \cdot U_{j-1}^{-1}$ satisfy the same recursive relation as the $\mathscr{A}_{j}$. Since $\mathcal{B}_{1} \cdot U_{0}^{-1}=\mathcal{A}_{1}$, we have $\mathcal{B}_{j} \cdot U_{j-1}^{-1}=\mathcal{A}_{j}$ for all $j$. Hence

$$
\mathscr{C}_{j}=\mathcal{B}_{j} \cdot u_{j-1}^{-1} \cdot u_{j}=A_{j} \cdot u_{j}
$$

and it follows that $\mathscr{C}_{k}$ and $\mathscr{A}_{k}$ are associated to bases of the same lattice, which is $\Gamma$. Moreover, since $\mathscr{C}_{k}$ is output of the $L^{3}$-algorithm, it is associated to a reduced basis of $\Gamma$.

Let us now analyse the computation time. For a matrix $M$ we denote by $L(M)$ the maximal number of (decimal) digits of its entries. If the $L^{3}$-algorithm is applied to a matrix $\mathcal{B}$, with as output a matrix $\mathscr{C}$, then according to the experiences of Lenstra, Odlyzko (cf. Lenstra [1984], p. 7) and ourselves, the computation time is proportional to $L(\mathscr{B})^{3}$ in practice. Since $\mathscr{C}$ is associated to a reduced basis, we assume that

$$
\mathrm{L}(\mathscr{C}) \cong{ }^{10} \log (\operatorname{det} \Gamma) / \mathrm{n}
$$

In our situation, $L\left(\mathscr{A}_{j}\right) \cong \ell \cdot j, L\left(\mathcal{D}_{j}\right) \cong \ell$, and by $\operatorname{det} \mathscr{G}_{j}=\operatorname{det} \mathscr{A}_{j}=\Theta_{\mathrm{n}}^{(\mathrm{j})}$ we have $L\left(\mathscr{C}_{j}\right) \cong \ell \cdot j / n$. Put $\mathscr{C}_{j}=\left(c_{i, h}^{(j)}\right), \quad u_{j}=\left(u_{i, h}^{(j)}\right)$. Then by $\mathscr{G}_{j}=\mathcal{A}_{j} \cdot U_{j}$ and the special shape of $\mathcal{A}_{j}$ we have $c_{i, h}^{(j)}=u_{i, h}^{(j)}$ for $\mathrm{i}=1, \ldots, \mathrm{n}-1$ and $\mathrm{h}=1, \ldots, \mathrm{n}$, and

$$
u_{n, h}^{(j)}=\left(-c_{1, h}^{(j)} \cdot \Theta_{1}^{(j)}-\ldots-c_{n-1, h}^{(j)} \cdot \Theta_{n-1}^{(j)}+c_{n, h}^{(j)}\right) / \Theta_{n}^{(j)}
$$

It follows that $L\left(U_{j}\right) \cong L\left(\mathscr{C}_{j}\right)$. So

$$
L\left(\mathcal{B}_{j}\right) \cong \max \left(L\left(\mathscr{E}_{j-1}\right), L\left(\mathcal{D}_{j-1} \cdot U_{j-1}\right)\right) \cong \ell+\ell \cdot(j-1) / n
$$

Instead of applying the $L^{3}$-algorithm once with $\mathscr{A}$ as input, we apply it $k$ times, with $\mathcal{B}_{1}, \ldots, \mathcal{B}_{k}$ as input. Thus we reduce the computation time by a factor

$$
\frac{L(\mathscr{A})^{3}}{\sum_{j=1}^{k} L\left(\mathcal{B}_{j}\right)^{3}} \cong \frac{(\ell \cdot k)^{3}}{\sum_{j=1}^{k} \ell^{3} \cdot\left(1+\frac{j-1}{n}\right)^{3}}=\frac{k^{3} \cdot n^{3}}{\sum_{j=0}^{k-1}(n+j)^{3}}
$$

For $k$ between $2.5 \cdot n$ and $3 \cdot n$ this expression is maximal, about $0.4 \cdot n^{2}$. So the reduction in computation time is considerable (a factor 10 already for $\mathrm{n}=5$ ). The storage space that is required is also reduced, since the largest numbers that appear in the input have $\ell \cdot(1+(k-1) / n)$ instead of $\ell \cdot k$ digits.

### 3.6. Finding all short lattice points: the Fincke and Pohst algorithm.

Sometimes it is not sufficient to have only a lower bound for $\ell(\Gamma)$ or $\ell(\Gamma, y)$. It may be useful to know exactly all vectors $\underline{x} \in \Gamma$ such that $|\underline{x}| \leqslant C$ or $|\underline{x}-\underline{y}| \leqslant C$ for a given constant $C$. There exists an efficient algorithm for finding all the solutions to these problems. This algorithm was devised by Fincke and Pohst [1985], cf. their (2.8) and (2.12). We give a description of this algorithm below.

The input of the algorithm is a matrix $\mathcal{B}$ whose column vectors span the lattice $\Gamma$, and a constant $C>0$. The output is a list of all lattice points $\underline{x} \in \Gamma$ with $|\underline{x}| \leqslant C$, apart from $\underline{x}=\underline{0}$. We give the algorithm in Figure 2. We use the notation $X=\left(\mathrm{X}_{\mathrm{ij}}\right)$ for matrices $X=\mathscr{A}, \mathcal{B}, \mathcal{R}, \mathscr{P}, \mathcal{U}$, and $\underline{x}_{i}$ for the column vectors of $X$.

The algorithm can also be used for finding all vectors $\underline{x} \in \Gamma$ of which the distance to a given non-lattice point $Z$ is at most a given constant $C$. Namely, let

$$
y=\sum_{i=1}^{n} s_{i} \cdot \underline{b}_{i}
$$

and let $r_{i}$ be the integer nearest to $s_{i}$ for all $i$. Put
$\mathcal{A}:=\mathcal{B}^{\mathrm{T}} \cdot \mathcal{B}$;
$q_{i j}:=a_{i j}$ for $1 \leqslant i \leqslant j \leqslant n ;$
$q_{j i}:=q_{i j}, \quad q_{i j}:=q_{i j} / q_{i i}$ for $1 \leqslant i<j \leqslant n ;$
$q_{k \ell}:=q_{k \ell}-q_{k i} \cdot q_{i \ell}$ for $i+1 \leqslant k \leqslant \ell \leqslant n$ for $1 \leqslant i \leqslant n$;
$r_{i i}:=V_{q_{i i}}$ for $1 \leqslant i \leqslant n$;
$r_{i j}:=r_{i i_{-1}} q_{i j}, \quad r_{j i}:=0$ for $1 \leqslant j<i \leqslant n ;$
compute $\mathcal{R}^{-1}$;
compute a row-reduced version $\mathscr{S}^{-1}$ of $\mathcal{R}^{-1}$, and $U, U^{-1}$ such that $\mathscr{\varphi}^{-1}=U^{-1} \cdot \mathcal{R}^{-1}$;
compute $\mathscr{S}=\mathcal{R} \cdot U 4$;
determine a permutation $\pi$ such that $\left|\underline{s}_{\pi(1)}\right| \geqslant \ldots \geqslant \underline{s}_{\pi(n)} \mid$, let $\varphi$, be the matrix with columns $\underline{s}_{\pi}^{-1}(\mathrm{i})$ for $i=1, \ldots, n$;
$\mathscr{A}:=\varphi, \mathrm{T} \cdot \varphi, \quad ;$
$q_{i j}:=a_{i j}$ for $1 \leqslant i \leqslant j \leqslant n ;$
$q_{j i}:=q_{i j}, \quad q_{i j}:=q_{i j} / q_{i i}$ for $1 \leqslant i<j \leqslant n$;
$q_{k \ell}:=q_{k \ell}-q_{k i} \cdot q_{i \ell}$ for $i+1 \leqslant k \leqslant \ell \leqslant n$ for $1 \leqslant i \leqslant n$;
i : = n ;
$\mathrm{T}_{\mathrm{i}}:=\mathrm{C}$;
$U_{i}:=0 ;$
(1) $Z:=V\left(T_{i} / q_{i i}\right)$;
$\mathrm{UB}\left(\mathrm{x}_{\mathrm{i}}\right):=\left\lfloor\mathrm{Z}-\mathrm{U}_{\mathrm{i}}\right\rfloor$;
$\mathrm{x}_{\mathrm{i}}:=\left\lceil-\mathrm{Z}-\mathrm{U}_{\mathrm{i}}\right\rceil-1$;
(2) $x_{i}:=x_{i}+1$;
if $X_{i} \leqslant U B\left(x_{i}\right)$, go to (4) ;
(3) $\mathrm{i}:=\mathrm{i}+1$;
go to (2) ;
(4) if i $=1$, go to (5) ;
i : = i - 1 ;
$U_{i}:=\sum_{j=}^{m} q_{i+1} \cdot x_{j}$;
$T_{i}:=T_{i+1}-q_{i+1, i+1} \cdot\left(x_{i+1}+U_{i+1}\right)^{2}$;
go to (1) ;
(5) if $x_{i}=0$ for $1 \leqslant i \leqslant n$, terminate;
compute and print $\left.\underline{x}=U_{\pi}^{-1}(1), \ldots, x_{\pi^{-1}(n)}\right)^{T}$;
go to (2).

Figure 2. The Fincke and Pohst Algorithm.

$$
\underline{z}=\sum_{i=1}^{n} r_{i} \cdot \underline{b}_{i}
$$

Then $|\underline{y}-\underline{z}|<C, ~ f o r ~ s o m e ~ c o n s t a n t ~ C, ~\left(~ C ' ~=~ \frac{n}{2} \cdot \sum\left|\underline{b}_{i}\right|\right.$ will do). Since $\underline{z} \in \Gamma$ it suffices to search for all lattice points $\underline{u}$ with $|\underline{u}| \leqslant C+C^{\prime}$, and compute for each such $\underline{u}$ also $\underline{x}=\underline{z}+\underline{u}$, since $|\underline{x}-\underline{y}|<C$ implies

$$
|\underline{u}| \leqslant|\underline{x}-\underline{y}|+|\underline{y}-\underline{z}| \leqslant C+C^{\prime} .
$$

### 3.7. Homogeneous multi-dimensional approximation in the real case: real approximation lattices.

Let the linear form $\Lambda$ have the form

$$
\Lambda=\sum_{i=1}^{n} x_{i} \cdot \vartheta_{i}
$$

We assume that $n \geqslant 2$. The case $n=2$ has already been discussed in Section 3.2, but the method of this section works also for $n=2$. In fact, it is in this case essentially the same method.

Let $C$ be a large enough integer, that is of the order of magnitude of $X_{0}^{n}$. Let $\gamma \in \mathbb{N}$ be a constant (we will explain its use later). We define the approximation lattice $\Gamma$ by the matrix

$$
\mathcal{B}=\left(\begin{array}{ccccc}
\gamma & & & \\
& \cdot & \cdot & & \\
\emptyset & & & \\
{\left[\gamma \cdot \mathrm{C} \cdot \vartheta_{1}\right]} & \ldots & {\left[\gamma \cdot \mathrm{C} \cdot \vartheta_{\mathrm{n}-1}\right]} & {\left[\gamma \cdot \mathrm{C} \cdot \vartheta_{\mathrm{n}}\right]}
\end{array}\right)
$$

of which the column vectors $\underline{b}_{1}, \ldots, \underline{b}_{n}$ are a basis of the lattice. Then $\Gamma$ is a sublattice of $\mathbb{Z}^{n}$ of determinant $\gamma^{n-1} \cdot\left[\gamma \cdot C \cdot \vartheta_{n}\right]$, which is of size $C$. A lattice point $\underline{x}$ has the form

$$
\underline{x}=\sum_{i=1}^{n} x_{i} \cdot \underline{b}_{i}=\left(\gamma \cdot x_{1}, \ldots, \gamma \cdot x_{n-1}, \tilde{\Lambda}\right)^{T}
$$

where the $x_{i}$ are integers, and

$$
\tilde{\Lambda}=\sum_{i=1}^{n} x_{i} \cdot\left[\gamma \cdot C \cdot \vartheta_{i}\right]
$$

Clearly, $\tilde{\Lambda}$ is close to $\gamma \cdot C \cdot \Lambda$. The length of the vector $\underline{x}$ now measures both $X_{0}$ and $|\Lambda|$, which are exactly the two numbers we want to balance with each other. Heuristics (cf. Section 1.3) tell us that in a generic case we expect $|\Lambda| \cong X_{0}^{-n}$. We now can prove easily the following useful lemma.

LEMMA 3.7. Let $X_{1}$ be a positive number such that

$$
\begin{equation*}
\ell(\Gamma) \geqslant V\left((n+1)^{2}+(n-1) \cdot \gamma^{2}\right) \cdot X_{1} . \tag{3.21}
\end{equation*}
$$

Then (3.1) has no solutions with

$$
\begin{equation*}
\frac{1}{\delta} \cdot \log \left(\gamma \cdot C \cdot c / X_{1}\right) \leqslant X \leqslant X_{1} . \tag{3.22}
\end{equation*}
$$

Remark. We apply this lemma for $X_{1}=X_{0}$. If condition (3.21) then fails, we must take a larger constant $C$. If it holds for a constant $C$ of the size $X_{0}^{n}$, then (3.22) yields a reduced lower bound for $X$ of size $\log X_{0}$.

Proof. Let $X_{1}, \ldots, x_{n}$ be a solution of (3.1) with $0<X \leqslant X_{1}$. Consider the lattice point

$$
\underline{x}=\sum_{i=1}^{n} x_{i} \cdot \underline{b}_{i}=\left(\gamma \cdot x_{1}, \ldots, \gamma \cdot x_{n-1}, \tilde{\Lambda}\right)^{T}
$$

with $\tilde{\Lambda}$ as above. Then

$$
|\underline{x}|^{2}=\gamma^{2} \cdot \sum_{i=1}^{n-1} x_{i}^{2}+\tilde{\Lambda}^{2} \leqslant(n-1) \cdot \gamma^{2} \cdot x_{1}^{2}+\tilde{\Lambda}^{2}
$$

and

$$
\begin{equation*}
|\tilde{\Lambda}-\gamma \cdot C \cdot \Lambda| \leqslant \sum_{i=1}^{n}\left|x_{i}\right| \cdot\left|\left[\gamma \cdot C \cdot \vartheta_{i}\right]-\gamma \cdot C \cdot \vartheta_{i}\right| \leqslant \sum_{i=1}^{n}\left|x_{i}\right| \tag{3.23}
\end{equation*}
$$

which is $\leqslant n \cdot X_{1}$. By (3.1), (3.21) and the definition of $\ell(\Gamma)$ we have

$$
\begin{gathered}
\gamma \cdot C \cdot c \cdot \exp (-\delta \cdot X)>|\gamma \cdot C \cdot \Lambda| \geqslant|\tilde{\Lambda}|-|\tilde{\Lambda}-\gamma \cdot C \cdot \Lambda| \\
\geqslant V\left(\ell(\Gamma)^{2}-(n-1) \cdot \gamma^{2} \cdot X_{1}^{2}\right)-n \cdot X_{1} \geqslant X_{1},
\end{gathered}
$$

and (3.22) follows at once.

Condition (3.21) can be checked by computing a reduced basis of the lattice $\Gamma$ by the $\mathrm{L}^{3}$-algorithm, and applying Lemma 3.4. The parameter $\gamma$ is used to keep the "rounding-off error"

$$
\left|\left[\gamma \cdot C \cdot \vartheta_{i}\right]-\gamma \cdot C \cdot \vartheta_{i}\right|
$$

relatively small. This is of importance only if $C$ is not very large, usually only if one wants to make a further reduction step after the first step has already been made. For large $C$, simply take $\gamma=1$.

It may be necessary, if $C$ is not very large, to use a more refined method of reducing the upper bound. To do so, we use the following lemma, which is a slight refinement of Lemma 3.7, together with the algorithm of Fincke and Pohst (cf. Section 3.6). It is particularly useful in the situation that one has different upper bounds for the $\left|x_{i}\right|$ for different $i$.

LEMMA 3.8. Suppose that for a solution of (3.1)

$$
\begin{equation*}
|\tilde{\Lambda}|>\sum_{i=1}^{n}\left|x_{i}\right| \tag{3.24}
\end{equation*}
$$

holds. Then

$$
\begin{equation*}
X<\frac{1}{\delta} \cdot \log \left(\gamma \cdot C \cdot c /\left(|\tilde{\Lambda}|-\sum_{i=1}^{n}\left|x_{i}\right|\right)\right) . \tag{3.25}
\end{equation*}
$$

Proof. Define the lattice point $\underline{x}$ as in the proof of Lemma 3.7. By (3.23) and (3.24)

$$
|\Lambda| \geqslant\left(|\tilde{\Lambda}|-\sum_{i=1}^{n}\left|x_{i}\right|\right) / \gamma \cdot C>0
$$

The result follows at once by (3.1).

We proceed as follows. Choose a constant $C_{0}$ such that if $|\tilde{\Lambda}|>C_{0}$ then the upper bounds for $\left|x_{i}\right|$ imply (3.24). In that case we have a new upper bound for $X$ from (3.25). In case $|\tilde{\Lambda}| \leqslant C_{0}$ we have an upper bound for the length of the vector $\underline{x}$. We compute all lattice points satisfying this bound by the algorithm of Fincke and Pohst, and check them for (3.1).

Summarizing, the reduction method presented above is based on the fact that a large solution of (3.1) corresponds to an extremely short vector in an appropriate approximation lattice. Since we can actually prove by computations that such short vectors do not exist, it follows that such large solutions do not exist. We will apply these techniques in Chapter 5.
3.8. Inhomogeneous multi-dimensional approximation in the real case: an alternative for the generalized Davenport lemma.

Let $\Lambda$ be the most general linear form that we will study, viz.

$$
\Lambda=\beta+\sum_{i=1}^{n} x_{i} \cdot \vartheta_{i}
$$

where $n \geqslant 2$ (the case $n=2$ has been dealt with in Section 3.3, but can be incorporated here also). To deal with this inhomogeneous case, two methods are available. The first method is a generalization of the method of Davenport that we discussed in Section 3.3. The second method is closer to the homogeneous case of the previous section.

First we explain briefly the generalized Davenport method. See Ellison [1971 ${ }^{\text {a }] ~(w h e r e ~ o n l y ~ t h e ~ c a s e ~} n=3$ is treated). Put

$$
\begin{aligned}
& \vartheta_{i}^{\prime}=\vartheta_{i} / \vartheta_{n} \text { for } i=1, \ldots, n-1, \quad \beta^{\prime}=\beta / \vartheta_{n} \\
& \Lambda^{\prime}=\Lambda / \vartheta_{n}=\beta^{\prime}+\sum_{i=1}^{n-1} x_{i} \cdot \vartheta_{i}^{\prime}+x_{n} .
\end{aligned}
$$

Let $\left(p_{1}, \ldots, p_{n-1}, q\right)$ be a simultaneous approximation to $\vartheta_{1}^{\prime}, \ldots, \vartheta_{n-1}^{\prime}$ with $q$ of the size of $X_{0}^{n-1}$, such that, for $i=1, \ldots, n-1$,

$$
\left|\vartheta_{i}^{\prime}-p_{i} / q\right|<c^{\prime} / q^{1+1 /(n-1)}
$$

for a small constant $c$ '.

LEMMA 3.9. (Davenport, Ellison). Suppose that

$$
\left\|q \cdot \beta^{\prime}\right\|>2 \cdot(n-1) \cdot X_{0} \cdot c^{\prime} / q^{1 /(n-1)}
$$

Then the solutions of (3.1), (3.2) satisfy

$$
X<\frac{1}{\delta} \cdot \log \left(q^{1+1 /(n-1)} \cdot c /\left|\vartheta_{n}\right| \cdot c \cdot \cdot(n-1) \cdot X_{0}\right)
$$

Proof. The result follows at once from

$$
\begin{aligned}
& \left\|q \cdot \beta^{\prime}\right\| \leqslant\left|q \cdot \Lambda^{\prime}+\sum_{i=1}^{n-1} x_{i} \cdot\left(p_{i}-q \cdot \vartheta_{i}^{\prime}\right)\right| \leqslant \\
& q \cdot\left|\vartheta_{n}\right|^{-1} \cdot c \cdot \exp (-\delta \cdot X)+(n-1) \cdot X_{0} \cdot c^{\prime} / q^{1 /(n-1)} .
\end{aligned}
$$

To apply this generalized Davenport method in practice, it is necessary to compute the simultaneous approximations $\left(p_{1}, \ldots, p_{n-1}, q\right)$. We indicated in Section 1.4 how this can be done with the $L^{3}$-algorithm. As lattice we take the one associated to the following matrix:

$$
\left(\begin{array}{cccc}
1 & & & \\
{\left[C \cdot \vartheta_{1}^{\prime}\right]} & -C & & \varnothing \\
\vdots & & \ddots & \\
{\left[C \cdot \vartheta_{n-1}^{\prime}\right]} & \varnothing & & -C
\end{array}\right)
$$

where $C$ is a constant of size $X_{0}^{n}$. Then $c_{1}$, the first basis vector of a reduced basis, will have length of the size of $C^{(n-1) / n} \cong X_{0}^{n-1}$. But $\underline{c}_{1}$ can be written as

$$
\underline{c}_{1}=\left(\mathrm{q}, \mathrm{q} \cdot\left[\mathrm{C} \cdot \vartheta_{1}^{\prime}\right]-\mathrm{C} \cdot \mathrm{p}_{1}, \ldots, \quad \mathrm{q} \cdot\left[\mathrm{C} \cdot \vartheta_{\mathrm{n}-1}^{\prime}\right]-\mathrm{C} \cdot \mathrm{p}_{\mathrm{n}-1}\right)^{\mathrm{T}}
$$

for some $p_{1}, \ldots, p_{n-1}$, $q$. It is expected that $q$ is of size $X_{0}^{n-1}$, and

$$
q \cdot C \cdot\left|\vartheta_{i}^{\prime}-p_{i} / q\right| \cong\left|q \cdot\left[C \cdot \vartheta_{i}^{\prime}\right]-C \cdot p_{i}\right|
$$

are of the size $X_{0}^{n-1}$, so that $\left|\vartheta_{i}^{\prime}-p_{i} / q\right|$ are of the size

$$
X_{0}^{n-1} / C \cdot X_{0}^{n-1}=C^{-1} \cong X_{0}^{-n} \cong q^{-(1+1 /(n-1))}
$$

as desired.

The above method has been applied in practice to solve Thue and Thue-Mahler equations by Agrawal, Coates, Hunt and van der Poorten [1980] (using multidimensional continued fractions instead of the $L^{3}$-algorithm), Pethö and Schulenberg [1987], and Blass, Glass, Meronk and Steiner [1987 ${ }^{\text {a }], ~[1987 ~}{ }^{\text {b }}$ ]. So it has proved to be useful. However, we prefer another method, for several reasons. Firstly, it is close to the homogeneous case as described in the previous section, whereas the generalized Davenport method has no obvious counterpart for the homogeneous case. Secondly, it actually produces solutions for which the linear form $\Lambda$ is almost as near to zero as possible under the condition $X \leqslant X_{0}$. Specifically, if a linear relation between the $\vartheta_{i}$ exists, but had not been noticed before (a situation that may occur in practice, cf. Agrawal, Coates, Hunt and van der Poorten [1980]), the method detects these relations, by finding explicitly an extremely short lattice vector (resp. a lattice vector extremely near to a given point) giving the coefficients of the relation. Thirdly, an analogous method for the p-adic case can be given (see Section 3.11). Finally, variations as indicated in Section 1.4 are possible. Concerning computation time we think that the two
methods are about equally fast.

The method works as follows. We take the approximation lattice $\Gamma$ exactly as in the homogeneous case (cf. the previous section), with constants $\gamma, \mathrm{C}$ chosen properly, i.e. $C$ is of the size $X_{0}^{n}$. Compute with the $L^{3}$-algorithm a reduced basis $\underline{c}_{1}, \ldots, \underline{c}_{n}$ of $\Gamma$. Let $\mathscr{C}$ be the matrix associated to this basis, and compute also the transformation matrix $U$ with $\mathscr{C}=\mathscr{B} \cdot \mathcal{U}$, and its inverse $U^{-1}$. Note that $\mathcal{B}^{-1}$, and hence also $\mathscr{C}^{-1}$, are easy to compute, namely by

$$
\mathcal{B}^{-1}=\left(\begin{array}{rlll}
1 / \gamma & & & \varnothing \\
\emptyset & & & \\
-\frac{\left[\gamma \cdot C \cdot \vartheta_{1}\right]}{\gamma \cdot\left[\gamma \cdot C \cdot \vartheta_{n}\right]} & \ldots & -\frac{\left[\gamma \cdot C \cdot \vartheta_{n-1}\right]}{\gamma \cdot\left[\gamma \cdot C \cdot \vartheta_{n}\right]} & \frac{1}{\left[\gamma \cdot C \cdot \vartheta_{n}\right]}
\end{array}\right)
$$

and our version of the $L^{3}$-algorithm (Fig. 1). Let $y \in \mathbb{Z}^{n}$ be defined by

$$
y=(0, \ldots, 0,-[\gamma \cdot C \cdot \beta])^{T}=\sum_{i=1}^{n} s_{i} \cdot \subseteq_{i}
$$

where the coefficients $\mathbf{s}_{\mathbf{i}} \in \mathbb{R}$ can be computed by

$$
\left(s_{1}, \ldots, s_{n}\right)^{T}=\mathscr{E}^{-1} \cdot y
$$

To be more precise, if $u^{-1}$ has $\underline{u}$ as $n$th column, then $\mathscr{C}^{-1}$ has $\underline{\mathbf{u}} /\left[\gamma \cdot \mathrm{C} \cdot \vartheta_{\mathrm{n}}\right]$ as n th column, so

$$
\left(s_{1}, \ldots, s_{n}\right)^{T}=-\underline{u} \cdot[\gamma \cdot C \cdot \beta] /\left[\gamma \cdot C \cdot \vartheta_{n}\right] .
$$

Now we apply Lemma 3.5 or 3.6 , that provide a lower bound for $\ell(\Gamma, y)$. Then we can apply the following lemma.

LEMMA 3.10. Let $X_{1}$ be a positive constant such that

$$
\begin{equation*}
\ell(\Gamma, y) \geqslant V\left((n+2)^{2}+(n-1) \gamma^{2}\right) \cdot x_{1} \tag{3.26}
\end{equation*}
$$

Then (3.1) has no solutions with

$$
\begin{equation*}
\frac{1}{\delta} \cdot \log \left(\gamma \cdot C \cdot c / X_{1}\right) \leqslant X \leqslant X_{1} . \tag{3.27}
\end{equation*}
$$

Remark. We apply this lemma for $X_{1}=X_{0}$. If condition (3.26) then fails, we must take a larger constant $C$. If it holds for a constant $C$ of the
size $X_{0}^{n}$, then (3.27) yields a reduced lower bound for $X$ of size $\log X_{0}$. Proof. Let $x_{1}, \ldots, x_{n}$ be a solution of (3.1) with $0<X \leqslant X_{1}$. Consider the lattice point

$$
\underline{x}=\sum_{i=1}^{n} x_{i} \cdot \underline{b}_{i}=\left(\gamma \cdot x_{1}, \ldots, \gamma \cdot x_{n-1}, \tilde{\Lambda}_{0}\right)^{T},
$$

with

$$
\tilde{\Lambda}_{0}=\sum_{i=1}^{n} x_{i} \cdot\left[\gamma \cdot C \cdot \vartheta_{i}\right] .
$$

Put $\tilde{\Lambda}=[\gamma \cdot C \cdot \beta]+\tilde{\Lambda}_{0}$. Then

$$
|\underline{x}-\underline{y}|^{2}=\gamma^{2} \cdot \sum_{i=1}^{n-1} x_{i}^{2}+\tilde{\Lambda}^{2} \leqslant(n-1) \cdot \gamma^{2} \cdot x_{1}^{2}+\tilde{\Lambda}^{2}
$$

and

$$
\begin{aligned}
|\tilde{\Lambda}-\gamma \cdot C \cdot \Lambda| & \leqslant|[\gamma \cdot C \cdot \beta]-\gamma \cdot C \cdot \beta|+\sum_{i=1}^{n}\left|x_{i}\right| \cdot\left|\left[\gamma \cdot C \cdot \vartheta_{i}\right]-\gamma \cdot C \cdot \vartheta_{i}\right| \\
& \leqslant 1+\sum_{i=1}^{n}\left|x_{i}\right| \leqslant 1+n \cdot X_{1} \leqslant(n+1) \cdot X_{1}
\end{aligned}
$$

By (3.1), (3.26) and the definition of $\ell(\Gamma, \underline{y})$ the result follows, since

$$
\begin{aligned}
& \gamma \cdot C \cdot c \cdot \exp (-\delta \cdot X)>|\gamma \cdot C \cdot \Lambda| \geqslant|\tilde{\Lambda}|-|\tilde{\Lambda}-\gamma \cdot C \cdot \Lambda| \\
& \geqslant V\left(\ell(\Gamma, \underline{y})^{2}-(n-1) \cdot \gamma^{2} \cdot X_{1}^{2}\right)-(n+1) \cdot X_{1} \geqslant X_{1} .
\end{aligned}
$$

Again we may prove refinements of the above lemma, similar to Lemma 3.8 in the homogeneous case. We explained in Section 3.5. how to apply the Fincke and Pohst algorithm in the inhomogeneous case. We do not work that out here.

Summarizing, the method described above is based on the fact that a large solution of (3.1) in the inhomogeneous case leads to a lattice point extremely near to a fixed point in $\mathbb{Z}^{n}$. We can actually prove by some computations that such lattice points do not exist, so that such extreme solutions do not exist. The method outlined in this section is used in Chapter 8. Note that in the case $n=2$ the method is essentially the same as the Davenport lemma.

### 3.9. Inhomogeneous zero-dimensional approximation in the p-adic case.

In the p-adic case we start with a very simple linear form $\Lambda$, to which also a very simple reduction method applies. Let $\Lambda$ be

$$
\Lambda=\beta+x \cdot \vartheta,
$$

for $\beta, \vartheta \in \Omega_{p}$ such that $\beta / \vartheta \in \mathbb{Q}_{p}$, and $x \in \mathbb{Z}, x>0$. It is obvious that in the real case with such a simple linear form $\Lambda$ inequality (3.1) has only finitely many solutions (we even don't need (3.2)), that are easy to compute. In the p-adic case however, inequality (3.3) may have infinitely many solutions, so we do need a bound like (3.4), and a reduction method.

Put $\vartheta^{\prime}=-\beta / \vartheta$. Then $\vartheta^{\prime} \in \mathbb{Q}_{p}$. Inequality (3.3) now becomes

$$
\begin{equation*}
\operatorname{ord}_{p}\left(\vartheta^{\prime}-x\right) \geqslant c_{1}^{\prime}+c_{2} \cdot x, \tag{3.28}
\end{equation*}
$$

where $c_{1}^{\prime}, c_{2}$ are constants with $c_{2}>0$. We assume that

$$
x \geqslant-c_{1}^{\prime} / c_{2} .
$$

Then (3.28) has no solutions if ord ${ }_{p}\left(\vartheta^{\prime}\right)<0$. Hence we may assume that $\vartheta^{\prime}$ is a p-adic integer. Let the p-adic expansion of $\vartheta$ ' be

$$
\vartheta^{\prime}=\sum_{i=0}^{\infty} u_{i} \cdot p^{i}
$$

where $u_{i} \in\{0,1, \ldots, p-1\}$ for all $i \in \mathbb{N}_{0}$. Compute the p-adic digits $u_{i}$ far enough to be able to apply the following reduction lemma.

LEMMA 3.11. Let $X_{1}$ be a positive constant. Let $r$ be the minimal index such that

$$
\begin{equation*}
p^{r}>x_{1}, \quad u_{r} \neq 0 \tag{3.29}
\end{equation*}
$$

Then (3.28) has no solutions with

$$
\begin{equation*}
\left(r-c_{1}^{\prime}\right) / c_{2}<x \leqslant X_{1} . \tag{3.30}
\end{equation*}
$$

Remark. We apply the lemma with $X_{1}=X_{0}$. The assumption behind the lemma is that in the p-adic expansion of $\vartheta^{\prime}$ no long sequences of zeroes appear. In fact, it seems that in our applications the numbers $u_{i}$ are distributed randomly over $\{0,1, \ldots, p-1\}$. Then the minimal $r$ satisfying (3.29)
will not be much larger than $\log X_{0} / \log p$, and then (3.30) yields a reduced upper bound of size $\log X_{0}$, as desired.

Proof. Let $x \leqslant X_{1}$ satisfy (3.28). Suppose that ord${ }_{p}\left(\vartheta{ }^{\prime}-x\right) \geqslant r+1$. Then $x \equiv \sum_{i=0}^{r} u_{i} \cdot p^{i}\left(\bmod p^{r+1}\right)$.

By $x \geqslant 0$ it follows from (3.29) that

$$
x \geqslant \sum_{i=0}^{r} u_{i} \cdot p^{i} \geqslant u_{r} \cdot p^{r} \geqslant p^{r}>x_{1}
$$

which contradicts the assumption $x \leqslant X_{1}$. Hence $\operatorname{ord}_{p}(\vartheta,-x) \leqslant r$, and (3.30) follows from (3.28).

Remark. In the above proof it is essential that $x \geqslant 0$. It is however not difficult to formulate a similar result that holds for all $x \in \mathbb{Z}$, by looking, if $p \neq 2$ for $p$-adic digits $u_{i}$ that are not only $\neq 0$ but also $\neq p-1$, and if $p=2$ for $p$-adic digits $u_{i}, u_{i+1}$ with $u_{i} \neq u_{i+1}$.

A method very similar to the one described above was used by Wagstaff [1979], [1981], a.o. for solving $5^{n} \equiv 2\left(\bmod 3^{n}\right)$. We apply the method in Chapter 4.
3.10. Homogeneous one-dimensional approximation in the p-adic case: p-adic continued fractions and approximation lattices of p-adic numbers.

Let $\Lambda$ have the form

$$
\Lambda=\mathrm{x}_{1} \cdot \vartheta_{1}+\mathrm{x}_{2} \cdot \vartheta_{2}
$$

where $\vartheta_{1}, \vartheta_{2} \in \Omega_{p}$ such that $\vartheta=-\vartheta_{1} / \vartheta_{2} \in \mathbb{Q}_{p}$, and $x_{1}, x_{2} \in \mathbb{Z}$. We may assume that $\operatorname{ord}_{p}(\vartheta) \geqslant 0$. Now

$$
\Lambda^{\prime}=\Lambda / \vartheta_{1}=-\mathrm{x}_{1} \cdot \vartheta+\mathrm{x}_{2} .
$$

So (3.3) now means that the rational number $x_{2} / x_{1}$ is $p$-adically close to the p -adic number $\vartheta$.

In analogy of the real case it seems reasonable to study p-adic continued fraction algorithms. However, a p-adic continued fraction algorithm that provides all best approximations to a p-adic number seems not to exist.

Therefore we introduce the concept of p-adic approximation lattices, as was done in de Weger $\left[1986^{\text {a }}\right]$. From this paper we adopt the best approximation algorithm, which is a generalization of the algorithm of Mahler [1961], Chapter IV. This algorithm goes back also on the euclidean algorithm, and thus is close to a continued fraction algorithm. But it is not a p-adic continued fraction algorithm in the sense that a p-adic number is expanded into a continued fraction, and that the approximations are then found by truncating the continued fraction.

Recall that for $\mu \in \mathbb{N}_{0}$ the rational integer $\vartheta^{(\mu)}$ is defined by $\operatorname{ord}_{p}\left(\vartheta-\vartheta^{(\mu)}\right) \geqslant \mu$ and $0 \leqslant \vartheta^{(\mu)}<p^{\mu}$. We define for any $\mu \in \mathbb{N}_{0}$ the $p$-adic approximation lattice $\Gamma_{\mu}$ by a matrix to which a basis of $\Gamma_{\mu}$ is associated, namely the matrix

$$
\left(\begin{array}{cc}
1 & 0 \\
\vartheta^{(\mu)} & p^{\mu}
\end{array}\right) .
$$

Then it is easy to see that

$$
\Gamma_{\mu}=\left\langle\left(x_{1}, x_{2}\right)^{T} \in \mathbb{Z}^{2} \mid \operatorname{ord}_{p}\left(x_{2}-x_{1} \cdot \vartheta\right) \geqslant \mu\right\rangle
$$

(cf. Lemma 3.13 in the next section, where we prove a more general result). The following algorithm computes a point of minimal length in $\Gamma_{\mu}$.

```
\(\underline{x}:=\left(1, \vartheta^{(\mu)}\right)^{\mathrm{T}} ; \underline{y}:=\left(0, \mathrm{p}^{\mu}\right)^{\mathrm{T}}\);
if \(|\underline{x}|>|y|\), interchange \(\underline{x}\) and \(\underline{y}\);
(1) compute \(K \in \mathbb{Z}\) such that \(|\underline{y}-K \cdot \underline{x}|\) is minimal ;
\(\underline{y}:=\underline{y}-K \cdot \underline{x}\);
if \(|\underline{x}|>|y|\), interchange \(\underline{x}\) and \(\underline{y}\), and go to (1) ;
print \(\underline{x}\).
```

Figure 3. p-adic approximation algorithm.

With this algorithm it is possible to compute $\ell\left(\Gamma_{\mu}\right)$ explicitly. Then we can apply the following lemma.

LEMMA 3.12. Let $X_{1}$ be a constant such that

$$
\begin{equation*}
\ell\left(\Gamma_{\mu}\right)>\sqrt{ } 2 \cdot X_{1} \tag{3.31}
\end{equation*}
$$

Then (3.3) has no solutions with

$$
\begin{equation*}
\left(\mu-1-c_{1}+\operatorname{ord}_{p}\left(\vartheta_{2}\right)\right) / c_{2}<X_{j} \leqslant X \leqslant X_{1} . \tag{3.32}
\end{equation*}
$$

Remark. We take $\mu$ such that $p^{\mu}$ is of the size of $X_{0}^{2}$, and apply the lemma for $X_{1}=X_{0}$. Then we expect that $\ell\left(\Gamma_{\mu}\right)$ is of the size of $X_{0}$, so that (3.31) is a reasonable condition.

Proof. Apply the proof of Lemma 3.14 (in the next section) for $n=2$.

A method like the one described above has been applied by Agrawal, Coates, Hunt and van der Poorten [1980]. We use it in Chapters 6 and 7.
3.11. Homogeneous multi-dimensional approximation in the p-adic case: p-adic approximation lattices.

We now study the case

$$
\Lambda=\sum_{i=1}^{n} x_{i} \cdot \vartheta_{i}
$$

where $\vartheta_{i} \in \Omega_{p}$ such that $\vartheta_{i} / \vartheta_{j} \in \mathbb{Q}_{p}, \quad x_{i} \in \mathbb{Z}$ for all $i, j$, and with $n \geqslant 2$. We may assume that $\operatorname{ord}_{p}\left(\vartheta_{i}\right)$ is minimal for $i=n$. Put

$$
\vartheta_{i}^{\prime}=-\vartheta_{i} / \vartheta_{\mathrm{n}} \text { for } \mathrm{i}=1, \ldots, \mathrm{n}-1 .
$$

Then $\vartheta_{i}^{\prime} \in \mathbb{Z}_{p}$ for all i. Put

$$
\Lambda^{\prime}=\Lambda / \vartheta_{n}=-\sum_{i=1}^{n-1} x_{i} \cdot \vartheta_{i}^{\prime}+x_{n}
$$

The definition of the p-adic approximation lattices can be generalized directly from the one-dimensional case. Namely, for any $\mu \in \mathbb{N}_{0}$ we define $\Gamma_{\mu}$ as the lattice associated to the matrix

$$
\mathcal{B}_{\mu}=\left(\begin{array}{rrrrr}
1 & & & & \varnothing \\
& & \ldots & & \\
& \varnothing & & 1 & \\
\vartheta_{1}^{\prime}(\mu) & & & \\
& \ldots & \vartheta_{n-1}^{\prime}(\mu) & p^{\mu}
\end{array}\right)
$$

Then we have the following result.

LEMMA 3.13. The lattice $\Gamma_{\mu}$, associated to the above defined matrix $\mathcal{B}_{\mu}$, is equal to the set

$$
\Gamma_{\mu}=\left\langle\left(x_{1}, \ldots, x_{n}\right)^{T} \in \mathbb{Z}^{\mathrm{n}} \mid \operatorname{ord}_{\mathrm{p}}\left(\Lambda^{\prime}\right) \geqslant \mu\right\rangle
$$

Proof. For any $\underline{x}=\left(x_{1}, \ldots, x_{n}\right)^{T} \in \Gamma_{\mu}$ there exists a $\underline{z}=\left(z_{1}, \ldots, z_{n}\right)^{T} \in \mathbb{Z}^{n}$ such that $\underline{x}=\mathcal{B}_{\mu} \cdot \underline{z}$. Then $x_{i}=z_{i}$ for $i=1, \ldots, n-1$, and

$$
x_{n}=\sum_{i=1}^{n-1} z_{i} \cdot \vartheta_{i}^{\prime}(\mu)+z_{n} \cdot p^{\mu} \equiv \sum_{i=1}^{n-1} x_{i} \cdot \vartheta^{\prime}{ }_{i}\left(\bmod p^{\mu}\right)
$$

Hence $\operatorname{ord}_{p}\left(\Lambda^{\prime}\right) \geqslant \mu$. Conversely, for any $\underline{x}=\left(x_{1}, \ldots, x_{n}\right)^{T}$ such that $\operatorname{ord}_{p}\left(\Lambda^{\prime}\right) \geqslant \mu$ there obviously exists $a \quad \underline{z} \in \mathbb{Z}^{n}$ such that $\underline{x}=\mathcal{B}_{\mu} \cdot \underline{z}$.

Using the $L^{3}$-algorithm we can compute a lower bound for $\ell\left(\Gamma_{\mu}\right)$. Then we can apply the following lemma, which is a direct generalization of Lemma 3. 12.

LEMMA 3.14. Let $X_{1}$ be a constant such that

$$
\begin{equation*}
\ell\left(\Gamma_{\mu}\right)>V_{\mathrm{n}} \cdot \mathrm{X}_{1} \tag{3.33}
\end{equation*}
$$

Then (3.3) has no solutions with

$$
\begin{equation*}
\left(\mu-1-c_{1}+\operatorname{ord}_{p}\left(\vartheta_{n}\right)\right) / c_{2}<x_{j} \leqslant X \leqslant X_{1} \tag{3.34}
\end{equation*}
$$

Remark. We take $\mu$ such that $p^{\mu}$ is of the size of $X_{0}^{n}$, and apply the lemma for $X_{1}=X_{0}$. Then we expect that $\ell\left(\Gamma_{\mu}\right)$ is of the size of $X_{0}$, so that (3.33) is a reasonable condition.

Proof. Let $x_{1}, \ldots, x_{n}$ be a solution of (3.3) with $X \leqslant X_{1}$. Then (3.33) prohibits the point $\left(x_{1}, \ldots, x_{n}\right)^{T}$ from being a lattice point in $\Gamma_{\mu}$. Hence, by Lemma 3.13, $\operatorname{ord}_{p}\left(\Lambda^{\prime}\right) \leqslant \mu-1$, and (3.34) follows from (3.3).

We will apply the results of this section in Chapters 6 and 7 .
3.12. Inhomogeneous one- and multi-dimensional approximation in the p-adic case.

Finally we study an inhomogeneous p-adic form

$$
\Lambda=\beta+\sum_{i=1}^{n} x_{i} \cdot \vartheta_{i}
$$

where $\beta, \vartheta_{i} \in \Omega_{p}$ such that $\beta / \vartheta_{j}, \vartheta_{i} / \vartheta_{j} \in \mathbb{Q}_{p}$ and $X_{i} \in \mathbb{Z}$ for all $i$, $j$, and $n \geqslant 2$. We assume that $\operatorname{ord}_{p}\left(\vartheta_{i}\right)$ is minimal for $i=n$, and that $\operatorname{ord}_{p}(\beta) \geqslant \operatorname{ord}_{p}\left(\vartheta_{n}\right)$. Put

$$
\begin{gathered}
\vartheta_{i}^{\prime}=-\vartheta_{i} / \vartheta_{n} \text { for } i=1, \ldots, n-1, \quad \beta^{\prime}=\beta / \vartheta_{n}, \\
\Lambda^{\prime}=\Lambda / \vartheta_{n}=\beta^{\prime}-\sum_{i=1}^{n-1} x_{i} \cdot \vartheta_{i}^{\prime}+x_{n} .
\end{gathered}
$$

Then $\beta^{\prime}, \vartheta_{i}^{\prime} \in \mathbb{Z}_{p}$ for all $i$. As $p$-adic approximation lattices we take the lattices $\Gamma_{\mu}$ that were defined for the homogeneous case, i.e. for any $\mu \in \mathbb{N}_{0}$ the lattice $\Gamma_{\mu}$ that is associated to the matrix $\mathcal{B}_{\mu}$ (see Section 3.11). Further put

$$
y=\left(0, \ldots, 0, \beta^{\prime}(\mu)\right)^{\mathrm{T}}=\sum_{i=1}^{\mathrm{n}} \mathrm{~s}_{\mathrm{i}} \cdot \underline{c}_{i} \in \mathbb{Z}^{\mathrm{n}}
$$

where $\underline{c}_{1}, \ldots, \underline{c}_{n}$ is a reduced basis of $\Gamma_{\mu}$, and $s_{i} \in \mathbb{R}$. By Lemma 3.5 or 3.6 we can compute a lower bound for $\ell(\Gamma, y)$. This is useful in view of the following lemma.

LEMMA 3.15. The $\operatorname{set} \Gamma_{\mu}(y)=\Gamma_{\mu}+y$ is equal to the set

$$
\Gamma_{\mu}(y)=\left\langle\left(x_{1}, \ldots, x_{n}\right)^{T} \in \mathbb{Z}^{\mathrm{n}} \mid \operatorname{ord}_{\mathrm{p}}\left(\Lambda^{\prime}\right) \geqslant \mu\right\rangle
$$

Proof. Let $\underline{x}=\left(x_{1}, \ldots, x_{n}\right)^{T}$ satisfy $\underline{x}-\underline{y} \in \Gamma_{\mu}$. Note that

$$
\underline{x}-\underline{y}=\left(x_{1}, \ldots, x_{n-1}, x_{n}-\beta,(\mu)\right)^{T}
$$

By Lemma 3.13 we have

$$
\operatorname{ord}_{p}\left(\sum_{i=1}^{n-1} x_{i} \cdot \vartheta_{i}-\left(x_{n}-\beta^{\prime}(\mu)\right)\right) \geqslant p^{\mu}
$$

The left hand side is just ord ${ }_{p}\left(\Lambda^{\prime}\right)$, which proves the lemma.

Obviously, the length of the shortest vector in $\Gamma_{\mu}(y)$ (a translated lattice) is equal to $\ell\left(\Gamma_{\mu}, y\right)$ (unless in the case $y \in \Gamma_{\mu}$, i.e. $s_{i} \in \mathbb{Z}$ for all i ). We have the following useful lemma.

LEMMA 3.16. Let $X_{1}$ be a constant such that

$$
\begin{equation*}
\ell\left(\Gamma_{\mu}, \underline{y}\right)>V_{\mathrm{n}} \cdot \mathrm{X}_{1} . \tag{3.35}
\end{equation*}
$$

Then (3.3) has no solutions with

$$
\begin{equation*}
\left(\mu-1-c_{1}+\operatorname{ord}_{p}\left(\vartheta_{n}\right)\right) / c_{2}<x_{j} \leqslant X \leqslant X_{1} \tag{3.36}
\end{equation*}
$$

Remark. We take $\mu$ such that $p^{\mu}$ is of the size of $X_{0}^{n}$, and apply the lemma for $X_{0}=X_{1}$. Then we expect that $\ell\left(\Gamma_{\mu}, \underline{y}\right)$ is of the size of $X_{0}$, so that (3.35) is a reasonable condition.

Proof. Let $x_{1}, \ldots, x_{n}$ be a solution of (3.3) with $X \leqslant X_{1}$. Then (3.35) prohibits the point $\left(x_{1}, \ldots, x_{n}\right)^{T}$ from being in $\Gamma_{\mu}(y)$. Hence, by Lemma 3.15, $\operatorname{ord}_{p}\left(\Lambda^{\prime}\right) \leqslant \mu-1$, and (3.36) follows from (3.3).

We will not apply the above lemma in this book. It is included here only for the sake of completeness. However, when solving Thue-Mahler equations (see Section 8.6), it will be of use.

### 3.13. Useful sublattices of p-adic approximation lattices.

In our p-adic applications of solving diophantine equations via linear forms, we always have linear forms in logarithms of algebraic numbers, i.e. in

$$
\Lambda=\beta+\sum_{i=1}^{n} x_{i} \cdot \vartheta_{i}
$$

the $\beta$ and $\vartheta_{i}$ 's are p-adic logarithms of algebraic numbers, say

$$
\beta=\log _{p}\left(\alpha_{0}\right), \quad \vartheta_{i}=\log _{p}\left(\alpha_{i}\right) \text { for } i=1, \ldots, n
$$

In Section 2.3 we have seen that for a $\xi \in \mathbb{Q}_{p}$ if $\operatorname{ord}_{p}(1 \pm \xi)>1 /(p-1)$ then $\operatorname{ord}_{p}\left(\log _{p}(\xi)\right)=\operatorname{ord}_{p}(1 \pm \xi)$. In our applications we apply this to

$$
\xi=\alpha_{0} \cdot \prod_{i=1}^{n} \alpha_{i}^{x_{i}}
$$

for which ord $(\xi-1)$ is large. This implies that ord ${ }_{p}\left(\log _{p}(\xi)\right)$ is large too, on which we based the definition of our approximation lattices. However, the converse is not necessarily true: ord $\left(\log _{p}(\xi)\right)$ being large does not imply that $\operatorname{ord}_{p}(\xi-1)$ is large. This is due to the fact that the $p$-adic
logarithm is a multi-branched function. To be more precise, for any root of unity $\zeta \in \mathbb{Q}_{p}$ we have $\log _{p}(\zeta)=0$ (cf. Section 2.3). In $\mathbb{Q}_{p}$ there exist only the $(p-1)$ th roots of unity if $p$ is odd, and only $\pm 1$ as roots of unity if $p=2$. Let $\zeta$ be a primitive $(p-1)$ th root of unity if $p$ is odd, and $\zeta=-1$ if $p=2$. It follows that ord ${ }_{p}\left(\log _{p}(\xi)\right)$ being large implies that for some $k \in\{0,1, \ldots, p-2\}$ (or $k \in\{0,1\}$ if $p=2$ )

$$
\operatorname{ord}_{p}\left(\log _{p}(\xi)\right)=\operatorname{ord}_{p}\left(\xi-\zeta^{k}\right)
$$

The set of $x_{1}, \ldots, x_{n}$ such that ord $_{p}(\xi-1)$ (or ord ${ }_{p}(\xi \pm 1)$ if one wishes) is large, turns out to be a sublattice $\Gamma_{\mu}^{*}$ (or $\Gamma_{\mu}^{\#} \quad$ respectively) of $\Gamma_{\mu}$. In the following lemma we shall prove this fact, and indicate how a basis of such a sublattice can be found. Then we can work with this sublattice instead of $\Gamma_{\mu}$ itself. Of course, in Lemmas $3.12,3.14$ and 3.16 we can replace $\Gamma_{\mu}$ by these sublattices $\Gamma_{\mu}^{*}, \Gamma_{\mu}^{\#}$. For simplicity we assume that $\alpha_{i} \in \mathbb{Q}_{p}$ for all i. We take $\alpha_{0}=1$ (corresponding to $\beta=0$, thus to the homogeneous case), and leave it to the reader to define appropriate translated lattices $\Gamma_{\mu}^{*}(y), \Gamma_{\mu}^{\#}(y)$ for the case $\alpha_{0} \neq 1$ (the inhomogeneous case).

LEMMA 3.17. (i). Let $\alpha_{1}, \ldots, \alpha_{n} \in \mathbb{Q}_{p}$ be given numbers with ord ${ }_{p}\left(\alpha_{i}\right)=0$ for all $i$, and $\operatorname{ord}_{p}\left(\log _{p}\left(\alpha_{i}\right)\right)$ minimal for $i=n$. Let $x_{1}, \ldots, x_{n} \in \mathbb{Z}$. Put

$$
\xi=\prod_{i=1}^{n} \alpha_{i}^{x_{i}}, \quad \mu_{0}=\operatorname{ord}_{p}\left(\log _{p}\left(\alpha_{n}\right)\right)
$$

For any $\mu \in \mathbb{N}_{0}$ put

$$
\begin{aligned}
& \Gamma_{\mu}=\left\langle\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{Z}^{n} \mid \operatorname{ord}_{p}\left(\log _{p}(\xi)\right) \geqslant \mu+\mu_{0}\right\rangle \\
& \Gamma_{\mu}^{*}=\left\langle\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{Z}^{n} \mid \operatorname{ord}_{p}(\xi \pm 1) \geqslant \mu+\mu_{0}\right\rangle \\
& \Gamma_{\mu}^{\#}=\left\langle\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{Z}^{n} \mid \operatorname{ord}_{p}(\xi-1) \geqslant \mu+\mu_{0}\right\rangle
\end{aligned}
$$

Then $\Gamma_{\mu}^{\#} \subseteq \Gamma_{\mu}^{*} \subseteq \Gamma_{\mu}$ are lattices. If $\mathrm{p}=2$ they are all equal. If $\mathrm{p}=3$ then $\Gamma_{\mu}^{*}=\Gamma_{\mu}$. If $\mathrm{p} \geqslant 3$ then $\#\left(\Gamma_{\mu} / \Gamma_{\mu}^{*}\right)=(\mathrm{p}-1) / 2, \quad \#\left(\Gamma_{\mu} / \Gamma_{\mu}^{\#}\right)=\mathrm{p}-1$, $\#\left(\Gamma_{\mu}^{*} / \Gamma_{\mu}^{\#}\right)^{\mu}=2{ }^{\mu}$.
(ii). Let $\underline{b}_{1}, \ldots, \underline{b}_{n}$ be a basis of $\Gamma_{\mu}$. Define $k(\underline{x})$ for any $\underline{x}=\left(x_{1}, \ldots, x_{n}\right)^{\frac{\mathrm{T}}{1}} \in \Gamma_{\mu}$ by

$$
\xi \equiv \zeta^{k(\underline{x})}\left(\bmod p^{\mu+\mu_{0}}\right), \quad k(\underline{x}) \in\{0,1, \ldots, p-2\}
$$

Let $\underline{b}_{1}^{\prime}, \ldots, \underline{b}_{\mathrm{n}}$ be a basis of $\Gamma_{\mu}$ such that

$$
k\left(\underline{b}_{n}^{\prime}\right)=\operatorname{gcd}\left(k\left(\underline{b}_{1}\right), \ldots, k\left(\underline{b}_{n}\right)\right)
$$

Put for $\mathrm{i}=1, \ldots, \mathrm{n}-1$ and $\mathrm{p} \geqslant 5$

$$
\begin{gathered}
\gamma_{i}^{*} \equiv k\left(\underline{b}_{i}^{\prime}\right) / k\left(\underline{b}_{n}^{\prime}\right)(\bmod (p-1) / 2), \quad\left|\gamma_{i}^{*}\right| \leqslant(p-1) / 4, \\
\underline{b}_{i}^{*}=\underline{b}_{i}^{\prime}-\gamma_{i}^{*} \cdot \underline{b}_{n}^{\prime},
\end{gathered}
$$

and for $p \geqslant 3$ also

$$
\begin{gathered}
\gamma_{i}^{\#} \equiv k\left(\underline{b}_{i}^{\prime}\right) / k\left(\underline{b}_{n}^{\prime}\right)(\bmod (p-1)), \quad\left|\gamma_{i}^{\#}\right| \leqslant(p-1) / 2, \\
\underline{b}_{i}^{\#}=\underline{b}_{i}^{\prime}-\gamma_{i}^{\#} \cdot \underline{b}_{n}^{\prime} .
\end{gathered}
$$

Further put for $p \geqslant 5$

$$
\gamma_{\mathrm{n}}^{*}=\operatorname{lcm}\left(\mathrm{k}\left(\underline{b}_{\mathrm{n}}^{\prime}\right),(\mathrm{p}-1) / 2\right) / \mathrm{k}\left(\underline{b}_{\mathrm{n}}^{\prime}\right), \quad \underline{b}_{\mathrm{n}}^{*}=\gamma_{\mathrm{n}}^{*} \cdot \underline{b}_{\mathrm{n}}^{\prime},
$$

and for $p \geqslant 3$ also

$$
\gamma_{n}^{\#}=\operatorname{lcm}\left(k\left(\underline{b}_{n}^{\prime}\right), p-1\right) / k\left(\underline{b}_{n}^{\prime}\right), \quad \underline{b}_{n}^{\#}=\gamma_{n}^{\#} \cdot \underline{b}_{n}^{\prime} .
$$

Then $\underline{b}_{1}^{*}, \ldots, \underline{b}_{n}^{*}$ is a basis of $\Gamma_{\mu}^{*}$, and $\underline{b}_{1}^{\#}, \ldots, \underline{b}_{n}^{\#}$ is a basis of $\Gamma_{\mu}^{\#}$.
Proof. (i). It is trivial that $\Gamma_{\mu}^{\#} \subseteq \Gamma_{\mu}^{*} \subseteq \Gamma_{\mu}$, and that they are lattices. The equalities of the lattices for $p=2,3$ follow from the fact that $\pm 1$ are the only roots of unity in $\mathbb{Q}_{p}$ for $p=2,3$. The values of $\#\left(\Gamma_{\mu} / \Gamma_{\mu}^{*}\right)$, etc., follow from (ii).
(ii). Note that $k(\underline{x})$ is $(\bmod (p-1))$ a linear function on $\Gamma_{\mu}$. The points $\underline{x}$ of $\Gamma_{\mu}^{*}$ are characterized by $(p-1) / 2 \mid k(\underline{x})$, and the points $\underline{x}$ of $\Gamma_{m}^{\#}$ are characterized by $(p-1) \mid k(\underline{x})$. It follows from the definitions in the lemma that for $\mathrm{i}=1, \ldots, \mathrm{n}-1$

$$
\begin{aligned}
& k\left(\underline{b}_{i}^{*}\right) \equiv k\left(\underline{b}_{i}^{\prime}\right)-\gamma_{i}^{*} \cdot k\left(\underline{b}_{n}^{\prime}\right) \equiv 0(\bmod (p-1) / 2), \\
& k\left(\underline{b}_{i}^{\#}\right) \equiv k\left(\underline{b}_{i}^{\prime}\right)-\gamma_{i}^{\#} \cdot k\left(\underline{b}_{n}^{\prime}\right) \equiv 0(\bmod (p-1))
\end{aligned}
$$

Note that $\underline{b}_{1}^{*}, \ldots, \underline{b}_{n-1}^{*}, \underline{b}_{n}^{\prime}$ and $\underline{b}_{1}^{\#}, \ldots, \underline{b}_{n-1}^{\#}, \underline{b}_{n}^{\prime}$ are both bases of $\Gamma_{\mu}$. Write $\quad \underline{x} \in \Gamma_{\mu}$ as

$$
\underline{x}=\sum_{i=1}^{n-1} y_{i}^{*} \cdot \underline{b}_{i}^{*}+y_{n}^{*} \cdot \underline{b}_{n}^{\prime}=\sum_{i=1}^{n-1} y_{i}^{\#} \cdot \underline{b}_{i}^{\#}+y_{n}^{\#} \cdot \underline{b}_{n}^{\prime}
$$

for integers $y_{i}^{*}, y_{i}^{\#}$. Then it follows that

$$
\begin{aligned}
& k(\underline{x}) \equiv y_{n}^{*} \cdot k\left(\underline{b}_{n}^{\prime}\right)(\bmod (p-1) / 2) \\
& k(\underline{x}) \equiv y_{n}^{\#} \cdot k\left(\underline{b}_{n}^{\prime}\right)(\bmod (p-1))
\end{aligned}
$$

So $\underline{x} \in \Gamma_{\mu}^{*}$ if and only if $\gamma_{n}^{*} \mid y_{n}^{*}$, and $\underline{x} \in \Gamma_{\mu}^{\#}$ if and only if $\gamma_{n}^{\#} \mid y_{n}^{\#}$. This proves the result.

## Chapter 4. S-integral elements of binary recurrence sequences.

Acknowledgements. The research for this chapter has been done partly in cooperation with A. Pethö from Debrecen. The results have been published in Pethö and de Weger [1986] and de Weger [1986].

### 4.1. Introduction.

In this chapter we present a reduction algorithm for the following problem. Let $A, B, G_{0}, G_{1}$ be integers, and let the recurrence sequence $\left\{G_{n}\right\}_{n=0}^{\infty}$ be defined by

$$
G_{n+1}=A \cdot G_{n}-B \cdot G_{n-1} \text { for } n=1,2, \ldots
$$

Assume that $\Delta=A^{2}-4 \cdot B$ is not a square, and that the sequence is not degenerate (this will be explained below). Let $w$ be a nonzero integer, and let $p_{1}, \ldots, p_{s}$ be distinct primes. We study the diophantine equation

$$
\begin{equation*}
G_{n}=w \cdot \prod_{i=1}^{s} p_{i}^{m_{i}} \tag{4.1}
\end{equation*}
$$

in nonnegative integers $n, m_{1}, \ldots, m_{s}$. We will study both the cases of positive and negative discriminant $\Delta$ (the 'hyperbolic' and 'elliptic' cases). It was shown by Mahler [1934] that (4.1) has only finitely many solutions. For the case $\Delta>0$ Schinzel [1967] has given an effectively computable upper bound for the solutions.

Mignotte $\left[1984^{\mathrm{a}}\right]$, $\left[1984^{\mathrm{b}}\right.$ ] indicated how in some instances (4.1) with $\mathrm{s}=1$ can be solved by congruence techniques. It is however not clear that his method will work for any equation (4.1) with $s=1$. Moreover, his method seems not to be generalizable for $s>1$. Pethö [1985] has given a reduction algorithm, based on the Gelfond-Baker method, to treat (4.1) in the case $\Delta>0, \quad \mathrm{w}=\mathrm{s}=1$.

Our reduction algorithms are based on a simple case of $p$-adic diophantine approximation, namely the zero-dimensional case, cf. Section 3.9. In the
hyperbolic case this suffices to be able to find all solutions of (4.1). This is based on a trivial observation on the exponential growth of $\left|G_{n}\right|$ in this case. In the elliptic case the situation is essentially more complicated. Then information on the growth of $\left|G_{n}\right|$ can be obtained from the complex Gelfond-Baker theory. Therefore in this case we have to combine the p-adic arguments with the one-dimensional homogeneous or inhomogeneous real diophantine approximation method, cf. Sections 3.2 and 3.3.

We shall give explicit upper bounds for the solutions of (4.1) which are small enough to admit the practical application of the reduction algorithms, if the parameters of the equation are not too large. Pethö [1985] pointed out that essentially better upper bounds hold for all but possibly one solutions. His reasoning is essentially the same as our reduction technique.

The generalized Ramanujan-Nagell equation

$$
\begin{equation*}
x^{2}+k=\prod_{i=1}^{s} p_{i}^{z_{i}} \tag{4.2}
\end{equation*}
$$

where $k \in \mathbb{Z}$ is fixed, and $x, z_{1}, \ldots, z_{s} \in \mathbb{N}_{0}$ are the unknowns, can be reduced to a finite number of equations of type (4.1) with $\Delta>0$. Equation (4.2) with $s=1$ has a long history (cf. Hasse [1966], Beukers [1981] for a survey), and interesting applications in coding theory (cf. Bremner, Calderbank, Hanlon, Morton and Wolfskill [1983], MacWilliams and Sloane [1977], and Tzanakis and Wolfskill [1986], [1987]). Examples of (4.2) have been solved using the Gelfond-Baker theory by Hunt and van der Poorten (unpublished). They used real or complex, not p-adic linear forms in logarithms. As far as we know, none of the proposed methods to treat (4.2) gives rise to an algorithm which works for arbitrary values of $k$ and the $p_{i}^{\prime} s$, whereas Tzanakis' elementary method (cf. Tzanakis [1983]) seems to be the only one that can be generalized to $s>1$. Our method has both properties.

This chapter is organized as follows. In Section 4.2 we give some preliminaries on binary recurrence sequences. In Section 4.3 we study the growth of $\left|G_{n}\right|$, both in the hyperbolic and the elliptic case. The hyperbolic case is trivial, and in the elliptic case we give a method for solving $\left|G_{n}\right|<v$ for a fixed $v \in \mathbb{R}$, by proving an upper bound for $n$ that has particularly good dependence on $v$, and by showing how to reduce such a bound. Section 4.4 gives upper bounds for the solutions of (4.1).

Section 4.5 gives a lemma on which the p-adic part of the reduction procedure is based. Then Section 4.6 treats some special cases, a.o. the 'symmetric' recurrences. For this special type of recurrence sequences our reduction algorithms fail, but elementary arguments will always work for solving (4.1) in these cases. In Section 4.7 we give the algorithm for reducing upper bounds for the solutions of (4.1) in the case $\Delta>0$, with some elaborated examples. The same is done for the case $\Delta<0$ in Section 4.8.

Section 4.9 shows how to treat the generalized Ramanujan-Nagell equation (4.2), as an application of the hyperbolic case of (4.1). As an example we determine all integers $x$ such that $x^{2}+7$ has no prime factors larger than 20, thus extending the result of Nagell [1948] on the equation $x^{2}+7=2^{n}$ (the original Ramanujan-Nagell equation). Finally in Section 4.10 we give an application of the elliptic case of (4.1) to a certain type of mixed quadratic-exponential diophantine equation, analogous to the application of the hyperbolic case to solving (4.2). As an example, we determine the solutions $X, m_{1}, m_{2}, n$ of

$$
X^{2}-3^{m} 1 \cdot 7^{m} \cdot x+2 \cdot\left(3^{m} \cdot 7^{m}\right)^{2}=11 \cdot 2^{n}
$$

### 4.2. Binary recurrence sequences.

Let $A, B, G_{0}, G_{1} \in \mathbb{Z}$ be given. Let the sequence $\left\{G_{n}\right\}^{\infty}=0$ be defined by

$$
\begin{equation*}
G_{n+1}=A \cdot G_{n}-B \cdot G_{n-1} \text { for } n=1,2, \ldots \tag{4.3}
\end{equation*}
$$

Let $\alpha, \beta$ be the roots of $x^{2}-A \cdot x+B=0$. We assume that $\Delta=A^{2}-4 \cdot B$ is not a square, and that $\alpha / \beta$ is not a root of unity (i.e. the sequence is not degenerate). Put

$$
\begin{equation*}
\lambda=\frac{G_{1}-G_{0} \cdot \beta}{\alpha-\beta}, \quad \mu=\frac{G_{0} \cdot \alpha-G_{1}}{\alpha-\beta} . \tag{4.4}
\end{equation*}
$$

Then $\lambda$ and $\mu$ are conjugates in $K=\mathbb{Q}(\sqrt{ } \Delta)$. We now have for all $n \geqslant 0$

$$
\begin{equation*}
\mathrm{G}_{\mathrm{n}}=\lambda \cdot \alpha^{\mathrm{n}}+\mu \cdot \beta^{\mathrm{n}} \tag{4.5}
\end{equation*}
$$

(cf. Shorey and Tijdeman [1986], Theorem C.1). We will show that when we are solving (4.1), we may assume without loss of generality that

$$
\left(G_{0}, G_{1}\right)=\left(G_{1}, B\right)=(A, B)=1
$$

Namely, if $d=\left(G_{0}, G_{1}\right)$ then $d \mid G_{n}$ for all $n \geqslant 0$, and thus we may study (4.1) with $G_{n}^{\prime}=G_{n} / d$ instead of with $G_{n}$. Next suppose that $d=(A, B)$. If also $d^{2} \mid B$ then it is easy to show that $d^{n-1} \mid G_{n}$ for all $n \geqslant 2$. Then we study (4.1) with $G_{n}^{\prime}=G_{n+1} / d^{n}$ instead of with $G_{n}$. The $A^{\prime}, B^{\prime}$ such that $G_{n+1}^{\prime}=A^{\prime} \cdot G_{n}^{\prime}-B^{\prime} \cdot G_{n-1}^{\prime}$ are $A^{\prime}=A / d, B^{\prime}=B / d^{2}$, and thus $\left(A^{\prime}, B^{\prime}\right)=1^{n+1}$. If however $d^{2} \not{ }^{n-1} B$, then we split the sequence into two parts. We study (4.1) first with $G_{n}^{\prime}=G_{2 \cdot n}$ and then with $G_{n}^{\prime}=G_{2 \cdot n+1}$, instead of with $G_{n}$. For both sequences $\left\{G_{n}^{\prime}\right\}$ the $A^{\prime}, B^{\prime}$ such that $G_{n+1}^{\prime}=A^{\prime} \cdot G_{n}^{\prime}-B^{\prime} \cdot G_{n-1}^{\prime}$ are given by $A^{\prime}=A^{2}-2 \cdot B, B^{\prime}=B^{2}$. Then $\left(A^{\prime}, B^{\prime}\right)=d$, and $d^{2} \mid B^{\prime}$, so we are in the previous case. Finally, let $p$ be a prime such that $p \mid\left(G_{1}, B\right)$, and let $p$ be a prime ideal of $\mathbb{Q}(\sqrt{ } \Delta)$ lying above $p$. By $p \mid B=\alpha \cdot \beta$ we have $p \mid(\alpha)$ or $p \mid(\beta)$. Suppose $\mathfrak{p} \mid(\alpha)$. Then $\mathfrak{p} \not(\beta)$ by $(A, B)=1$ (note that $A=\alpha+\beta$ ). Hence

$$
\operatorname{ord}_{\mathfrak{p}}\left(\lambda \cdot \alpha^{n}+\mu \cdot \beta^{n}\right)=\min \|\left[\operatorname{ord}_{\mathfrak{p}}\left(\lambda \cdot \alpha^{n}\right), \quad \operatorname{ord}_{\mathfrak{p}}\left(\mu \cdot \beta^{n}\right)\right]=\operatorname{ord}_{\mathfrak{p}}(\mu)
$$

if $n \geqslant n_{0}$ for some $n_{0}$. Thus ord $p_{n}$ ) is constant for $n \geqslant n_{0}$, and the same is true if $\mathfrak{p} \mid(\beta)$. Thus we may assume that $\left(G_{1}, B\right)=1$.

LEMMA 4.1. Let $n, m_{1}, \ldots, m_{s}$ be a solution of (4.1). Then, with the above assumptions, we have for $i=1$, ..., $s$ either $m_{i}=0$ or $n=0$ or

$$
\begin{align*}
& \operatorname{ord}_{p_{i}}(\alpha)=\operatorname{ord}_{p_{i}}(\beta)=0  \tag{4.6}\\
& \operatorname{ord}_{p_{i}}(\lambda)=\operatorname{ord}_{p_{i}}(\mu)=-\frac{1}{2} \cdot \operatorname{ord}_{p_{i}}(\Delta) \leqslant 0
\end{align*}
$$

Proof. Suppose $p_{i} \mid B$. Then $p_{i} \nmid A$, hence, from (4.3) and ( $B, G_{1}$ ) $=1$, $p_{i} \nmid G_{n}$ for all $n \geqslant 1$. Thus, $m_{i}=0$ or $n=0$. Next suppose $p_{i} \nmid B$. Then, by $\alpha \cdot \beta=B$,

$$
\operatorname{ord}_{p_{i}}(\alpha)+\operatorname{ord}_{p_{i}}(\beta)=\operatorname{ord}_{p_{i}}(B)=0
$$

Now, $\alpha$ and $\beta$ are algebraic integers, so their $p_{i}$-adic orders are nonnegative. It follows that they are zero. Put $E=-\lambda \cdot \mu \cdot \Delta$. Note that $E \in \mathbb{Z}$, and for all $n \geqslant 0$

$$
G_{n+1}^{2}-A \cdot G_{n} \cdot G_{n+1}+B \cdot G_{n}^{2}=E \cdot B^{n}
$$

Suppose that $p_{i} \mid E$, then we infer that $p_{i} \nless G_{n}$ for all $n$, since $\left(G_{0}, G_{1}\right)=1$. Hence $m_{i}=0$. Next suppose $p_{i} \nmid E$, then

$$
\operatorname{ord}_{p_{i}}(\lambda \cdot V \Delta)+\operatorname{ord}_{p_{i}}(\mu \cdot V \Delta)=\operatorname{ord}_{p_{i}}(E)=0
$$

Since $\lambda \cdot \sqrt{ } \Delta$ and $\mu \cdot \sqrt{ } \Delta$ are algebraic integers (note that $V \Delta=\alpha-\beta$ ), the result follows.

From Lemma 2.1 it follows that we may assume without loss of generality that (4.6) holds for $i=1, \ldots$, $s$. We may also assume that $\operatorname{ord}_{p_{i}}(w)=0$ for i = 1, ..., s. The special case $s=0$ in equation (4.1) is trivial if $\Delta>0$, and will be treated implicitly in the next section for all $\Delta$.

### 4.3. The growth of the recurrence sequence.

First we treat the hyperbolic case $\Delta>0$. Note that $|\alpha| \neq|\beta|$, since the sequence is not degenerate. So we may assume $|\alpha|>|\beta|$. We have the following, almost trivial, result on the exponentiality of the growth of the sequence $\left\{G_{n}\right\}_{n=0}^{\infty}$. Let

$$
\begin{aligned}
& n_{0}>\max \left(2, \log \left|\frac{\mu}{\lambda}\right| / \log \left|\frac{\alpha}{\beta}\right|\right), \\
& \gamma=|\lambda|-|\mu| \cdot\left|\frac{\alpha}{\beta}\right|^{-n_{0}} .
\end{aligned}
$$

Note that $\gamma>0$.

LEMMA 4.2. Let $\Delta>0$. If $n \geqslant n_{0}$ then $\left|G_{n}\right| \geqslant \gamma \cdot|\alpha|^{n}$.

Proof. By (4.5), $|\alpha|>|\beta|$ and $n_{0}>0$ it follows for $n \geqslant n_{0}$ that

$$
\left|G_{\mathrm{n}}\right| \cdot|\alpha|^{-\mathrm{n}}=\left|\lambda+\mu \cdot\left(\frac{\alpha}{\beta}\right)^{-\mathrm{n}}\right| \geqslant|\lambda|-|\mu| \cdot\left|\frac{\alpha}{\beta}\right|^{-\mathrm{n}} \geqslant \gamma .
$$

We apply this to (4.1) as follows.

COROLLARY 4.3. Let $\Delta>0$. Any solution $n, m_{1}, \ldots, m_{s}$ of (4.1) with $n \geqslant n_{0}$ satisfies

$$
n<\sum_{i=1}^{S} m_{i} \cdot \frac{\log p_{i}}{\log |\alpha|}-\frac{\log (\gamma /|w|)}{\log |\alpha|} .
$$

Proof. Clear, from Lemma 4.2 and (4.1).

Next we study the elliptic case $\Delta<0$. Since $\alpha / \beta$ is not a root of unity, $B \geqslant 2$. Since $(\alpha, \beta)$ and $(\lambda, \mu)$ are pairs of complex conjugates, $|\alpha|=|\beta|$ and $|\lambda|=|\mu|$. Let $v \in \mathbb{R}, v \geqslant 1$ be given. We study the inequality

$$
\begin{equation*}
\left|G_{n}\right| \leqslant v \tag{4.7}
\end{equation*}
$$

in the variable $n \in \mathbb{N}_{0}$. We apply a result of Waldschmidt (see Section 2.3) from the complex theory of linear forms in logarithms, which gives an upper bound for $n$ that is particularly good in $v$. See also Kiss [1979]. Let

$$
\begin{aligned}
& \mathrm{E}=-\lambda \cdot \mu \cdot \Delta, \\
& \mathrm{U}_{2}= \frac{1}{2} \cdot \max (\pi, \log \mathrm{~B}), \mathrm{U}_{3}=\frac{1}{2} \cdot \max (\pi, \log \mathrm{E}), \\
& \mathrm{U}_{2}^{+}= \min \left(\mathrm{U}_{2}, \mathrm{U}_{3}\right), \mathrm{U}_{3}^{+}=\max \left(\mathrm{U}_{2}, \mathrm{U}_{3}\right), \\
& \mathrm{C}_{1}= 3.362 \times 10^{21} \cdot \mathrm{U}_{2} \cdot \mathrm{U}_{3} \cdot \log \left(2 \cdot \mathrm{e} \cdot \mathrm{U}_{2}^{+}\right), \mathrm{C}_{2}=\log \left(4 \cdot \mathrm{e} \cdot \mathrm{U}_{3}^{+}\right), \\
& \mathrm{C}_{3}= \max \left(\log (\pi / 2 \cdot|\mu|)+\mathrm{C}_{1} \cdot \mathrm{C}_{2}+\mathrm{C}_{1} \cdot \log \left(4 \cdot \mathrm{C}_{1} / \log \mathrm{B}\right),\right. \\
&\left.\frac{1}{2} \cdot \log |\lambda \cdot V \Delta|\right) \cdot 4 / \log \mathrm{B} .
\end{aligned}
$$

THEOREM 4.4. Let $\Delta<0, v \in \mathbb{R}, v \geqslant 1$. If $n \geqslant 0$ satisfies (4.7) then

$$
\mathrm{n}<\mathrm{C}_{3}+\frac{4}{\log \mathrm{~B}} \cdot \log \mathrm{v}
$$

Remark. Note that $C_{3}$ does not depend on $v$.

The following corollary of Theorem 4.4 is immediate.

COROLLARY 4.5. Let $\Delta<0$. Any solution $n, m_{1}, \ldots, m_{s}$ of (4.1) satisfies

$$
n<C_{3}+\frac{4}{\log B} \cdot\left|\left[\log |w|+\sum_{i=1}^{S} m_{i} \cdot \log p_{i}\right]\right|
$$

Proof (of theorem 4.4). Note that $|\alpha|=|\beta|=\sqrt{ } B \geqslant \sqrt{ } 2$. First we treat the case $G_{n}=0$. Kiss [1979] gives an upper bound for such $n$, but since in our situation $\left(G_{0}, G_{1}\right)=\left(G_{1}, B\right)=(A, B)=1$, we can do much better. Namely, put $R_{n}=\left(\alpha^{n}-\beta^{n}\right) /(\alpha-\beta)$ for all $n \in \mathbb{Z}$. It is easy to show that $R_{n} \in \mathbb{Z}$
and $R_{-n}=-B^{-n} \cdot R_{n}$ for all $n \in \mathbb{Z}$. Now $G_{n_{0}}=\lambda \cdot \alpha^{n_{0}}+\mu \cdot \beta^{n_{0}}=0$ implies

$$
\begin{aligned}
G_{n}= & \lambda \cdot \alpha^{n_{0}} \cdot \alpha{ }^{n-n_{0}}+\mu \cdot \beta^{n_{0}} \cdot \beta^{n-n_{0}}=\lambda \cdot \alpha{ }^{n_{0}} \cdot \sqrt{ } \cdot R_{n-n_{0}} \\
= & -\lambda \cdot \beta^{-n_{0}} \cdot v_{\Delta} \cdot B^{n} \cdot R_{n_{0}-n} .
\end{aligned}
$$

Thus we have

$$
G_{0}=\left|\left[-\lambda \cdot \beta^{-n_{0}} \cdot \sqrt[V]{ }\right]\right| \cdot R_{n_{0}} \quad, \quad G_{1}=\left|\left[-\lambda \cdot \beta^{-n_{0}} \cdot \sqrt[V]{ }\right]\right| \cdot B \cdot R_{n_{0}-1} .
$$

Suppose that $\mathfrak{p} \mid\left(R_{n}, B \cdot R_{n-1}\right)$ for some prime ideal $\mathfrak{p}$ in $\mathbb{Q}(\sqrt{ } \Delta)$. Then $\mathfrak{p} \mid\left(\alpha \cdot R_{n}-B \cdot R_{n-1}\right)=(\alpha)^{n}$, and $\mathfrak{p} \mid\left(\beta \cdot R_{n}-B \cdot R_{n-1}\right)=(\beta)^{n}$, which contradicts $(A, B)=1$. Thus $\left(R_{n}, B \cdot R_{n-1}\right)=1$, and then by $\left(G_{0}, G_{1}\right)=1$ we must have

$$
\left|\lambda \cdot \beta^{-n_{0}} \cdot \vee \Delta\right|=1 .
$$

Thus we find that $G_{n}=0$ implies

$$
n=\frac{2}{\log B} \cdot \log |\lambda \cdot V \Delta|<C_{3} .
$$

Now we turn to the case $G_{n} \neq 0$. We have from (4.7)

$$
\begin{equation*}
\left|\left(\frac{-\lambda}{\mu}\right) \cdot\left(\frac{\alpha}{\beta}\right)^{\mathrm{n}}-1\right| \leqslant \frac{\mathrm{v}}{|\mu|} \cdot \mathrm{B}^{-\mathrm{n} / 2} . \tag{4.8}
\end{equation*}
$$

We may assume $n \geqslant 2$. Let $-\lambda / \mu=e^{2 \pi i \cdot \psi}, \quad \alpha / \beta=e^{2 \pi i \cdot \varphi}$, with $-\frac{1}{2}<\psi \leqslant \frac{1}{2}$ and $-\frac{1}{2}<\varphi \leqslant \frac{1}{2}$. Let $k \in \mathbb{Z}$ be such that $|\psi+n \cdot \varphi+k| \leqslant \frac{1}{2}$. Then $|k| \leqslant 1+\frac{1}{2} \cdot n \leqslant n$. Put

$$
\Lambda=2 \pi i \cdot(\psi+n \cdot \varphi+k)=\log \left(\frac{-\lambda}{\mu}\right)+n \cdot \log \left(\frac{\alpha}{\beta}\right)+2 \cdot k \cdot \log (-1) .
$$

By lemma 2.3 and (4.8) we have an upper bound for $|\Lambda|$ :

$$
\begin{array}{r}
|\Lambda|=2 \pi \cdot|\psi+n \cdot \varphi+k| \leqslant \frac{1}{2} \pi \cdot\left|e^{2 \pi i \cdot(\psi+n \cdot \varphi+\mathrm{k})}-1\right| \\
=\frac{1}{2} \pi \cdot\left|\left(\frac{-\lambda}{\mu}\right) \cdot\left(\frac{\alpha}{\beta}\right)^{\mathrm{n}}-1\right| \leqslant \frac{1}{2} \pi \cdot \frac{\mathrm{v}}{|\mu|} \cdot \mathrm{B}^{-\mathrm{n} / 2} . \tag{4.9}
\end{array}
$$

From $G_{n} \neq 0$ we derive $\Lambda \neq 0$. Then from lemma 2.4 we can derive a lower bound for $|\Lambda|$. Note that $\max (n, 2|k|) \leqslant 2 \cdot n$, so that $W=\log (2 \cdot n)$. We choose $V_{1}=\frac{1}{2}$. The number $z=\alpha / \beta$ satisfies

$$
B \cdot z^{2}-\left(A^{2}-2 \cdot B\right) \cdot z+B=0,
$$

hence $h(\alpha / \beta) \leqslant \frac{1}{2} \cdot \log B$. And $z=-\lambda / \mu$ satisfies

$$
E \cdot z^{2}-\left(2 \cdot E+\Delta \cdot G_{0}^{2}\right) \cdot z+E=0
$$

hence $h(-\lambda / \mu) \leqslant \frac{1}{2} \cdot \log E$. Thus $V_{2}=U_{2}^{+}, V_{3}=U_{3}^{+}$satisfy the requirements for Theorem 2.4. We find

$$
\begin{align*}
|\Lambda|>\exp & \left(-C_{1} \cdot\left(\log (2 \cdot n)+\log \left(2 \cdot e \cdot U_{3}^{+}\right)\right)\right)  \tag{4.10}\\
& =\exp \left(-C_{1} \cdot\left(\log n+C_{2}\right)\right)
\end{align*}
$$

Combining (4.9) and (4.10) we find $n<a+b \cdot \log n$, where

$$
\begin{aligned}
& \mathrm{a}=\frac{2}{\log \mathrm{~B}} \cdot\left(\log \mathrm{v}+\log \frac{\pi}{2 \cdot|\mu|}+\mathrm{C}_{1} \cdot \mathrm{C}_{2}\right), \\
& \mathrm{b}=2 \cdot \mathrm{C}_{1} / \log \mathrm{B} .
\end{aligned}
$$

The result now follows from Lemma 2.1, since

$$
b=2 \cdot C_{1} / \log B=1.681 \times 10^{21} \cdot \frac{\max (\pi, \log B)}{\log B} \cdot \max (\pi, \log E) \cdot \log \left(2 \cdot e \cdot U_{2}^{+}\right)
$$

which is certainly larger than $e^{2}$.

Remark. Note that $v$ may depend on $n$. Thus we can find an upper bound for the solutions $n \in \mathbb{N}_{0}$ of e.g. $\left|G_{n}\right| \leqslant n^{c}$ for any constant $c$.

We now want to reduce the bound found in Theorem 4.3. We do this by studying the diophantine inequality

$$
\begin{equation*}
|\psi+\mathrm{n} \cdot \varphi+\mathrm{k}|<\mathrm{v}_{0} \cdot \mathrm{~B}^{-\mathrm{n} / 2} \tag{4.11}
\end{equation*}
$$

which follows from (4.9), where $v_{0}=v / 4 \cdot|\mu|$. We have to distinguish between the homogeneous case $\psi=0$ and the inhomogeneous case $\psi \neq 0$. We apply the methods that have been described in Sections 3.2 and 3.3 respectively. Unlike in other chapters, here we give the results in the form of precisely defined algorithms.

First we study the homogeneous case $\psi=0$. We then use Algorithm H (see the next page). Let $N$ be an upper bound for $n$ for the solutions of (4.11), for example the bound found in Theorem 4.3.

Input: $\varphi, \mathrm{B},|\mu|, \mathrm{v}_{0}, \mathrm{~N}$.
Output: new, reduced bound $N^{*}$ for $n$.
(i) (initialization) Choose $n_{0} \geqslant 2 / \log B$ such that $B^{n_{0} / 2} / n_{0} \geqslant 2 \cdot v_{0}$; $\mathrm{N}_{0}:=\mathrm{N}$; compute the continued fraction

$$
|\varphi|=\left[0, a_{1}, a_{2}, \ldots, a_{\ell_{0}+1}, \ldots\right]
$$

and the denominators $q_{1}, \ldots, q_{\ell_{0}+1}$ of the convergents of $|\varphi|$, with $\ell_{0}$ so large that $\mathrm{q}_{\ell_{0}} \leqslant \mathrm{~N}_{0}<\mathrm{q}_{\ell_{0}+1} ; \mathrm{i}:=0$;
(ii) (compute new bound) $A_{i}:=\max \left(a_{1}, \ldots, a_{\ell_{i}}\right.$ ) ; compute the largest integer $N_{i+1}$ such that

$$
B^{N_{i+1} / 2} / N_{i+1} \leqslant v_{0} \cdot\left(A_{i}+2\right)
$$

and $\ell_{i+1}$ such that $q_{\ell_{i+1}} \leqslant N_{i+1}<q_{\ell_{i+1}+1}$;
(iii) (terminate loop)

$$
\begin{aligned}
& \text { if } n_{0} \leqslant N_{i+1}<N_{i} \text { then } i:=i+1 \text {, goto (ii); } \\
& \text { else } N^{*}:=\max \left(n_{0}, N_{i+1}\right) \text {, stop. }
\end{aligned}
$$

Figure 4. ALGORITHM H. (reduces upper bound for (4.11) in the case $\psi=0$ ).

LEMMA 4.6. Algorithm H terminates. Inequality (4.11) with $\psi=0$ has no solutions with $\mathrm{N}^{*}<\mathrm{n}<\mathrm{N}$.

Proof. Termination is obvious, since all $N_{i}$ are integers. Note that $B^{x / 2} / x$ is an increasing function for $x \geqslant 2 / \log B$. Hence, if $n \geqslant n_{0}$,

$$
|\varphi|-|k| / n \mid \leqslant v_{0} \cdot B^{-n / 2} / n<1 / 2 n^{2}
$$

It follows (cf. (3.6)) that $|k| / n$ is a convergent of $|\varphi|$, say $|k| / n=p_{m} / q_{m}$. Then $q_{m} \leqslant n$, and (cf. (3.5)),

$$
\left||\varphi|-p_{m} / q_{m}\right|>1 /\left(a_{m+1}+2\right) \cdot q_{m}^{2}
$$

Suppose $n \leqslant N_{i}$ for some $i \geqslant 0$. Then $m \leqslant \ell_{i}$. Hence,

$$
B^{n / 2} / n \leqslant v_{0} \cdot n^{-2} \cdot| | \varphi|-|k| / n|^{-1}<v_{0} \cdot\left(a_{m+1}+2\right) \leqslant v_{0} \cdot\left(A_{m}+2\right)
$$

It follows that if $N_{i+1} \geqslant n_{0}$ then $n \leqslant N_{i+1}$.

Next we study the inhomogeneous case $\psi \neq 0$. Again, let $N$ be an upper bound for $n$ satisfying (4.11). We now have the following Algorithm $I$.

Input: $\varphi, \psi, B, v_{0}, N$.
Output: new, reduced upper bound $N^{*}$ for all but a finite number of explicitly given n.
(i) (initialization) $\mathrm{N}_{0}:=[\mathrm{N}]$; compute the continued fraction

$$
|\varphi|=\left[0, a_{1}, a_{2}, \ldots, a_{\ell}, \ldots\right]
$$

and the convergents $p_{i} / q_{i}$ for $i=1, \ldots, \ell_{0}$, with $\ell_{0}$ so large that $\mathrm{q}_{\ell_{0}}>4 \cdot \mathrm{~N}_{0}$ and $\mathrm{Il}_{\ell_{0}} \cdot \psi \|>2 \cdot \mathrm{~N}_{0} / \mathrm{q}_{\ell_{0}}$. (If such $\ell_{0}$ cannot be found within reasonable time, take $\ell_{0}$ so large that
$\left.\mathrm{q}_{\ell_{0}}>4 \cdot \mathrm{~N}_{0}\right) ; \quad \mathrm{i}:=0 ;$
(ii) (compute new bound)
if $\mathrm{llq}_{\ell} \cdot \psi \|>2 \cdot \mathrm{~N}_{\mathrm{i}} / \mathrm{q}_{\ell}{ }_{\mathrm{i}}$

$$
\begin{aligned}
& \text { then } N_{i+1}:=\left[2 \cdot \log \left(q_{\ell}^{2} \cdot v_{0} / N_{i}\right) / \log B\right] \text {; } \\
& \text { else compute } \mathrm{K} \in \mathbb{Z} \text { with }\left|\mathrm{K}-\mathrm{q}_{\ell} \cdot \psi\right| \leqslant \frac{1}{2} \text {; compute } \\
& \mathrm{n}_{0} \in \mathbb{Z}, \quad 0 \leqslant \mathrm{n}_{0}<\mathrm{q}_{\ell_{\mathrm{i}}} \text {, with } \mathrm{K}=\mathrm{n}_{0} \cdot \mathrm{p}_{\ell_{\mathrm{i}}} \equiv 0\left(\bmod \mathrm{q}_{\ell_{\mathrm{i}}}\right) \text {; } \\
& \text { if } n=n_{0} \text { is a solution of (4.11), then print an } \\
& \text { appropriate message; } \\
& N_{i+1}:=\left[2 \cdot \log \left(4 \cdot q_{\ell_{i}} \cdot v_{0}\right) / \log B\right] ;
\end{aligned}
$$

(iii) (terminate loop)

$$
\begin{aligned}
& \text { if } N_{i+1}<N_{i} \\
& \text { then } i:=i+1 \text {; compute the minimal } \ell_{i}<\ell_{i-1} \text { such that } \\
& \mathrm{q}_{\ell_{i}}>4 \cdot \mathrm{~N}_{\mathrm{i}} \text { and } \mathrm{Ilq}_{\ell_{\mathrm{i}}} \cdot \psi \|>2 \cdot \mathrm{~N}_{\mathrm{i}} / \mathrm{q}_{\ell_{\mathrm{i}}} \text { (if such } \ell_{\mathrm{i}} \text { does } \\
& \text { not exist, choose the minimal } \left.\ell_{i} \text { with } q_{\ell_{i}}>4 \cdot N_{i}\right) \text {; } \\
& \text { goto (ii) ; } \\
& \text { else } N^{*}:=N_{i} \text {; stop. }
\end{aligned}
$$

Figure 5. ALGORITHM I. (reduces upper bound for (4.11) in the case $\psi \neq 0$ ).

LEMMA 4.7. Algorithm I terminates. Inequality (4.11) with $\psi \neq 0$ has for $\mathrm{N}^{*}<\mathrm{n}<\mathrm{N}$ only the finitely many solutions found by the algorithm.

Proof. It is clear that the algorithm terminates. Suppose that $n \leqslant N_{i}$ for
some $i \geqslant 0$. Then if $\operatorname{llq}_{\ell_{i}} \cdot \psi \|>2 \cdot N_{i} / q_{\ell_{i}}$, we have

$$
\begin{aligned}
&{ }^{\| l q_{\ell}} \\
&=\psi \| \ell_{i} \\
& \leqslant q_{\ell_{i}} \cdot|\psi+n \cdot \varphi+k|+n / q_{\ell_{i}} \leqslant q_{\ell_{i}} \cdot v_{0} \cdot B^{-n / 2}+N_{i} / q_{\ell_{i}}
\end{aligned} .
$$

It follows that $n \leqslant N_{i+1}$. If $\operatorname{llq}_{\ell_{i}} \cdot \psi \| \leqslant 2 \cdot N_{i} / q_{\ell_{i}}$, then

$$
\begin{gathered}
\left|K+n \cdot p_{\ell_{i}}+k \cdot q_{\ell_{i}}\right| \leqslant\left|K-q_{\ell_{i}} \cdot \psi\right|+q_{\ell_{i}} \cdot|\psi+n \cdot \varphi+k|+n \cdot\left|p_{\ell_{i}}-q_{\ell} \cdot \varphi\right| \\
\leqslant \frac{1}{2}+q_{\ell_{i}} \cdot v_{0} \cdot B^{-n / 2}+N_{i} / q_{\ell}<\frac{3}{4}+q_{\ell} \cdot v_{0} \cdot B^{-n / 2}
\end{gathered}
$$

If $q_{\ell_{i}} \cdot v_{0} \cdot B^{-n / 2} \leqslant \frac{1}{4}$, then $K+n \cdot p_{\ell_{i}}+k \cdot q_{\ell_{i}}=0$, since it is an integer. By $\left(p_{\ell_{i}}, q_{\ell_{i}}\right)=1$ it follows that $n \equiv n_{0}\left(\bmod q_{\ell}\right)$. Since $q_{\ell_{i}}>N_{i}$, the only possibility is $n=n_{0}$. If $q_{\ell} \cdot v_{0} \cdot B^{-n / 2}>\frac{1}{4}^{i}$, then $n \leqslant N_{i+1}$ follows immediately.

We remark that in practice one almost always finds an $\ell_{i}$ such that ${ }^{\| l q_{\ell}}{ }_{i} \cdot \psi \|>2 \cdot N_{i} / q_{\ell_{i}}$, if $N_{i}$ is large enough.

### 4.4. Upper bounds.

In this section we will derive explicit upper bounds for the solutions of (4.1), both in the hyperbolic and elliptic cases. Our first step is the application of the p-adic theory of linear forms in logarithms, which works the same way in both cases. We use it to find a bound for $m_{i}$ that is polynomial in $\log \mathrm{n}$. Then we combine this with the results of Section 4.3 on the growth of the recurrence sequence, which for the solutions of (4.1) yield a bound for $n$ that is linear in the $m_{i}$ (Corollaries 4.3 and 4.5).

Assume that $n_{0} \geqslant 2$. Let $D$ be the discriminant of $\mathbb{Q}(\sqrt{ } \Delta)$. Put

$$
L=\log \max \left(|e \cdot D|^{1 / 4},|\alpha \cdot \lambda \cdot \sqrt{ } \Delta|,|\alpha \cdot \mu \cdot \sqrt{ } \Delta|,|\beta \cdot \lambda \cdot \sqrt{ } \Delta|,|\beta \cdot \mu \cdot \sqrt{ } \Delta|\right)
$$

Let $d$ be the squarefree part of $\Delta$. For $i=1, \ldots$, s put

$$
\varphi_{\mathrm{i}}=2 \text { if } \mathrm{p}_{\mathrm{i}} \mathrm{l} \mathrm{~d}, \quad \varphi_{\mathrm{i}}=1 \text { otherwise, }
$$

$$
\begin{aligned}
& \rho_{i}=2 \text { if } p_{i}=2, d \equiv 5(\bmod 8) \text { or if } p_{i}>2,\left(\frac{d}{p_{i}}\right)=-1, \\
& \quad \rho_{i}=1 \text { otherwise, } \\
& C_{4, i}=10^{6} \cdot\left(\frac{2}{\rho_{i} \cdot \log p_{i}}\right)^{7} \cdot \varphi_{i}^{-3} \cdot L^{4} \cdot p_{i}^{4 \cdot \rho_{i}+4} \cdot\left(1+\frac{\varphi_{i} \cdot L \cdot p_{i}+2 / L}{\log n_{0}}\right)^{3} .
\end{aligned}
$$

LEMMA 4.8. The solutions of (4.1) with $n \geqslant n_{0}$ satisfy

$$
m_{i}<C_{4, i} \cdot(\log n)^{3} \text { for } i=1, \ldots, s
$$

Proof. Rewrite (4.1), using (4.5), as

$$
\left(\frac{\alpha}{\beta}\right)^{n}-\left(\frac{-\mu}{\lambda}\right)=\frac{w}{\lambda} \cdot \beta^{-n} \cdot \prod_{i=1}^{s} p_{i}^{m_{i}}
$$

Then, by (4.6),

$$
m_{i} \leqslant m_{i}-\operatorname{ord}_{p_{i}}(\lambda)=\operatorname{ord}_{p_{i}}\left(\frac{w}{\lambda} \cdot \beta^{-n} \cdot \prod_{i=1}^{s} p_{i}^{m}\right)=\operatorname{ord}_{p_{i}}\left(\left(\frac{\alpha}{\beta}\right)^{n}-\left(\frac{-\mu}{\lambda}\right)\right) .
$$

Apply Lemma 2.5 (Schinzel's result) with $\xi^{\prime \prime}=\alpha, \xi^{\prime}=\beta, \chi \prime=\mu \cdot \sqrt{\prime} \Delta$, $\chi^{\prime}=-\lambda \cdot V \Delta$. Then we find, using $\operatorname{ord}_{\mathfrak{p}_{i}}(\cdot)=\varphi_{i} \cdot \operatorname{ord}_{p_{i}}(\cdot)$,

$$
m_{i}<10^{6} \cdot\left(\frac{2}{\rho_{i} \cdot \log p_{i}}\right)^{7} \cdot \varphi_{i}^{-3} \cdot L^{4} \cdot p_{i}^{4} \cdot \rho_{i}^{+4} \cdot\left(\log n+\varphi_{i} \cdot L \cdot p_{i}^{\rho_{i}}+2 / L\right)^{3},
$$

from which the result follows, since $n \geqslant n_{0}$.

Put

$$
C_{4}=\max _{i}\left(C_{4, i}\right), \quad m=\max _{i}\left(m_{i}\right), \quad P=\prod_{i=1}^{s} p_{i} .
$$

In the case $\Delta>0$, let $n_{0}>\max (2, \log |\lambda / \mu| / \log |\alpha / \beta|)$, and put

$$
\begin{aligned}
& C_{5}=\log P /(\log |\alpha|+\min (0, \log (\gamma /|w|))), \\
& C_{6}=\max \left(8 \cdot C_{4} \cdot\left(\log 27 \cdot C_{4} \cdot C_{5}\right)^{3}, 841 \cdot C_{4}\right) .
\end{aligned}
$$

In the case $\Delta<0$, put

$$
C_{7}=\max \left\{C_{3}+\frac{4}{\log B} \cdot \log \left(2 \cdot\left|G_{0} \cdot \mu \cdot \sqrt{ }\right|\right),\right.
$$

$$
\begin{aligned}
& \left.\quad 8 \cdot\left(\left(\mathrm{C}_{3}+\frac{4 \cdot \log |\mathrm{w}|}{\log \mathrm{B}}\right)^{1 / 3}+\left(\frac{4 \cdot \mathrm{C}_{4} \cdot \log \mathrm{P}}{\log \mathrm{~B}}\right)^{1 / 3} \cdot \log \left(\frac{108 \cdot \mathrm{C}_{4} \cdot \log \mathrm{P}}{\log \mathrm{~B}}\right)\right)^{3}\right\}, \\
& \mathrm{C}_{8, i}=\mathrm{C}_{4, i} \cdot\left(\log \mathrm{C}_{7}\right)^{3} \text { for } \mathrm{i}=1, \ldots, \mathrm{~s} .
\end{aligned}
$$

Then we have the following result, giving explicit upper bounds for the solutions of (4.1).

THEOREM 4.9. Let $n, m_{1}, \ldots, m_{s}$ be a solution of (4.1).
(i). If $\Delta>0$ and $n \geqslant n_{0}$ then $n<C_{5} \cdot \mathrm{C}_{6}$ and $\mathrm{m}<\mathrm{C}_{6}$.
(ii). If $\Delta<0$ then $\mathrm{n}<\mathrm{C}_{7}$ and $\mathrm{m}_{\mathrm{i}}<\mathrm{C}_{8, \mathrm{i}}$ for $\mathrm{i}=1$, ..., s .

Proof. (i). Corollary 4.3 yields $n<C_{5} \cdot m$. By Lemma 4.8 we now have

$$
m<C_{4} \cdot(\log n)^{3}<C_{4} \cdot\left(\log C_{5} \cdot m\right)^{3} .
$$

If $C_{4} \cdot C_{5}>\left(e^{2} / 3\right)^{3}$, we apply Lemma 2.1 with $a=0, b=C_{4} \cdot C_{5}, h=3$, and we find $m<8 \cdot C_{4} \cdot\left(\log 27 \cdot C_{4} \cdot C_{5}\right)^{3}$. If $C_{4} \cdot C_{5} \leqslant\left(e^{2} / 3\right)^{3}$, then

$$
n<C_{5} \cdot m<C_{4} \cdot C_{5} \cdot(\log n)^{3} \leqslant\left(e^{2} / 3\right)^{3} \cdot(\log n)^{3},
$$

from which we deduce $n<12564$. Now, $m<C_{4} \cdot(\log n)^{3}<841 \cdot C_{4}$. (ii). From Lemma 4.8 and Corollary 4.5 we see that

$$
\mathrm{n}<\mathrm{C}_{3}+\frac{4}{\log \mathrm{~B}} \cdot \log \left(2 \cdot\left|\mathrm{G}_{0} \cdot \mu \cdot \sqrt{ } \Delta\right|\right),
$$

or

$$
n<C_{3}+\frac{4 \cdot \log |w|}{\log B}+\frac{4 \cdot C_{4} \cdot \log P}{\log B} \cdot(\log n)^{3}
$$

The result now follows from Lemma 2.1, since $4 \cdot C_{4} \cdot \log P / \log B>\left(e^{2} / 3\right)^{3}$. a

### 4.5. A basic lemma.

We introduce some notation, and then give an almost trivial lemma that is at the heart of our reduction methods for both the hyperbolic and the elliptic cases. Let for $\mathrm{i}=1, \ldots$, s

$$
\begin{aligned}
& e_{i}=-\operatorname{ord}_{p_{i}}(\lambda), \quad f_{i}=\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}\left(\frac{\alpha}{\beta}\right)\right), \quad g_{i}=f_{i}-e_{i}, \\
& \vartheta_{i}=-\log _{p_{i}}\left(\frac{-\lambda}{\mu}\right) / \log _{p_{i}}\left(\frac{\alpha}{\beta}\right) .
\end{aligned}
$$

By Lemma 4.1 the $p_{i}$-adic logarithms of $\alpha / \beta$ and $-\lambda / \mu$ exist. Note that $\log _{p_{i}}(\alpha / \beta) \neq 0$, since the sequence $\left\{G_{n}\right\}$ is not degenerate. Note that for conjugated $\xi, \quad \xi$ also $\log _{p} \xi$ and $\log _{p} \xi$, are conjugates, hence $\log _{p}\left(\xi / \xi^{\prime}\right) \in \sqrt{ } \Delta \cdot \mathbb{Q}_{p}$. Hence both numerator and denominator of $\vartheta_{i}$ are in $\sqrt{ } \triangle \cdot \mathbb{Q}_{p_{i}}$, so $\vartheta_{i} \in \mathbb{Q}_{p_{i}}$. Hence, if $\vartheta_{i} \neq 0$, we can write

$$
\vartheta_{i}=\sum_{\ell=\mathrm{k}_{\mathrm{i}}}^{\infty} \mathrm{u}_{\mathrm{i}, \ell} \cdot p_{\mathrm{i}}^{\ell},
$$

where $k_{i}=\operatorname{ord}_{p_{i}}\left(\vartheta_{i}\right) \quad$ and $u_{i, \ell} \in\left\{0,1, \ldots, p_{i}-1\right.$ for all $\ell$. The following lemma localizes the elements of $\left\{G_{n}\right\}$ with many factors $p_{i}$, in terms of the $p_{i}$-adic expansion of $\vartheta_{i}$.

LEMMA 4.10. Let $n \in \mathbb{N}_{0}$. If $\operatorname{ord}_{p_{i}}\left(G_{n}\right)+e_{i}>1 /\left(p_{i}-1\right)$ then

$$
\operatorname{ord}_{p_{i}}\left(G_{n}\right)=g_{i}+\operatorname{ord}_{p_{i}}\left(n-\vartheta_{i}\right)
$$

Proof. By Lemma 4.1 we have

$$
\operatorname{ord}_{p_{i}}\left(G_{n}\right)+e_{i}=\operatorname{ord}_{p_{i}}\left(\left(\frac{\alpha}{\beta}\right)^{n}-\left(\frac{-\mu}{\lambda}\right)\right)=\operatorname{ord}_{p_{i}}\left(\left(\frac{-\lambda}{\mu}\right) \cdot\left(\frac{\alpha}{\beta}\right)^{n}-1\right)
$$

With $\xi=(-\lambda / \mu) \cdot(\alpha / \beta)^{n}-1$ we have by assumption $\operatorname{ord}_{p_{i}}(\xi)>1 /\left(p_{i}-1\right)$.
Hence $\operatorname{ord}_{p_{i}}(\xi)=\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}(1+\xi)\right)$, and it follows that

$$
\begin{aligned}
\operatorname{ord}_{p_{i}}\left(G_{n}\right) & +e_{i}=\operatorname{ord}_{p_{i}}\left(n \cdot \log _{p_{i}}\left(\frac{\alpha}{\beta}\right)+\log _{p_{i}}\left(\frac{-\lambda}{\mu}\right)\right) \\
& =\operatorname{ord}_{p_{i}}\left(n-\vartheta_{i}\right)+f_{i}
\end{aligned}
$$

### 4.6. Trivial cases.

We have to exclude some trivial cases first. The first trivial case is that of $\operatorname{ord}_{p_{i}}\left(\vartheta_{i}\right)<0$. Then the solutions of (4.1) satisfy $m_{i} \leqslant 1 /\left(p_{i}-1\right)-e_{i}$, or, by Lemma 4.10,

$$
m_{i}=f_{i}-e_{i}+\operatorname{ord}_{p_{i}}\left(n-\vartheta_{i}\right)
$$

Since $n \in \mathbb{Z}$ and $\operatorname{ord}_{p_{i}}\left(\vartheta_{i}\right)<0$ we have $\operatorname{ord}_{p_{i}}\left(n-\vartheta_{i}\right)=\operatorname{ord}_{p_{i}}\left(\vartheta_{i}\right)$. Hence

$$
m_{i} \leqslant \max \left(f_{i}+\operatorname{ord}_{p_{i}}\left(\vartheta_{i}\right), 1 /\left(p_{i}-1\right)\right)-e_{i}
$$

The case where all $p_{i}$-adic digits of $\vartheta_{i}$ from a certain point on are all zero is a special case, because the reduction methods of the next sections then do not work. This is so because these reduction methods make use of zero-dimensional p-adic diophantine approximation, as explained in Section 3.9, applied to the p-adic linear form

$$
\log _{p}\left(\frac{\lambda}{\mu}\right)+n \cdot \log _{p}\left(\frac{\alpha}{\beta}\right)
$$

for $p=p_{1}, \ldots, p_{s}$. This means that we must study the $p$-adic number

$$
\vartheta=-\log _{\mathrm{p}}\left(\frac{\lambda}{\mu}\right) / \log _{\mathrm{p}}\left(\frac{\alpha}{\beta}\right) .
$$

If it happens that this number $\vartheta$ is zero, or that all digits in the p-adic expansion of $\vartheta$ are zero from a certain point on, then obviously the reduction process of Section 3.9 breaks down, since it is based on the assumption that the p-adic expansion of $\vartheta$ contains sufficiently many non-zero digits.

This case can be dealt with as follows. Note that $\vartheta_{i}=r$ holds for all $\mathrm{i}=1, \ldots$, s with the same r . Thus, by Lemma 4.10,

$$
\begin{equation*}
m_{i} \leqslant \max \left(g_{i}+\operatorname{ord}_{p_{i}}(n-r), 1-e_{i}\right) \leqslant g_{i}+1+\operatorname{ord}_{p_{i}}(n-r) \tag{4.12}
\end{equation*}
$$

Then we have, if $\Delta>0$, by Corollary 4.3,

$$
n \cdot \log |\alpha|<\sum_{i=1}^{S}\left(g_{i}+1\right) \cdot \log p_{i}-\log (\gamma /|w|)+\log |n-r|
$$

from which a good upper bound for $n$ can be derived (no application of the Gelfond-Baker theory is involved, so the constants are relatively small). And if $\Delta<0$, the proof of Lemma 4.11 below yields $\vartheta_{i}=0$, whence, by (4.12),

$$
\left|G_{n}\right|=|w| \cdot \prod_{i=1}^{s} p_{i}^{m_{i}} \leqslant v_{0} \cdot n
$$

for some constant $v_{0}$. Only minor changes in the results and algorithms of Section 4.3 suffice to deal with this inequality instead of (4.7).

There is however an elementary way of treating this case, using congruences only, that is guaranteed to work. We define the following special symmetric recurrences'. For $\alpha, \beta$ as defined in Section 4.2, let $d$ be the squarefree part of $\Delta$, and put

$$
R_{n}=\frac{\alpha^{n}-\beta^{n}}{\alpha-\beta}, \quad S_{n}=\alpha^{n}+\beta^{n}
$$

for $d=-1$ also

$$
\mathrm{T}_{\mathrm{n}}^{ \pm}=(1 \pm V(-1)) \cdot \alpha^{\mathrm{n}}+(1 \mp V(-1)) \cdot \beta^{\mathrm{n}}
$$

and for $d=-3$ also (with $\omega=\rho$ or $\bar{\rho}$ for $\rho=\frac{1}{2} \cdot(1+\sqrt{ }(-3)$ ) )

$$
\begin{aligned}
& \mathrm{U}_{\mathrm{n}}(\omega)=(1+\omega) \cdot \alpha^{\mathrm{n}}+(1+\bar{\omega}) \cdot \beta^{\mathrm{n}}, \\
& \mathrm{~V}_{\mathrm{n}}(\omega)=\omega \cdot \alpha^{\mathrm{n}}+\bar{\omega} \cdot \beta^{\mathrm{n}}
\end{aligned}
$$

for all $n \in \mathbb{Z}$. Note that

$$
\mathrm{T}_{\mathrm{n}}^{+} \cdot \mathrm{T}_{\mathrm{n}}^{-}=2 \cdot \mathrm{~S}_{2 \mathrm{n}}, \quad \mathrm{U}_{\mathrm{n}}(\omega) \cdot \mathrm{U}_{\mathrm{n}}(\bar{\omega}) \cdot \mathrm{R}_{\mathrm{n}}=3 \cdot \mathrm{R}_{3 \mathrm{n}}, \quad \mathrm{~V}_{\mathrm{n}}(\omega) \cdot \mathrm{V}_{\mathrm{n}}(\bar{\omega}) \cdot \mathrm{S}_{\mathrm{n}}=\mathrm{S}_{3 \mathrm{n}}
$$

We have the following lemma. We assume that $\operatorname{ord}_{\mathrm{p}}(\vartheta) \geqslant 0$.

LEMMA 4.11. If $\vartheta$ has only finitely many nonzero p-adic digits, then there exist an $r \in \mathbb{N}_{0}$ and $a k \in \mathbb{Q}$ such that $G_{n}=\kappa \cdot R_{n-r}$, or $G_{n}=\kappa \cdot S_{n-r}$, or (if $\mathrm{d}=-1$ ) $\mathrm{G}_{\mathrm{n}}=\kappa \cdot \mathrm{T}_{\mathrm{n}}^{ \pm}$, or (if $\mathrm{d}=-3$ ) $\mathrm{G}_{\mathrm{n}}=\kappa \cdot \mathrm{U}_{\mathrm{n}}(\omega)$ or $\kappa \cdot \mathrm{V}_{\mathrm{n}}(\omega)$, where $\omega=\rho$ or $\bar{\rho}$. Further, $r=0$ if $\Delta<0$.

Proof. By $\operatorname{ord}_{p}(\vartheta) \geqslant 0$ we have $\vartheta=r$ for some $r \in \mathbb{N}_{0}$. From the definition of $\vartheta$ we infer

$$
\log _{p}\left(\frac{\alpha}{\beta}\right)^{r} \cdot\left(\frac{-\lambda}{\mu}\right)=0
$$

hence $\eta=(\beta / \alpha)^{r} \cdot(\mu / \lambda)$ is a root of unity. It follows that we can write

$$
\mathrm{G}_{\mathrm{n}}=\lambda \cdot \alpha^{\mathrm{r}} \cdot\left(\alpha^{\mathrm{n}-\mathrm{r}}+\eta \cdot \beta^{\mathrm{n}-\mathrm{r}}\right)
$$

First let $B= \pm 1$. Then $\Delta>0$ and

$$
\begin{aligned}
& G_{0}=\lambda \cdot \alpha^{r} \cdot\left(\alpha^{-r} \pm \beta^{-r}\right)= \pm \lambda \cdot \alpha^{r} \cdot\left(\alpha^{r} \pm \beta^{r}\right) \\
& G_{1}=\lambda \cdot \alpha^{r} \cdot\left(\alpha^{1-r} \pm \beta^{1-r}\right)= \pm \lambda \cdot \alpha^{r} \cdot\left(\alpha^{r-1} \pm \beta^{r-1}\right)
\end{aligned}
$$

Note that

$$
\begin{aligned}
& \left(\alpha^{r-1}+\beta^{r-1}, \alpha^{r}+\beta^{r}\right)=(2, \alpha+\beta)=(1) \text { or }(2) \\
& \left(\alpha^{r-1}-\beta^{r-1}, \alpha^{r}-\beta^{r}\right)=(\alpha-\beta)
\end{aligned}
$$

By $\left(G_{0}, G_{1}\right)=1$ it follows that $\pm \lambda \cdot \alpha^{r}=1, \frac{1}{2}$ or $1 /(\alpha-\beta)$, respectively, and the assertion follows.
Next suppose $|B| \geqslant 2$. Then

$$
G_{0} \cdot B \cdot\left(\eta \cdot \alpha^{r-1}+\beta^{r-1}\right)=G_{1} \cdot\left(\eta \cdot \alpha^{r} \pm \beta^{r}\right)
$$

Since $\left(B, G_{1}\right)=1$, we have $\alpha \cdot \beta \mid \eta \cdot \alpha^{r} \pm \beta^{r}$. By $(A, B)=1$ we have $(\alpha, \beta)=(1)$, and from $\alpha \mid \beta^{r}$ it then follows that $\vartheta=r=0$. So $G_{0}=\lambda \cdot(1+\eta) \in \mathbb{Z}$. The result now follows easily, since for $\eta$ the only possibilities are $\pm 1$ for all $d$, and moreover $\pm \sqrt{ }(-1)$ if $d=-1$, and $\pm \rho, \pm \bar{\rho}$ if $\mathrm{d}=-3$.

In the cases of Lemma 4.11 we can treat (4.1) as follows. Lemma 4.10 shows that the smallest index $n=g\left(m \cdot p^{\ell}\right)>0$ such that $m \cdot p^{\ell} \mid G_{n}$ grows exponentially with $\ell$. Also, $G_{n}$ grows exponentially with $n$, as follows from Lemma 4.2 and Theorem 4.4. Hence $G_{g\left(m \cdot p^{\ell}\right)}$ grows doubly exponentially with $\ell$. It follows that $a=w \cdot p_{1}^{m_{1}} \cdot \ldots \cdot p_{s}^{m_{s}} \underset{m}{\text { cannot keep }} \underset{m}{ }$ up with $\quad G_{g(a)}$ as the $m_{i}$ tend to infinity. It follows that if $p_{1}{ }^{m} \cdot \ldots \cdot p_{s}{ }_{s}$ is large enough, there exists a prime $q$ such that $q \mid G_{g(a)}$ but $q \nmid a$. Now the sequences $\left\{R_{n}\right\},\left\{S_{n}\right\}$ have special divisibility properties, such as

$$
\begin{aligned}
& R_{n} \mid R_{m} \text { if and only if } n \mid m \\
& S_{n} \mid S_{k n} \text { for odd } k \\
& \operatorname{ord}_{2}\left(S_{n}\right) \leqslant \operatorname{ord}_{2}\left(S_{3}\right) \text { for all } n \geqslant 1 .
\end{aligned}
$$

Making use of this kind of properties it can be proved that $q \| G_{n}$ whenever a $I G_{n}$. This gives an upper bound for the solutions of (4.1), since for those solutions a $\mid G_{n}$ but $q \nmid G_{n}$. We give two examples.

Example. Let $A=16, B=1, G_{0}=1, G_{1}=8, w=1, p_{1}=2, p_{2}=11$. Then $\alpha=8+3 \cdot \sqrt{ } 7, \beta=8-3 \cdot \sqrt{ } 7, \lambda=\mu=\frac{1}{2}$, so $\lambda / \mu$ is a root of unity. Hence $\vartheta_{1}=\vartheta_{2}=0$. Note that we have a sequence of type $S_{n}$ here. We have

| $n$ |  | -3 | -2 | -1 | 0 | 1 | 2 |
| :--- | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| $G_{n}$ | 2024 | 127 | 8 | 1 | 8 | 127 | 2024 |
| $G_{n}(\bmod 16)$ | 8 | -1 | 8 | 1 | 8 | -1 | 8 |
| $G_{n}(\bmod 11)$ | 0 | 6 | 8 | 1 | 8 | 6 | 0 |
| $G_{n}\left(\bmod 11^{2}\right)$ | 88 | 6 | 8 | 1 | 8 | 6 | 88 |
| $G_{n}(\bmod 23)$ | 0 | 12 | 8 | 1 | 8 | 12 | 0 |

It follows by this table that $\operatorname{ord}_{2}\left(G_{n}\right)=0$ or 3 , according to $n$ even or odd, and $\operatorname{ord}_{11}\left(G_{n}\right)>0$ if and only if $n \equiv 3(\bmod 6)$. This can also be derived from Lemma 4.10, which yields: if $\operatorname{ord}_{2}\left(G_{n}\right) \geqslant 1$ (which happens exactly for odd $n$ ), then $\operatorname{ord}_{2}\left(G_{n}\right)=3+\operatorname{ord}_{2}(n)=3$. Further, if $\operatorname{ord}_{11}\left(G_{n}\right) \geqslant 1$ (which happens exactly when $n \equiv 3(\bmod 6)$ ), then $\operatorname{ord}_{11}\left(G_{n}\right)=1+\operatorname{ord}_{11}(n) \quad\left(e . g . \quad \operatorname{ord}_{11}\left(G_{33}\right)=2, \quad\right.$ but $\left.\operatorname{ord}_{11}\left(G_{11}\right)=0\right)$.

Now, $G_{3} \mid G_{3 k}$ holds for all odd $k$. Note that $G_{3}$ has exactly 3 factors 2 , and 1 factor 11 . But it is larger than $2^{3} \cdot 11=88$. Hence there is a prime $q$, distinct from 2 and 11 , such that $q \mid G_{n}$ whenever $11 \mathrm{I} \mathrm{G}_{\mathrm{n}}$. Thus $\mathrm{G}_{\mathrm{n}}=2^{\mathrm{m}_{1}} \cdot 11^{\mathrm{m}_{2}}$ has no solutions with $\mathrm{m}_{2} \neq 0$, so that there remain only three solutions: $n=-1,0,1$. Note that it is not necessary to know the value of $q$ explicitly. In this case it is 23 , and indeed it is easy to show directly that $23 \mid G_{n}$ if and only if $n \equiv 3(\bmod 6)$.

Example. Let $A=5, B=13, G_{0}=G_{1}=1$. Then $\Delta=-27, \alpha=1+3 \cdot \rho$, $\lambda=(1+\rho) / 3$. Then $\lambda / \bar{\lambda}=\rho$ is a root of unity, thus $\vartheta=0$. We will solve $G_{n}= \pm 2^{m}$. The sequence $G_{n}=\lambda \cdot \alpha^{n}+\bar{\lambda} \cdot \bar{\alpha}^{n}$ is related to the sequence $H_{n}=\bar{\lambda} \cdot \alpha^{n}+\lambda \cdot \bar{\alpha}^{n}$ and to $R_{n}=\left(\alpha^{n}-\bar{\alpha}^{n}\right) /(\alpha-\bar{\alpha})$ by $G_{n} \cdot H_{n} \cdot R_{n}=R_{3 n} / 3$. Since $R_{n}$ has nice divisibility properties, we have useful information on the prime divisors of $G_{n}$ and $H_{n}$. We find:

| n | 0 | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 |
| :--- | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| $G_{n}$ | 1 | 1 | -8 | -53 | -161 | -116 | 1513 | 9073 | 25696 |
| $H_{n}$ | 1 | 4 | 7 | -17 | -176 | -659 | -1007 | 3532 | 30751 |
| $R_{n}$ | 0 | 1 | 5 | 12 | -5 | -181 | -840 | -1847 | 1685 |

Now, $G_{n} \equiv 0(\bmod 16)$ if and only if $n \equiv 8(\bmod 12)$ (Lemma 4.10 yields: if $\operatorname{ord}_{2}\left(G_{n}\right) \geqslant 2$ (which happens exactly when $n \equiv 2(\bmod 3)$ ), then $\left.\operatorname{ord}_{2}\left(G_{n}\right)=2+\operatorname{ord}_{2}(n)\right), H_{n} \equiv 0(\bmod 16)$ if and only if $n \equiv 4(\bmod 12)$, and $R_{n} \equiv 0(\bmod 16)$ if and only if $n \equiv 0(\bmod 12)$. Note that
$G_{4} \cdot H_{4} \cdot R_{4}=R_{12} / 3=-2^{4} \cdot 5 \cdot 7 \cdot 11 \cdot 23$. Considering the sequences modulo $5,7,11$ and 23 we find that $2^{4} \cdot 7 \cdot 11 \cdot 23 \mid G_{n} \cdot H_{n}$ for all $n \equiv 0(\bmod 4)$, and in fact 11 I $G_{n}$ whenever $16 \mid G_{n}$. Thus $G_{n}= \pm 2^{m}$ implies $m \leqslant 3$. It follows from Section 4.3 how to solve $\left|G_{n}\right| \leqslant 8$.

We note that a process as described above can always be applied when dealing with a situation as in Lemma 4.11. This is guaranteed by Lemma 4. 10.

From now on we thus assume that $\operatorname{ord}_{p_{i}}\left(\vartheta_{i}\right) \geqslant 0$ for all $i=1, \ldots, s$, and that infinitely many $p_{i}-$ adic digits $u_{i, \ell}$ of $\vartheta_{i}$ are nonzero.

### 4.7. The reduction algorithm in the hyperbolic case.

First we give the reduction algorithm (Algorithm $P$, see the next page) for the case $\Delta>0$. It is based on Lemma 4.10 and Corollary 4.3 only. Let $N$ be an upper bound for $n$ for the solutions $n, m_{1}, \ldots, m_{s}$ of (4.1). For example, $N=C_{5} \cdot \mathrm{C}_{6}$ as in Theorem 4.9.

THEOREM 4.12. With all the above assumptions, Algorithm P terminates. Equation (4.1) with $\Delta>0$ has no solutions with $N^{*} \leqslant n<N, m_{i}>M_{i}$ for i $=1, \ldots, \mathrm{~s}$.

Proof. Since the $p_{i}$-adic expansion of $\vartheta_{i}$ is assumed to be infinite, there exist $r_{i}$ with the required properties. It is clear that $s_{i, 1} \leqslant r_{i}<s_{i, 0}$, and that $N_{j} \leqslant N_{j-1}$. So $\mathbf{s}_{\mathbf{i}, \mathbf{j}} \leqslant s_{i, j-1}$ holds for all $j \geqslant 1$. Since $s_{i, j} \geqslant 0$, there is a $j$ such that $N_{j} \leqslant n_{0}$ or $s_{i, j}=s_{i, j-1}$ for all $i=1, \ldots, s$. In the latter case, $K_{j}$ remains false. ; in both cases the algorithm terminates. We prove by induction on $j$ that $m_{i} \leqslant g_{i}+s_{i, j}$ for $\mathrm{i}=1, \ldots, \mathrm{~s}$, and $\mathrm{n}<\mathrm{N}_{\mathrm{j}}$ hold for all j . For $\mathrm{j}=0$, it is clear that $n<N_{0}$. Suppose $n<N_{j-1}$ for some $j \geqslant 1$. Suppose there exists an $i$ such that $m_{i}>g_{i}+s_{i, j}$. From Lemma 4.10 we have

$$
\operatorname{ord}_{p_{i}}\left(n-\vartheta_{i}\right)=m_{i}-g_{i} \geqslant s_{i, j}+1,
$$

hence, by $u_{i, s_{i, j}} \neq 0$,

$$
n \geqslant \sum_{\ell=0}^{s_{i}, j} u_{i, \ell} \cdot p^{\ell} \geqslant p^{s_{i}, j} \geqslant N_{j-1}
$$

Input: $\alpha, \beta, \lambda, \mu, w, p_{1}, \ldots, p_{s}, N$.
Output: new, reduced upper bounds $M_{i}$ for $m_{i}$ for $i=1, \ldots$, $s$, and $N^{*}$ for $n$.
(i) (initialization) Choose an $\mathrm{n}_{0} \geqslant 0$ such that
$\mathrm{n}_{0}>\log |\mu / \lambda| / \log |\alpha / \beta| ; \gamma:=|\lambda|-|\mu| \cdot|\alpha / \beta|^{-\mathrm{n}_{0}} ;$
$\left.g_{i}:=\operatorname{ord}_{p_{i}}(\lambda)+\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}(\alpha / \beta)\right)\right)$
$h_{i}:=\operatorname{ord}_{p_{i}}(\lambda)+\left\{\begin{array}{cl}3 / 2 & \text { if } p_{i}=2 \\ 1 & \text { if } \\ p_{i}=3 \\ 1 / 2 & \text { if } \quad p_{i} \geqslant 5\end{array}\right\}$ for $i=1, \ldots, s ;$
$g:=\gamma /|w| \cdot \prod_{i=1}^{s} p_{i}^{g_{i}} ; \quad N_{0}:=N ;$
(ii) (computation of the $\vartheta_{i}$ 's) Compute for $i=1$, ..., s the first $r_{i}$ $p_{i}$-adic digits $u_{i, \ell}$ of

$$
\vartheta_{i}=-\log _{p_{i}}\left(\frac{-\lambda}{\mu}\right) / \log _{p_{i}}\left(\frac{\alpha}{\beta}\right)=\sum_{\ell=0}^{\infty} u_{i}, \ell \cdot p_{i}^{\ell},
$$

where $r_{i}$ is so large that $p_{i}{ }_{i} \geqslant N_{0}$ and $u_{i, r_{i}} \neq 0$;
(iii) (further initialization, start outer loop) $s_{i, 0}:=r_{i}+1$ for $\mathrm{i}=1, \ldots, \mathrm{~s} ; \mathrm{j}:=1$;
(iv) (start inner loop) i := 1 ; $K_{j}:=$.false. ;
(v) (computation of the new bounds for $m_{i}$, terminate inner loop) $s_{i, j}:=\min \left\{s \in \mathbb{N}_{0} \mid p_{i}^{s} \geqslant N_{j-1}\right.$ and $\left.u_{i, s} \neq 0\right\} ;$
if $s_{i, j}<s_{i, j-1}$
then $K_{j}:=$ true. ;
if $i<s$

$$
\text { then } i:=i+1 \text {; goto (v) ; }
$$

(vi) (computation of the new bound for $n$, terminate outer loop)
$N_{j}:=\min \left(N_{j-1},\left(\sum_{i=1}^{S} s_{i, j} \cdot \log p_{i}-\log g\right) / \log |\alpha|\right) ;$
if $N_{j} \geqslant n_{0}$ and $K_{j}$
then $j:=j+1$; goto (iv) ;
else $N^{*}:=\max \left(N_{j}, n_{0}\right)$;

$$
M_{i}:=\max \left(h_{i}, g_{i}+s_{i, j}\right) \text { for } i=1, \ldots, s ; \text { stop. }
$$

Figure 6. ALGORITHM P. (reduces given upper bounds for (4.1) if $\Delta>0$ ).
which contradicts our assumption. Thus, $m_{i} \leqslant g_{i}+s_{i, j}$ for $i=1, \ldots, s$. Then from Corollary 4.3 it follows that

$$
n<\left(\sum_{i=1}^{s}\left(g_{i}+s_{i, j}\right) \cdot \log p_{i}-\log (\gamma /|w|)\right) / \log |\alpha|
$$

hence $n<N_{j}$.

Remark 1. In general, one expects that $p_{i}{ }_{i}, \mathbf{j}$ will not be much larger than $N_{j}$, i.e. not too many consecutive $p_{i}$-adic digits of $\vartheta_{i}$ will be zero. Then $\mathrm{N}_{\mathrm{j}}$ is about as large as $\log \mathrm{N}_{\mathrm{j}-1}$. In practice, the algorithm will often terminate in three or four steps, near to the largest solution. The computation time is polynomial in $s$, the bottleneck of the algorithm is the computation of the $p_{i}$-adic logarithms.

Remark 2. Pethö [1985] gives for $s=1$ a different reduction algorithm. For a prime $p_{i}$ he computes the function $g(u)$, defined for $u \in \mathbb{N}$ as the smallest index $n \geqslant 0$ such that $G_{n} \neq 0$ and $p_{i}^{u} \mid G_{n}$. Note that if the $p_{i}$-adic limit $\lim _{u \rightarrow \infty} g(u)$ exists, then by Lemma 4.10 it is equal to $\vartheta_{i}$. $u \rightarrow \infty$

Remark 3. If $B= \pm 1$ (hence $\Delta>0$ ), we can extend the sequence $\left\{G_{n}\right\}_{n=0}^{\infty}$ to negative indices by the recursion formula

$$
G_{n-1}=A \cdot B \cdot G_{n}-B \cdot G_{n+1} \text { for } n=0,-1,-2, \ldots
$$

(cf. (4.3)). Then (4.5) is true for $n<0$ also. We can solve equation (4.1) with $n \in \mathbb{Z}$ not necessarily nonnegative, by applying Algorithm P twice: once for $\left\{G_{n}\right\}_{n=0}^{\infty}$, and once for the sequence $\left\{G_{n}^{\prime}\right\}_{n=0}^{\infty}$, defined by $G_{n}^{\prime}=G_{-n}$. Note that $G_{n}^{\prime}=B^{n} \cdot\left(\mu \cdot \alpha^{n}+\lambda \cdot \beta^{n}\right)$, and

$$
\vartheta_{i}^{\prime}=-\frac{\log _{p_{i}}(-\mu / \lambda)}{\log _{p_{i}}(\alpha / \beta)}=+\frac{\log _{p_{i}}(-\lambda / \mu)}{\log _{p_{i}}(\alpha / \beta)}=-\vartheta_{i} \quad \text { for } \quad i=1, \ldots, s
$$

Now, instead of applying Algorithm $P$ twice, we can modify it, so that it works for all $n \in \mathbb{Z}$, as follows. Lemmas 4.8 and 4.10 remain correct if we replace $n$ by $|n|$. In Theorem 4.9 the lower bound for $n_{0}$ must be replaced by

$$
\mathrm{n}_{0}>\max (2, \quad|\log | \mu / \lambda| | / \log |\alpha / \beta|, \quad|\log | \lambda / \mu| | / \log |\alpha / \beta|),
$$

and $\gamma$ has to be replaced by

$$
\gamma=\min \left(|\lambda|-|\mu| \cdot|\alpha / \beta|^{-\mathrm{n}_{0}},|\mu|-|\lambda| \cdot|\alpha / \beta|^{-\mathrm{n}_{0}}\right)
$$

Similar modifications should be made in step (i) of Algorithm P. Further, in step (ii), $r_{i}$ should be chosen so large that

$$
\begin{gathered}
\text { if } p_{i} \neq 2 \text { then } p_{i} \geqslant N_{0} \text { and } u_{i, r_{i}} \neq 0, u_{i, r_{i}} \neq p-1 ; \\
\text { else } p_{i} r_{i}^{-1} \geqslant N_{0} \text { and } u_{i, r_{i}} \neq u_{i, r_{i}-1} ;
\end{gathered}
$$

and similar modifications have to be made in step (v) for $\mathrm{s}_{\mathrm{i}, \mathrm{j}}$. With these changes, Theorem 4.12 remains true with $n$ replaced by $|n|$.

We conclude this section with an example.

Example. Let $A=6, B=1, G_{0}=1, G_{1}=4, w=1, p_{1}=2, p_{2}=11$. Then $\alpha=3+2 \cdot \sqrt{ } 2, \quad \beta=3-2 \cdot \sqrt{2}, \lambda=(1+2 \cdot \sqrt{2}) / 4 \cdot \sqrt{ } 2, \mu=(-1+2 \cdot \sqrt{2}) / 4 \cdot \sqrt{2}$, and $\Delta=32$. With $n_{0}=e^{60}=1.142 \times 10^{26}$ we find $C_{4}<2.49 \times 10^{20}$. With the modifications of Remark 3 above we have $\gamma>0.323, C_{5}<1.76$, $C_{6}<2.62 \times 10^{26}, C_{5} \cdot C_{6}<4.62 \times 10^{26}$. Hence all solutions of $G_{n}=2^{m_{1}} \cdot 11^{m_{2}}$ satisfy $|n|<4.62 \times 10^{26}, \max \left(m_{1}, m_{2}\right)<2.62 \times 10^{26}$. We perform the reduction Algorithm $P$ step by step. (We write the $p$-adic number $\sum_{\ell=0}^{\infty} u_{\ell} \cdot p^{\ell}$ as $0 . u_{0} u_{1} u_{2} \ldots$, and if $p>10$ we denote the digits larger than 9 by the symbols A, B, C, ... ).

$$
\begin{align*}
& n_{0}=2, \gamma>0.303, g_{1}=0, g_{1}=1, g>0.0275 \text {, }  \tag{i}\\
& h_{1}=-1, h_{2}=\frac{1}{2}, N_{0}=4.62 \times 10^{26} \text {. } \\
& \vartheta_{1}=0.1011110111010001110010100010011000110010  \tag{ii}\\
& 0000111101010001000001001100111010101101 \\
& 1110001011000011101000011010010101000101 \\
& 1000101011000001100101011111011010001011 \\
& \text { 001.... , } \\
& \vartheta_{2}=0 . A 935905530 \text { 7330A 1A223 } 96230 \text { 3A006 A3366 } 83368 \\
& \text { 8270.... , } \\
& \text { so } r_{1}=90 \text { (since } u_{1,89}=1, u_{1,90}=0,2^{89}>N_{0} \text { ), } \\
& r_{2}=29 \text { (since } u_{2,29} \text { ) }=6,11^{29}>N_{0} \text { ). }  \tag{iii}\\
& \text { (v)-(vi) } s_{1,1}=90, s_{2,1}=29, K_{1}=\text {.true., } N_{1}<76.9 \text {; }
\end{align*}
$$

$$
\begin{aligned}
& \text { (v)-(vi) } s_{1,2}=10, s_{2,2}=2, K_{2}=\text {.true., } N_{2}<8.7 \text {; } \\
& \text { (v)-(vi) } s_{1,3}=6, s_{2,3}=1, K_{3}=\text {.true., } N_{3}<5.8 \text {; } \\
& \text { (v)-(vi) } s_{1,4}=6, s_{2,4}=1, K_{4}=\underline{\text { false., }} N_{4}<5.8 \text {. }
\end{aligned}
$$

Hence $|n| \leqslant 5, m_{1} \leqslant 6, m_{2} \leqslant 2$. We have

| $n$ | -5 | -4 | -3 | -2 | -1 | 0 | 1 | 2 | 3 | 4 | 5 |
| :--- | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| $G_{n}$ | 2174 | 373 | 64 | 11 | 2 | 1 | 4 | 23 | 134 | 781 | 4552 |

So there are 5 solutions: with $n=-3,-2,-1,0,1$.

### 4.8. The reduction algorithm in the elliptic case.

We now present an algorithm to reduce upper bounds for the solutions of (4.1) in the case $\Delta<0$. The idea is to apply alternatingly Algorithms $P$ and one of $H$ and $I$. Let $N$ be an upper bound for $n$, for example $n=C_{7}$ as in Theorem 4.9.

Input: $\alpha, \beta, \lambda, \mu, w, p_{1}, \ldots, p_{s}, N$.
Output: new, reduced upper bounds $N^{*}$ for $n$, and $M_{i}$ for $m_{i}$ for i $=1, \ldots, \mathrm{~s}$.
(i) (initialization) $\mathrm{N}_{\mathrm{O}}:=[\mathrm{N}]$; $\mathrm{j}:=1$;
$g_{i}:=\operatorname{ord}_{p_{i}}(\lambda)+\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}(\alpha / \beta)\right)$
$h_{i}:=\operatorname{ord}_{p_{i}}(\lambda)+\left\{\begin{array}{cl}3 / 2 & \text { if } \quad p_{i}=2 \\ 1 & \text { if } \\ 1 / 2 & p_{i}=3 \\ \text { if } & p_{i} \geqslant 5\end{array}\right\}$ for $i=1, \ldots, s ;$
(ii) (computation of the $\vartheta_{i}{ }^{\prime} s, \varphi, \psi$ ) Compute for $i=1, \ldots$, $s$ the first $\mathrm{r}_{\mathrm{i}} \quad \mathrm{p}_{\mathrm{i}}-\mathrm{adic}$ digits $\mathrm{u}_{\mathrm{i}}, \ell$ of

$$
\vartheta_{i}=-\log _{p_{i}}\left(\frac{-\lambda}{\mu}\right) / \log _{p_{i}}\left(\frac{\alpha}{\beta}\right)=\sum_{\ell=0}^{\infty} u_{i}, \ell \cdot p_{i}^{\ell}
$$

where $r_{i}$ is so large that $\mathrm{p}_{\mathrm{i}} \geqslant \mathrm{N}_{0}$ and $\mathrm{u}_{\mathrm{i}, \mathrm{r}_{\mathrm{i}}} \neq 0$; compute $\psi=\log (-\lambda / \mu) / 2 \pi i$, and the continued fraction

$$
|\varphi|=\left|\frac{1}{2 \pi i} \cdot \log (\alpha / \beta)\right|=\left[0, a_{1}, \ldots, a_{\ell_{0}}, \ldots\right]
$$

with the convergents $\mathrm{p}_{\mathrm{i}} / \mathrm{q}_{\mathrm{i}}$ for $\mathrm{i}=1, \ldots, \ell_{0}$, where $\ell_{0}$ is so large that $\mathrm{q}_{\ell_{0}-1} \leqslant \mathrm{~N}_{0}<\mathrm{q}_{\ell_{0}}$ if $\psi=0 ; \mathrm{q}_{\ell_{0}}>4 \cdot \mathrm{~N}_{0}$ and ${ }^{\mathrm{IIq}} \ell_{0} \mathrm{II}>2 \cdot \mathrm{~N}_{0} / \mathrm{q}_{\ell_{0}}$ if $\psi \neq 0$ and such $\ell_{0}$ can be found in a reasonable amount of time, $q_{\ell_{0}}>4 \cdot N_{0}$ otherwise;
(iii) (one step of Algorithm P) For $\mathrm{i}=1, \ldots$, s put
$M_{i, j}:=\max \left(h_{i}, g_{i}+\min \left\{s \in \mathbb{N}_{0} \mid p_{i}^{s} \geqslant N_{j-1}\right.\right.$ and $\left.\left.u_{i, s} \neq 0\right\}\right) ;$
(iv) (one step of Algorithm $H$ or I)
if $\psi=0$

$$
\text { then } A:=\max \left(a_{1}, \ldots, a_{\ell}{ }_{j}\right) ; \quad v:=|w| \cdot \prod_{i=1}^{s} p_{i}^{M_{i}, j}
$$

$$
\text { choose } n_{0} \geqslant 2 / \log B \text { such that } B^{n_{0} / 2} / n_{0} \geqslant v / 2 \cdot|\mu|
$$

$$
\text { compute the largest integer } N_{j} \text { such that }
$$

$$
B^{N_{j} / 2} / N_{j} \leqslant(A+2) \cdot v / 4 \cdot|\mu|
$$

$$
N_{j}:=\max \left(n_{0}, N_{j}\right) ;
$$

$$
\text { if } N_{j}<N_{j-1} \text { then compute } \ell_{j} \text { with } q_{\ell}-1 \leqslant N_{j}<q_{\ell} \text {; }
$$

$$
j:=j+1 \text {; goto (iii) ; }
$$

else if $\mathrm{Ilq}_{\ell}{ }_{j-1} \cdot \psi \|>2 \cdot N_{j-1} / \mathrm{q}_{\ell}{ }_{j-1}$

$$
\text { then } N_{j}:=\left[2 \cdot \log \left(q_{\ell}^{2}{ }_{j-1}^{2} \cdot v / 4 \cdot|\mu| \cdot N_{j-1}\right) / \log B\right] ;
$$

else compute $K \in \mathbb{Z}$ with $\left|K-q_{\ell}{ }_{j-1} \cdot \psi\right| \leqslant \frac{1}{2}$;

$$
\begin{aligned}
& \text { compute } \mathrm{n}_{0} \in \mathbb{Z}, \quad 0 \leqslant \mathrm{n}_{0}<\mathrm{q}_{\ell}, \text { with } \\
& \mathrm{K}+\mathrm{n}_{0} \cdot \mathrm{p}_{\ell} \quad \equiv 0\left(\bmod \mathrm{q}_{\ell} \text {, }\right) \text {; } \\
& \text { if } \mathrm{n}=\mathrm{n}_{0} \text { is a solution of (4.1) }
\end{aligned}
$$

then print an appropriate message;

$$
N_{j}:=\left[2 \cdot \log \left(q_{\ell}^{j-1}<\cdot v /|\mu|\right) / \log B\right]
$$

if $N_{j}<N_{j-1}$
then compute the minimal $\ell_{j}<\ell_{j-1}$ such that $q_{\ell}>4 \cdot N_{j}$ and $\left\|l q_{\ell} \cdot \psi\right\|>2 \cdot N_{j} / q_{\ell}$ (if such $\ell_{j}$ does not exist, choose the minimal $\ell_{j}$ such that $\left.q_{\ell}>4 \cdot N_{j}\right) ; j:=j+1$; goto (iii) ;
(v) (termination) $N^{*}:=N_{j-1} ; M_{i}:=M_{i, j}$ for $i=1, \ldots, s$; stop.

Figure 7. ALGORITM C. (reduces upper bounds for (4.1) in the case $\Delta<0$ ).

The following theorem now follows at once from the proofs of Lemmas 4.6, 4.7 and Theorem 4. 12.

THEOREM 4.13. Algorithm C terminates. Equation (4.1) with $\Delta<0$ has no solutions with $N^{*}<n<N$ and $\mathrm{m}_{\mathrm{i}}>\mathrm{M}_{\mathrm{i}}$ for $\mathrm{i}=1$, ..., s , apart from those spotted by the algoritm.

We conclude this section with an example.

Example. Let $A=1, B=2, G_{0}=2, G_{1}=3$, then $\Delta=-7, \alpha=(1+\sqrt{ }-7) / 2$ and $\lambda=(2+\sqrt{ }-7) / \sqrt{ }-7$. Let $w= \pm 1, p_{1}=3, p_{2}=7$. We have with $n_{0}=2$ the following results: $\mathrm{C}_{4}<6.40 \times 10^{16}, \mathrm{C}_{3}<9.14 \times 10^{29}, \mathrm{C}_{7}<7.42 \times 10^{30}$, $\max \left(\mathrm{C}_{8,1}, \mathrm{C}_{8,2}\right)<2.30 \times 10^{22}$. Further, $\mathrm{g}_{1}=1, \mathrm{~g}_{2}=0, \mathrm{~h}_{1}=1, \mathrm{~h}_{2}=0$. By Theorem 4.9 we may choose $N_{0}=7.42 \times 10^{30}$. We have

$$
\begin{aligned}
& \varphi=\mid[\pi-\arctan (\sqrt{ } 7 / 3)] / 2 \pi \\
& =[0,2,1,1,2,16,6,1,2,2,13, \\
& 1,1,3,1,1,2,1,2,1,1 \text {, } \\
& 1,1,1,9,2,1,2,1,7,1 \text {, } \\
& 6,269,4,3,1,1,50,2,1,6 \text {, } \\
& 1,1,2,1,1,7,1,61,1,12 \text {, } \\
& 3,7,4,7,3,121,1,21,2,1,7, \ldots] \text {, } \\
& \psi=|[\pi-\arctan (4 \cdot \sqrt{ } / / 3)]| / 2 \pi \\
& =0.2939628336996454026789566605200190806203 . . \text {, } \\
& \vartheta_{1}=0.2001012210000110210200211002220222012021 \\
& 1002020202211020012101000010021110020122 \\
& 11111222022102102212 \text { 2200... , } \\
& \vartheta_{2}=0.3254212042435613402061561134521011633152 \\
& 253364504411254 \text { 55033... . }
\end{aligned}
$$

Now we choose $\ell_{0}=61$, since

$$
q_{61}=142511833114244361193755123881743>4 \cdot N_{0}
$$

and $\quad \mathrm{ll}_{61} \bullet \psi \|=0.24487 \ldots>2 \cdot \mathrm{~N}_{0} / \mathrm{q}_{61}=0.104 \ldots$. We have $\mathrm{M}_{1,1}=67$, $M_{2,1}=37$, and we find $N_{1}=637$. Next we choose $\ell_{1}=9$, since $\mathrm{q}_{9}=10102>4 \times 637$ and $\mathrm{Iq}_{9} \cdot \psi \|=0.38745 \ldots>2 \times 637 / 10102$. We have $M_{1,2}=7, M_{2,2}=4$, and we find $N_{2}=74$. Next we choose $\ell_{2}=6$, since $q_{6}=1291>4 \times 74$, and $\mathrm{llq}_{6} \cdot \psi \|=0.49398>2 \times 74 / 1291$. We have $\mathrm{M}_{1,3}=6$,
$M_{2,3}=3$, and we find $N_{3}=60$. In the next step we find no improvement. Hence $n \leqslant 60, m_{1} \leqslant 6, m_{2} \leqslant 3$. It is a matter of straightforward computation to check that there are only the following 6 solutions of $G_{n}= \pm 3^{m} 1 \cdot 7^{m} 2$ : $G_{1}=3, G_{2}=-1, G_{3}=-7, G_{5}=3^{2}, G_{7}=1, G_{17}=3^{2} \cdot 7^{2}$.

### 4.9. The generalized Ramanujan-Nagell equation.

The most interesting application of the reduction algorithms of the preceding section seems to be the solution of the generalized Ramanujan-Nagell equation (4.2). Let $k$ be a nonzero integer, and let $p_{1}, \ldots, p_{s}$ be distinct prime numbers. Then we ask for all nonnegative integers $x, z_{1}, \ldots, z_{s}$ with

$$
x^{2}+k=\prod_{i=1}^{s} p_{i}^{z_{i}}
$$

First we note that $z_{i}=0$ whenever $-k$ is a quadratic nonresidue $\left(\bmod p_{i}\right)$. Thus we assume that this is not the case for all $i$. Let $p_{i} \| k$ for $i=1, \ldots, t$ and $p_{i} \nmid k$ for $i=t+1, \ldots, s$. Let ord $p_{i}(k)$ be odd for $i=1, \ldots, r$ and even for $i=r+1, \ldots, t$. Dividing by large enough powers of $p_{i}$ for $i=1, \ldots, t$, (4.2) reduces to a finite number of equations

$$
\begin{equation*}
D_{0} \cdot x_{1}^{2}+k_{1}=\prod_{i=r+1}^{s} p_{i}^{z_{i}^{\prime}} \tag{4.13}
\end{equation*}
$$

with $p_{i} \not k_{1}$ for $i=1, \ldots, s$, and $D_{0}$ composed of $p_{1}, \ldots, p_{r}$ only, and squarefree. We distinguish between the $2^{s-r}$ combinations of $z_{i}^{\prime}$ odd or even for $i=r+1, \ldots, s$. Suppose that $z_{i}^{\prime}$ is odd for $i=r+1, \ldots$, u and even for $i=u+1, \ldots, s$. Put

$$
\begin{equation*}
y=\prod_{i=r+1}^{u} p_{i}^{\left(z_{i}^{\prime}-1\right) / 2} \cdot \prod_{i=u+1}^{s} p_{i}^{z_{i}^{\prime} / 2} \tag{4.14}
\end{equation*}
$$

Then, from (4.13),

$$
\begin{equation*}
\mathrm{D}_{0} \cdot \mathrm{x}_{1}^{2}-\left(\prod_{i=r+1}^{u} p_{i}\right) \cdot y^{2}=-k_{1} \tag{4.15}
\end{equation*}
$$

Put $D=D_{0} \cdot \prod_{i=r+1}^{u} p_{i} . \operatorname{Then~(4.14)~and~(4.15)~lead~to~}$

$$
\left\{\begin{array}{c}
v^{2}-D \cdot w^{2}=k_{2}  \tag{4.16}\\
v=\prod_{i=r+1}^{s} p_{i}^{m_{i}}
\end{array}\right.
$$

with $v=y \cdot \prod_{i=r+1}^{u} p_{i}, \quad w=x_{1}, \quad k_{2}=k_{1} \cdot \prod_{i=r+1}^{u} p_{i}$, and also to

$$
\left\{\begin{array}{c}
v^{2}-D \cdot w^{2}=k_{2}  \tag{4.17}\\
w=\prod_{i=r+1}^{s} p_{i}^{m}{ }_{i}
\end{array}\right.
$$

with $v=D_{0} \cdot x_{1}, w=y, k_{2}=-k_{1} \cdot D_{0}$. We proceed with either (4.16) or (4.17), which is the most convenient (e.g. the one with the smaller $\left|k_{2}\right|$ ).

If $D=1$, then (4.16) and (4.17) are trivial. So assume $D>1$. Let $\varepsilon$ be the smallest unit in $\mathbb{Z}+\sqrt[V]{ } \cdot \mathbb{Z}$ with $\varepsilon>1$. It is well known that the solutions $v, w$ of $v^{2}-D \cdot w^{2}=k_{2}$ fall apart into a finite number of classes of associated solutions. Let there be $T$ such classes, and choose for $\tau=1, \ldots, \mathrm{~T}$ in the $\tau$ th class the solution ${ }^{v_{\tau, 0}},{ }^{\mathrm{w}}{ }_{\tau, 0}$ such that $\gamma_{\mathrm{t}}=\mathrm{v}_{\tau, 0}+{ }^{\mathrm{w}} \mathrm{r}_{, 0} \cdot \sqrt{ } \mathrm{D}>1$ is minimal. Then all solutions of $\mathrm{v}^{2}-\mathrm{D} \cdot \mathrm{w}^{2}=\mathrm{k}_{2}$ are given by $v= \pm v_{\tau, n}, w= \pm w_{\tau, n}$, with

$$
\left\{\begin{array}{l}
\mathrm{v}_{\tau, \mathrm{n}}=\left(\gamma_{\tau} \cdot \varepsilon^{\mathrm{n}}+\gamma_{\tau}^{\prime} \cdot \varepsilon^{-\mathrm{n}}\right) / 2  \tag{4.18}\\
\mathrm{w}_{\tau, \mathrm{n}}=\left(\gamma_{\tau} \cdot \varepsilon^{\mathrm{n}}-\gamma_{\tau}^{\prime} \cdot \varepsilon^{-\mathrm{n}}\right) / 2 \cdot v_{\mathrm{D}}
\end{array}\right.
$$

for ${\underset{\sim}{n}}_{\infty}^{\infty} \in \mathbb{Z}$, where $\gamma_{t}^{\prime}=\mathrm{v}_{\tau, 0}-\mathrm{w}_{\tau, 0} . \sqrt{ } \mathrm{D}$. That is, $\left\{\mathrm{v}_{\tau, \mathrm{n}}\right\}_{\mathrm{n}=-\infty}^{\infty}$ and $\left\{W_{\tau, n}\right\}^{\infty}=-\infty$ are linear binary recurrence sequences. Now, (4.16) and (4.17) reduce to $T$ equations of type (4.1). If $\mathrm{k}_{2}=1$, then $T=1, \gamma_{1}=\varepsilon$, $\gamma_{1}^{\prime}=\varepsilon^{-1}$. If $\mathrm{k}_{2} \mid 2 \cdot \mathrm{D}, \mathrm{k}_{2} \neq 1$, then it is easy to prove that $\gamma_{\tau}^{2}=\left|\mathrm{k}_{2}\right| \cdot \varepsilon$, $\gamma_{t}^{\prime 2}=\left|k_{2}\right| \cdot \varepsilon^{-1}$, so that

$$
\begin{aligned}
& \mathrm{v}_{\tau, \mathrm{n}}=\sqrt[V]{ }\left|\mathrm{k}_{2}\right| \cdot\left(\left(\gamma_{\tau} / \sqrt{ }\left|\mathrm{k}_{2}\right|\right)^{2 \mathrm{n}+1}+\left(\gamma_{\tau}^{\prime} / \sqrt{ }\left|\mathrm{k}_{2}\right|\right)^{2 \mathrm{n}+1}\right) / 2, \\
& \mathrm{w}_{\tau, \mathrm{n}}=\sqrt{ }\left|\mathrm{k}_{2}\right| \cdot\left(\left(\gamma_{\tau} / V\left|\mathrm{k}_{2}\right|\right)^{2 \mathrm{n}+1}-\left(\gamma_{\tau}^{\prime} / \sqrt{ }\left|\mathrm{k}_{2}\right|\right)^{2 \mathrm{n}+1}\right) / 2 \cdot \sqrt{ } .
\end{aligned}
$$

In both cases, (4.16) and (4.17) can be solved by elementary means (see Section 4.6, of related interest are Størmer [1897], Mahler [1935], Lehmer [1964], Rumsey and Posner [1964] and Mignotte [1985]). If $\mathrm{k}_{2} \nmid 2 \cdot \mathrm{D}$, then we apply the reduction algorithm to one of the equations $v_{\tau, n}=\prod_{i=r+1}^{s}{ }^{m}{ }_{\mathrm{m}}^{\mathrm{i}}{ }^{\text {i }}$,
$w_{\tau, n}=\prod_{i=r+1}^{s} p_{i}^{m_{i}}$. Note that $n$ is allowed to be negative, since $B= \pm 1$, so we can use the modified algorithm of Remark 3, Section 4.7.

Thus we have a procedure for solving (4.2) completely. It is well known how the unit $\varepsilon$ and the minimal solutions $v_{\tau, 0},{ }^{\mathrm{w}}{ }_{\tau, 0}$ for $\tau=1, \ldots, \mathrm{~T}$ can be computed by the continued fraction algorithm for $\sqrt{ } \mathrm{D}$. We conclude this section with an example. It extends the result of Nagell [1948] (also proved by many others) on the original Ramanujan-Nagell equation $x^{2}+7=2^{z}$.

THEOREM 4.14. The only nonnegative integers $x$ such that $x^{2}+7$ has no prime divisors larger than 20 are the 16 in the following table.

| $x$ | $x^{2}+7$ | $x$ | $x^{2}+7$ | $x$ | $x^{2}+7$ |
| :--- | :--- | :---: | :--- | ---: | :---: |
| 0 | 7 | 7 | $56=2^{3} \cdot 7$ | 31 | $968=2^{3} \cdot 11^{2}$ |
| 1 | $8=2^{3}$ | 9 | $88=2^{3} \cdot 11$ | 35 | $1232=2^{4} \cdot 7 \cdot 11$ |
| 2 | 11 | 11 | $128=2^{7}$ | 53 | $2816=2^{8} \cdot 11$ |
| 3 | $16=2^{4}$ | 13 | $176=2^{4} \cdot 11$ | 75 | $5632=2^{9} \cdot 11$ |
| 5 | $32=2^{5}$ | 21 | $448=2^{6} \cdot 7$ | 181 | $32768=2^{15}$ |
|  |  |  |  |  |  |
|  |  |  |  |  |  |

Proof. Since -7 is a quadratic nonresidue modulo 3, 5, 13, 17 and 19 , we have only the primes 2,7 and 11 left. Only one factor 7 can occur in $x^{2}+7$, thus we have to solve the two equations

$$
\begin{align*}
& x^{2}+7=2^{z_{1}} \cdot 11^{z_{2}}  \tag{4.19}\\
& x^{2}+7=7 \cdot 2^{z_{1}} \cdot 11^{z_{2}} \tag{4.20}
\end{align*}
$$

Equation (4.20) can be solved in an elementary way. We distinguish four cases, each leading to an equation of the type

$$
y^{2}-D \cdot z^{2}=c
$$

with $\mathrm{c} \| 2 \cdot \mathrm{D}$, and either y or z composed of factors 2 and 11 only. We have:
(i) $z_{1}$ even, $z_{2}$ even, $y=2^{z_{1} / 2} \cdot 11^{z_{2} / 2}, \quad z=x / 7, \quad c=1, D=7$;
(ii) $\mathrm{z}_{1}$ odd, $\mathrm{z}_{2}$ even, $\mathrm{y}=2^{\left(\mathrm{z}_{1}+1\right) / 2} \cdot 11^{\mathrm{z}_{2} / 2}, \mathrm{z}=\mathrm{x} / 7, \quad \mathrm{c}=2, \mathrm{D}=14$;
(iii) $z_{1}$ even, $z_{2}$ odd, $y=x, z=2_{1}^{z_{1} / 2} \cdot\left(z_{2}-11\right) / 2\left(z_{2}-1\right) / 2, \quad c=-7, \quad D=77$; (iv) $z_{1}$ odd, $z_{2}$ odd, $y=x, z=2^{\left(z_{1}-1\right) / 2} \cdot 11^{\left(z_{2}-1\right) / 2}, c=-7, D=154$.

In the first example of Section 4.5 we have worked out case (i). We leave the other cases to the reader.

Equation (4.19) can be solved by the reduction algorithm. Again we have four cases, each leading to an equation of the type

$$
y^{2}-D \cdot z^{2}=c
$$

with $z$ composed of factors 2 and 11 only. We have
(i) $z_{1}$ even, $z_{2}$ even, $y=x, z=2^{z_{1} / 2} \cdot 11^{z_{2} / 2}, \quad c=-7, D=1$;
(ii) $z_{1}$ odd, $z_{2}$ even, $y=x, z=2^{\left(z_{1}-1\right) / 2} \cdot 11_{2}^{z_{2} / 2}, \quad c=-7, D=2$;
(iii) $z_{1}$ even, $z_{2}$ odd, $y=x, z=2^{z_{1} / 2} \cdot 11\left(z_{2}-1\right) / 2, \quad c=-7, D=11$;
(iv) $z_{1}$ odd, $z_{2}$ odd, $y=x, z=2^{\left(z_{1}-1\right) / 2} \cdot 11^{\left(z_{2}-1\right) / 2}, c=-7, D=22$.

Case (i) is trivial. The other three cases each lead to one equation of type (4.1). In the example in Section 4.7 we have worked out case (ii). With the following data the reader should be able to perform Algorithm P by hand for the cases (iii) and (iv), thus completing the proof. In these cases $\mathrm{N}<10^{30}$ is a correct upper bound.

Case (iii): $\alpha=10+3 \cdot \sqrt{ } 11, \quad \lambda=(2+\sqrt{ } 11) / 2 \cdot \sqrt{ } 11$,

$$
\begin{aligned}
& \vartheta_{1}=0.1001101000001101010000110101100100111110 \\
& 1101110010000011011010111101000011001101 \\
& 0101010010111011100110000100100101011011 \\
& 0001000111011100010101101011111010111110 \text { 10... , } \\
& \vartheta_{2}=0.23075764253900426090 \text { A92A1 } 037570731458414 \text { 7A238.... . }
\end{aligned}
$$

Case (iv): $\alpha=197+42 \cdot \sqrt{ } 22, \quad \lambda=(9+2 \cdot \sqrt{ } 22) / 2 \cdot \sqrt{ } 22$,

$$
\begin{array}{rl}
\vartheta_{1}= & \begin{array}{llllllllll}
11101 & 01101 & 01110 & 01010 & 10111 & 10001 & 00100 & 00011 \\
10000 & 00110 & 10101 & 01100 & 01101 & 01111 & 01101 & 10101
\end{array} \\
01011 & 10100 \\
01100 & 11101
\end{array} 10011000110001011110
$$

Remarks. 1. The computation time for the above proof was less than 2 sec .
2. Let $\Phi(X, Y)=a \cdot X^{2}+b \cdot X \cdot Y+c \cdot Y^{2}$ be a quadratic form with integral coefficients, and $\Delta=b^{2}-4 \cdot a \cdot c$ positive or negative. Let $k$ be a nonzero integer, and $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{s}}$ distinct prime numbers. Then we note that

$$
4 \cdot a \cdot \Phi(X, Y)=(2 \cdot a \cdot X+b \cdot Y)^{2}-\Delta \cdot Y^{2}
$$

so that the diophantine equations

$$
\Phi(X, k)=\prod_{i=1}^{s} p_{i}^{Z_{i}}, \quad \Phi\left(X, \prod_{i=1}^{S} p_{i}^{z_{i}}\right)=k
$$

in integers $X \neq 0$ and $z_{1}, \ldots, z_{S} \in \mathbb{N}_{0}$, can both be solved by our method.

### 4.10. A mixed quadratic-exponential equation.

In this section we give an application of Algorithm $C$ to the following diophantine equation. Let

$$
\Phi(X, Y)=a \cdot X^{2}+b \cdot X \cdot Y+c \cdot Y^{2}
$$

be a quadratic form with integral coefficients, such that $D=b^{2}-4 \cdot a \cdot c$ is negative. Let $q, v, w$ be nonzero integers, and $p_{1}, \ldots, p_{s}$ distinct prime numbers. Consider the equation

$$
\begin{equation*}
\Phi\left(\mathrm{X}, \mathrm{w} \cdot \prod_{\mathrm{i}=1}^{\mathrm{s}} \mathrm{p}_{\mathrm{i}}^{\mathrm{m}}\right)=\mathrm{v} \cdot \mathrm{q}^{\mathrm{n}} \tag{4.21}
\end{equation*}
$$

in integers $X$, and $n, m_{1}, \ldots, m_{s} \in \mathbb{N}_{0}$.

Let $\beta, \bar{\beta}$ be the roots of $\Phi(x, 1)=0$. Let $h$ be the class number of $\mathbb{Q}(\sqrt{D})$. There exists a $\pi \in \mathbb{Q}(\sqrt{D})$ such that we have the principal ideal equation $(\pi) \cdot(\bar{\pi})=\left(q^{h}\right)$. Put $n=n_{1}+h \cdot n_{2}$, with $0 \leqslant n_{1}<h$. Then $\Phi(\mathrm{X}, \mathrm{Y})=\mathrm{v} \cdot \mathrm{q}^{\mathrm{n}}$ is equivalent to finitely many ideal equations

$$
(\mathrm{a} \cdot \mathrm{X}-\mathrm{a} \cdot \beta \cdot \mathrm{Y}) \cdot(\mathrm{a} \cdot \mathrm{X}-\mathrm{a} \cdot \bar{\beta} \cdot \mathrm{Y})=(\sigma) \cdot(\bar{\sigma}) \cdot(\pi)^{\mathrm{n}_{2}} \cdot(\bar{\pi})^{\mathrm{n}_{2}}
$$

with $(\sigma) \cdot(\bar{\sigma})=\left(a \cdot v \cdot q^{n}\right)$. Hence we have the equations in algebraic numbers

$$
\left\{\begin{array}{l}
a \cdot X-a \cdot \beta \cdot Y=\gamma \cdot \pi^{n_{2}} \\
a \cdot X-a \cdot \bar{\beta} \cdot Y=\bar{\gamma} \cdot \bar{\pi}^{n_{2}}
\end{array}, \quad\left\{\begin{array}{l}
a \cdot X-a \cdot \beta \cdot Y=\gamma \cdot \bar{\pi}_{2}^{n_{2}} \\
a \cdot X-a \cdot \bar{\beta} \cdot Y=\bar{\gamma} \cdot \pi_{2},
\end{array}\right.\right.
$$

where $\gamma$ is composed of $\sigma$, units, and common divisors of $a \cdot X-a \cdot \beta \cdot Y$ and $a \cdot X-a \cdot \bar{\beta} \cdot Y$. Note that there are only finitely many choices for $\gamma$ possible. Thus, (4.21) is equivalent to a finite number of equations

$$
a \cdot(\bar{\beta}-\beta) \cdot w \cdot \prod_{i=1}^{s} p_{i}^{m_{i}}=\gamma \cdot \pi^{n_{2}}-\bar{\gamma} \cdot \bar{\pi}^{n_{2}}
$$

or, if we put $\lambda=\gamma / a \cdot(\bar{\beta}-\beta)$ and $G_{n_{2}}=\lambda \cdot \pi^{n_{2}}+\bar{\lambda} \cdot \bar{\pi}^{n_{2}}$,

$$
\begin{equation*}
G_{n_{2}}=w \cdot \prod_{i=1}^{s} p_{i}^{m_{i}} \tag{4.22}
\end{equation*}
$$

Here, $\quad\left\{G_{n_{2}}\right\}_{n_{2}}^{\infty}=0$ is a recurrence sequence with negative discriminant. So (4.22) is of type (4.1), and can thus be solved by the reduction algorithm of Section 4.8.

Before giving an example we remark that (4.21) with $D>0$ is not solvable with the methods of this chapter. This is due to the fact that in $\mathbb{Q}(\sqrt{D})$ with $D>0$ there are infinitely many units, hence infinitely many possibilities for $\gamma$. Another generalization of equation (4.21) is to replace $q^{n}$ by $\prod_{i=1}^{t} q_{i}^{n}$. This problem is also not solvable by the method of this chapter, since it does not lead to a binary recurrence sequence if $t \geqslant 2$. These problems can however be dealt with by using multi-dimensional approximation methods, as presented in Chapter 3 and applied in Chapter 7.

We finally present an example.

THEOREM 4.15. The equation

$$
x^{2}-3^{m_{1}} \cdot 7^{m_{2}} \cdot x+2 \cdot\left(3^{m_{1}} \cdot 7^{m_{2}}\right)^{2}=11 \cdot 2^{n}
$$

in $X \in \mathbb{Z}, n, m_{1}, m_{2} \in \mathbb{N}_{0}$ has only the following 24 solutions:

| n | $\mathrm{m}_{1}$ | $\mathrm{~m}_{2}$ | X |  | n | $\mathrm{m}_{1}$ | $\mathrm{~m}_{2}$ | X |
| :---: | :---: | :---: | :---: | ---: | ---: | ---: | ---: | ---: |
| 1 | 1 | 0 | -1, | 4 | 5 | 2 | 0 | -10, |
| 1 | 0 | 0 | -4, | 5 | 6 | 19 |  |  |
| 2 | 0 | 0 | -6, | 7 | 7 | 0 | 0 | -37, |
| 3 | 0 | 1 | 2, | 5 | 78 |  |  |  |
| 3 | 1 | 0 | -7, | 10 | 7 | 3 | 0 | 2, |
| 4 | 0 | 1 | -6, | 13 | 11 | 1 | 1 | -137, |
| 4 | 158 |  |  |  |  |  |  |  |
| 4 |  |  | 27 | 2 | -829, | 1270 |  |  |

Proof. Put $\beta=(1+\sqrt{ }-7) / 2$. Then

$$
X^{2}-X \cdot Y+2 \cdot Y^{2}=(X-\beta \cdot Y) \cdot(X-\bar{\beta} \cdot Y)
$$

Note that $\mathbb{Q}(\sqrt[V]{ }-7)$ has class number 1 , and that

$$
2=\frac{1+\sqrt{ }-7}{2} \cdot \frac{1-\sqrt{ }-7}{2}, \quad 11=(2+\sqrt{ }-7) \cdot(2-\sqrt{ }-7)
$$

Suppose $\gamma \mid X-\beta \cdot Y$ and $\gamma \mid X-\bar{\beta} \cdot Y$. Then $\gamma \mid(\bar{\beta}-\beta) \cdot Y=-V-7 \cdot 3^{m_{1}} \cdot 7^{m_{2}}$. On the other hand, $\gamma \mid 11 \cdot 2^{\mathrm{n}}$. It follows that $\gamma= \pm 1$, hence $X-\beta \cdot Y$ and $X-\bar{\beta} \cdot Y$ are coprime. Thus we have two possibilities:

$$
\begin{aligned}
& X-\beta \cdot Y= \pm(2 \pm \sqrt{ }-7) \cdot\left(\frac{1 \pm \sqrt{ }-7}{2}\right)^{n}, \\
& X-\beta \cdot Y= \pm(2 \mp \sqrt{ }-7) \cdot\left(\frac{1 \pm \sqrt{ }-7}{2}\right)^{n},
\end{aligned}
$$

in each equation the 2 nd and $3 \mathrm{rd} \pm$ being independent. Hence we have to solve

$$
G_{n}^{(j)}=\lambda^{(j)} \cdot \beta^{n}+\bar{\lambda}^{(j)} \cdot \bar{\beta}^{n}=3^{m} \cdot 7^{m_{2}} \text { for } j=1,2 \text {, }
$$

with $G_{n+1}^{(j)}=G_{n}^{(j)}-2 \cdot G_{n-1}^{(j)}$ for $j=1,2$, and $\lambda^{(1)}=\bar{\lambda}^{(2)}=(2+\sqrt{ }-7) / \sqrt{ }-7$, so that $G_{0}^{(1)}=G_{0}^{(2)}=1, G_{1}^{(1)}=3, G_{1}^{(2)}=-1$. Note that $\vartheta_{i}^{(1)}=-\vartheta_{i}^{(2)}$ for i $=1,2$, and $\psi^{(1)}=-\psi^{(2)}$. For $j=1$ we have solved (4.22) in the example of Section 4.8. It is left to the reader to solve (4.22) for $j=2$. This can be done with the numerical data given for the case $j=1$.

Remark. The computation time for the above proof was less than 3 sec.

## Chapter 5. The inequality $0<x-y<y^{\delta}$ in S-integers.

The results of this chapter have been published in de Weger [1987].

### 5.1. Introduction.

Let $S$ be the set of all positive integers composed of primes from a fixed finite set $\left\{p_{1}, \ldots, p_{s}\right\}$, where $s \geqslant 2$, and let $\delta \in(0,1)$. In this chapter we study the diophantine inequality

$$
\begin{equation*}
0<x-y<y^{\delta} \tag{5.1}
\end{equation*}
$$

in $x, y \in S$. We give explicit upper bounds for the solutions, and we show how the algorithms for homogeneous, one- and multi-dimensional diophantine approximation in the real case, that were presented in Chapter 3, can be used for finding all solutions of (5.1) for any set of parameters $p_{1}$, ..., $p_{s}$, $\delta$. For $s=2$ the continued fraction method (cf. Section 3.2) is used. For $s \geqslant 3$ we use the $L^{3}$-algorithm for reducing upper bounds (cf. Section 3.7).

Tijdeman [1973] (see also Shorey and Tijdeman [1986], Theorem 1.1) showed that there exists a computable number $c$, depending on max ( $p_{i}$ ) only, such that for all $x, y \in S$ with $x>y \geqslant 3$,

$$
x-y>y /(\log y)^{c}
$$

Thus, for any solution of (5.1) a bound for $x$, $y$ follows. Størmer [1897] showed how to solve the equation $x-y=k$ with $k=1$, 2 with an elementary method (see also Mahler [1935], Lehmer [1964]). Our method can solve this equation for arbitrary $k \in \mathbb{Z}$. For the one-dimensional case $s=2$, Ellison $\left[1971^{b}\right.$ ] has proved the following result: for all but finitely many explicitly given exceptions, $\left|2^{x}-3^{y}\right|>\exp (x \cdot(\log 2-1 / 10)$ ) for all $x, y \in \mathbb{N}$. Cijsouw, Korlaar and Tijdeman (appendix to Stroeker and Tijdeman [1982]) have found all the solutions $x, y \in \mathbb{N}$ of the inequality

$$
\left|p^{x}-q^{y}\right|<p^{\delta \cdot x}
$$

for all primes $p, q$ with $p<q<20$, and with $\delta=\frac{1}{2}$. We shall extend
these results for many more values of $\mathrm{p}, \mathrm{q}$ and with $\delta=0.9$. Further, we determine all the solutions of (5.1) for the multi-dimensional case $s=6$, $\left\{p_{1}, \ldots, p_{6}\right\}=\{2,3,5,7,11,13\}$ with $\delta=\frac{1}{2}$.

In Section 5.2 we derive upper bounds for the solutions of (5.1). In Sections 5.3 and 5.4 we give a method for reducing such upper bounds in the one- and multi-dimensional cases respectively, and work them out explicitly for some examples. Section 5.5 contains tables with numerical data.

### 5.2. Upper bounds for the solutions.

We assume that the primes are ordered as $p_{1}<\ldots<p_{s}$. For a solution $x, y$ of (5.1), the finitely many $z \in \mathbb{N}$ for which $z \cdot x, z \cdot y$ is also a solution of (5.1) can be found without any difficulty. Therefore we may assume that $(x, y)=1$. Put

$$
x=\max _{1 \leqslant i \leqslant s} \operatorname{ord}_{p}(x \cdot y)
$$

Put

$$
\begin{aligned}
& C_{1}=2^{9 \cdot s+26} \cdot s^{s+4} \cdot \max \left(1, \frac{1}{\log p_{1}}\right) \cdot\left(\prod_{i=2}^{s} \log p_{i}\right) \cdot \log \left(e \cdot \log p_{s-1}\right) /(1-\delta) \\
& C_{2}=2 \cdot \log 2 / \log p_{1}+2 \cdot C_{1} \cdot \log \left(e \cdot C_{1} \cdot \log p_{s}\right)
\end{aligned}
$$

THEOREM 5.1. The solutions of (5.1) satisfy $X<C_{2}$.
Proof. If $y \leqslant \frac{1}{2} \cdot x$, then $y^{\delta}>x-y \geqslant y$, which contradicts $y \geqslant 1$. So $y>\frac{1}{2} \cdot x$. Put $\Lambda=\log (x / y)$. Then

$$
\begin{equation*}
0<\Lambda<x / y-1<y^{-(1-\delta)}<\left(\frac{1}{2} \cdot x\right)^{-(1-\delta)} \tag{5.2}
\end{equation*}
$$

By $x=\max (x, y) \geqslant p_{1}^{X}$, we obtain

$$
\begin{equation*}
0<\Lambda<2^{1-\delta} \cdot p_{1}^{-(1-\delta) \cdot X} \tag{5.3}
\end{equation*}
$$

We apply Waldschmidt's result, Lemma 2.4, to $\Lambda$, with $n=s, q=2$. Note that the 'independence condition' $\left[\mathbb{Q}\left(V p_{1}, \ldots, V p_{n}\right): \mathbb{Q}\right]=2^{n}$ holds. Since $p_{i} \geqslant 3$ we have $V_{i}=\log p_{i}$ for $i \geqslant 2$. Thus

$$
\Lambda>\exp \left(-\left(\log X+\log \left(e \cdot \log p_{s}\right)\right) \cdot C_{1} \cdot(1-\delta) \cdot \log p_{1}\right)
$$

Combining this with (5.3) we find

$$
X<C_{1} \cdot \log \left(e \cdot \log p_{S}\right)+\log 2 / \log p_{1}+C_{1} \cdot \log X
$$

The result now follows from Lemma 2.1, since $C_{1}>e^{2}$.

Examples. With $s=2,2 \leqslant p_{i} \leqslant 199, \delta=0.9$ we have $C_{1}<2.30 \times 10^{17}$, $C_{2}<1.97 \times 10^{19}$.
With $s=6,2 \leqslant p_{i} \leqslant 13, \delta=\frac{1}{2}$ we find $C_{1}<8.37 \times 10^{33}, C_{2}<1.35 \times 10^{36}$.

### 5.3. Reducing the upper bounds in the one-dimensional case.

In this section we work out the examples $s=2, \delta=0.9$, and $p_{1}, p_{2}$ run through either the set of primes below 200, or the set of non-powers below 50 (we did not use that the $p_{i}$ are primes). We note that for any other triple $p_{1}, p_{2}, \delta$ the method works similarly. We prove the following result.

THEOREM 5.2. (a) The diophantine inequality

$$
\begin{equation*}
\left|p_{1}^{x_{1}}-p_{2}^{x_{2}}\right|<\min \left(p_{1}^{x_{1}}, p_{2}^{x_{2}}\right)^{\delta} \tag{5.4}
\end{equation*}
$$

with $\mathrm{p}_{1}, \mathrm{p}_{2}$ primes such that $\mathrm{p}_{1}<\mathrm{p}_{2}<200$, and

$$
\begin{gather*}
\mathrm{x}_{1}, \mathrm{x}_{2} \in \mathbb{Z}, \mathrm{x}_{1} \geqslant 2, \mathrm{x}_{2} \geqslant 2 \text {, and either } \delta=\frac{1}{2} \\
\text { or } \delta=0.9, \min \left(\mathrm{p}_{1}, \mathrm{p}_{2}^{2}\right)>10^{15} \tag{5.5}
\end{gather*}
$$

has only the 77 solutions listed in Table I.
(b) The diophantine inequality (5.4) with $\mathrm{p}_{1}, \mathrm{p}_{2}$ non-powers such that $2 \leqslant p_{1}<p_{2} \leqslant 50$ and conditions (5.5), has only the 74 solutions listed in Table II.

Remarks. The Tables are given in Section 5.5. In Tables I, II the column "delta" gives the real number with

$$
\left|p_{1}^{x_{1}}-p_{2}^{x_{2}}\right|=\min \left(p_{1}^{x_{1}}, p_{2}^{x_{2}}\right)^{\text {delta }}
$$

Note that in Theorem 5.2 we do not demand $\left(x_{1}, x_{2}\right)=1$, and in Theorem
5.2(b) we do not demand $p_{1}, p_{2}$ to be primes. The conditions (5.5) are chosen such that the numerous solutions of (5.4) with $\delta=0.9$ and $\min \left(p_{1}^{x_{1}}, p_{2}^{x_{2}}\right) \leqslant 10^{15}$ can be found without much effort.

Proof. Write

$$
\Lambda=\left|x_{1} \cdot \log p_{1}-x_{2} \cdot \log p_{2}\right|, \quad X=\max \left(x_{1}, x_{2}\right)
$$

We assume that

$$
\begin{equation*}
p_{1}^{X}>10^{25} \tag{5.6}
\end{equation*}
$$

since it is easy to find the remaining solutions. Let $\log \mathrm{p}_{1} / \log \mathrm{p}_{2}$ have the simple continued fraction expansion (cf. Section 3.2)

$$
\log p_{1} / \log p_{2}=\left[0, a_{1}, a_{2}, \ldots\right]
$$

and let the convergents be $r_{n} / q_{n}$ for $n=1,2, \ldots$. We may assume that $\left(x_{1}, x_{2}\right)=1$. First we show that $x_{1} \geqslant x_{2}$. For if $x_{1}<x_{2}$, then

$$
\Lambda=x_{2} \cdot \log p_{2}-x_{1} \cdot \log p_{1}>X \cdot\left(\log p_{2}-\log p_{1}\right) \geqslant X \cdot \log \frac{199}{197}
$$

and from (5.3) and (5.6) we then infer

$$
0.0101 \leqslant 0.0101 \cdot X<X \cdot \log \frac{199}{197}<\Lambda<2^{0.1} \cdot 10^{-5 / 2}<0.0034
$$

which is contradictory. Thus $x_{1} \geqslant x_{2}$, hence $X=x_{1}$. Next we prove that

$$
\begin{equation*}
\mathrm{p}_{1}^{\mathrm{X} / 10}>3.1 \cdot \mathrm{X} \tag{5.7}
\end{equation*}
$$

Namely, suppose the contrary. Then $2^{X / 10} \leqslant 3.1 \cdot X$, and it follows that $X \leqslant 80$. This contradicts $3.1 \cdot X \geqslant p_{1}^{X / 10}>10^{5 / 2}$. From (5.3) we infer

$$
\begin{equation*}
\left|\frac{x_{2}}{X}-\frac{\log p_{1}}{\log p_{2}}\right|<\frac{2^{0.1}}{\log p_{2}} \cdot p_{1}^{-X / 10} \cdot \frac{1}{X} \tag{5.8}
\end{equation*}
$$

It follows from (5.7) that

$$
\left|\frac{x_{2}}{X}-\frac{\log p_{1}}{\log p_{2}}\right|<\frac{2^{0.1}}{\log 2} \cdot \frac{1}{3.1 \cdot x^{2}}<\frac{1}{2 \cdot x^{2}}
$$

Hence $x_{2} / X$ is, by Lemma 3.1 , a convergent of $\log p_{1} / \log p_{2}$, say $r_{k} / q_{k}$. From the example at the end of Section 5.2 we see that $X \leqslant X_{0}<1.97 \times 10^{19}$. We find from (3.7) that $k \leqslant 92.996$, hence $k \leqslant 92$. Lemma 3.1 further yields: if (5.3) holds then

$$
\begin{equation*}
\mathrm{a}_{\mathrm{k}+1}>-2+\mathrm{p}_{1}^{\mathrm{q}_{\mathrm{k}} / 10} \cdot \frac{1}{\mathrm{q}_{\mathrm{k}}} \cdot \frac{\log \mathrm{p}_{2}}{2^{0.1}} \tag{5.9}
\end{equation*}
$$

and if

$$
\begin{equation*}
a_{k+1}>p_{1}^{q_{k} / 10} \cdot \frac{1}{q_{k}} \cdot \frac{\log p_{2}}{2^{0.1}} \tag{5.10}
\end{equation*}
$$

then (5.3) holds for $\left(x_{1}, x_{2}\right)=\left(q_{k}, r_{k}\right)$. We computed the continued fraction expansions and the convergents of all numbers $\log p_{1} / \log p_{2}$ in the mentioned ranges for $p_{1}, p_{2}$ exactly up to the index $n$ such that $q_{n-1} \leqslant 1.97 \times 10^{19}<q_{n} \quad$ (cf. Section 2.5 for details of the computational method). Note that $n \leqslant 93$. We checked all convergents for (5.9), and subsequently for (5.10). It is possible, though unlikely, that there is a convergent that satisfies (5.9) but fails (5.10). We met only one such a case: $p_{1}=15, p_{2}=23$, with $\log 15 / \log 23=[0,1,6,2,1,51, \ldots]$, so that $a_{5}=51, r_{4}=19, q_{4}=22$. Now, (5.9) holds but (5.10) fails, since

$$
15^{2.2} \cdot \frac{1}{22} \cdot(\log 19) / 2^{0.1}=51.4 \ldots \in[51,53)
$$

We have in this case $\Lambda=0.002714 \ldots<0.002771 \ldots=2^{0.1} \cdot 15^{-2.2}$, so (5.3) is true. But $\log \left(15^{22}-23^{19}\right) / \log \left(23^{19}\right)=0.9008 \ldots>\delta$, so (5.1) is not true. This example illustrates that (5.3) is weaker than (5.1). Therefore all found solutions of (5.3) have been checked for (5.1) as well. The proof is now completed by the details of the computations, which we omit here.

Remarks. 1. Theorem 5.2(a) is used in the proof of Theorem 6.2.
2. The computations for the proof of Theorem 5.2 took 35 sec .

### 5.4. Reducing the upper bounds in the multi-dimensional case.

Now let $s \geqslant 3$. Put $x_{i}=\operatorname{ord}_{p_{i}}(x / y)$ for $i=1, \ldots, s$. Then $X=\max \left|x_{i}\right|$, and

$$
\Lambda=\sum_{i=1}^{S} x_{i} \cdot \log p_{i}
$$

Note that (5.3) is of the form (3.1). Hence by Theorem 5.1 we can use the method described in Section 3.7 for solving (5.3). We shall do so for the example $s=6,\left\{p_{1}, \ldots, p_{6}\right\}=\{2,3,5,7,11,13\}$ (the first six primes), and $\delta=\frac{1}{2}$.

We use small refinements of Lemmas 3.7 and 3.8 , devised specially for this application, as follows. Let notation be as in Section 3.7.

LEMMA 5.3. Let $X_{1}$ be a positive number such that

$$
\begin{equation*}
\ell(\Gamma) \geqslant V\left(4 \cdot n^{2}+(n-1) \cdot \gamma^{2}\right) \cdot x_{1} \tag{5.11}
\end{equation*}
$$

Then (5.3) has no solutions with for $i=1, \ldots$, $s$

$$
\begin{equation*}
\log \left(\gamma \cdot C \cdot \sqrt{2} / s \cdot X_{1}\right) / \frac{1}{2} \cdot \log p_{i} \leqslant\left|x_{i}\right| \leqslant X \leqslant X_{1} \tag{5.12}
\end{equation*}
$$

LEMMA 5.4. Suppose that

$$
\begin{equation*}
|\tilde{\Lambda}|>\sum_{i=1}^{S}\left|x_{i}\right| \tag{5.13}
\end{equation*}
$$

Then

$$
\begin{equation*}
\left|x_{i}\right|<\log \left(\gamma \cdot C \cdot \sqrt{ } 2 /\left(|\lambda|-\sum_{i=1}^{S}\left|x_{i}\right|\right)\right) /(1-\delta) \cdot \log p_{i} \tag{5.14}
\end{equation*}
$$

Remark. Lemmas 5.3 and 5.4 are refinements of Lemma 3.8, in that they differentiate between the different $x_{i}$. Moreover, Lemma 5.3 has slightly sharper condition and conclusion than Lemma 3.7.

Proofs (of Lemmas 5.3 and 5.4). Analogous to the proofs of Lemmas 3.7 and 3.8, using (5.2) and

$$
p_{i}^{\left|x_{i}\right|} \leqslant \max (x, y)=x<2 \cdot|\Lambda|^{-1 / 2}
$$

THEOREM 5.5. The diophantine inequality

$$
\begin{gathered}
0<x-y<\sqrt{y} \\
\text { in } x, y \in S=\left\{2^{x_{1}} \cdot \ldots \cdot 13^{x_{6}} \mid x_{i} \in \mathbb{N}_{0} \text { for } i=1, \ldots, 6\right\} \text { with } \\
(x, y)=1 \text { has exactly } 605 \text { solutions. Among those, } 571 \text { satisfy } \\
\operatorname{ord}_{2}(x \cdot y) \leqslant 19, \operatorname{ord}_{3}(x \cdot y) \leqslant 12, \operatorname{ord}_{5}(x \cdot y) \leqslant 8, \\
\operatorname{ord}_{7}(x \cdot y) \leqslant 7, \operatorname{ord}_{11}(x \cdot y) \leqslant 5, \operatorname{ord}_{13}(x \cdot y) \leqslant 5 .
\end{gathered}
$$

The remaining 34 solutions are listed in Table III.

Remark. The upper bounds for $\operatorname{ord}_{p_{i}}(x \cdot y)$ given for the 571 solutions not listed in Table III are chosen such that it takes a reasonable amount of computer time to find them all by a brute force method. The list of all 605 solutions is too extensive to be reproduced here.

Proof. By the example at the end of Section 5.2 we know that $X<X_{0}$ for $X_{0}=1.35 \times 10^{36}$. We apply the method described in Section 3.7. Take $C=10^{240}$ (which is chosen so that it is somewhat larger than $X_{0}^{6}$ ), and $\gamma=1$. We applied the $L^{3}$-algorithm to the corresponding lattice $\Gamma_{1}$, and found a reduced basis $\underline{c}_{1}, \ldots, \underline{c}_{6}$ with $\left|\underline{c}_{1}\right|>9.40 \times 10^{39}$. By Lemma 3.4,

$$
\ell\left(\Gamma_{1}\right)>2^{-5 / 2} \cdot 9.40 \times 10^{39}>1.66 \times 10^{39}
$$

This is larger than $V\left(4 \cdot 6^{2}+5 \cdot 1^{2}\right) \cdot X_{0}=1.64 \ldots \times 10^{37}$, so (5.11) holds with $X_{1}=X_{0}$. By Lemma 5.3 we find

$$
X<\log \left(10^{240} \cdot \sqrt{ } 2 / 6 \cdot 1.35 \times 10^{36}\right) / \frac{1}{2} \cdot \log 2<1350.4
$$

so $X \leqslant 1350$. Next we choose $C=10^{32}, \gamma=1$, and $X_{0}=1350$. The reduced basis of the corresponding lattice $\Gamma_{2}$ was computed, and we found $\left|\underline{c}_{1}\right|>2.71 \times 10^{5}$. Hence $\ell\left(\Gamma_{2}\right)>4.79 \times 10^{4}$, which is larger than $\sqrt{ } 149 \cdot 1350=1.64 \ldots \times 10^{4}$. Hence Lemma 5.3 yields for all $i=1, \ldots, 6$

$$
\left|x_{i}\right|<\log \left(10^{32} \cdot \sqrt{ } 2 / 6 \cdot 1350\right) / \frac{1}{2} \cdot \log p_{i}
$$

and it follows that

$$
\begin{align*}
& \left|x_{1}\right| \leqslant 187, \quad\left|x_{2}\right| \leqslant 118, \quad\left|x_{3}\right| \leqslant 80,  \tag{5.15}\\
& \left|x_{4}\right| \leqslant 66, \quad\left|x_{5}\right| \leqslant 54, \quad\left|x_{6}\right| \leqslant 50 .
\end{align*}
$$

Next we choose $C=10^{12}, \gamma=10^{4}$. We use Lemma 5.4 as follows. If $|\lambda|>10^{6}$ then (5.13) holds by (5.15), and Lemma 5.4 yields

$$
\begin{align*}
& \left|x_{1}\right| \leqslant 67, \quad\left|x_{2}\right| \leqslant 42, \quad\left|x_{3}\right| \leqslant 29  \tag{5.16}\\
& \left|x_{4}\right| \leqslant 24, \quad\left|x_{5}\right| \leqslant 19, \quad\left|x_{6}\right| \leqslant 18
\end{align*}
$$

All vectors in the corresponding lattice $\Gamma_{3}$ satisfying (5.15) and $|\lambda|<10^{6}$ have been computed with the Fincke and Pohst algorithm, cf. Section 3.6. We omit details. We found that there exist only two such vectors, but they do not correspond to solutions of (5.1). Hence all solutions of (5.1) satisfy (5.16). Next, we choose $C=10^{8}, \gamma=10^{4}$. If
$|\lambda|>5 \times 10^{5}$ then Lemma 5.4 yields

$$
\begin{align*}
& \left|x_{1}\right| \leqslant 42, \quad\left|x_{2}\right| \leqslant 27, \quad\left|x_{3}\right| \leqslant 18  \tag{5.17}\\
& \left|x_{4}\right| \leqslant 15, \quad\left|x_{5}\right| \leqslant 12, \quad\left|x_{6}\right| \leqslant 11 .
\end{align*}
$$

There are 143 vectors in the corresponding lattice $\Gamma_{4}$ satisfying (5.16) and $|\lambda| \leqslant 5 \times 10^{5}$. Of them, 2 correspond to solutions of (4.1), namely those with

$$
\begin{aligned}
& \left(x_{1}, \ldots, x_{6}\right)=(7,-5, \quad 3,-9,-3, \quad 8), \quad \lambda=257674 \\
& \left(x_{1}, \ldots, x_{6}\right)=(-10,10,-6, \quad 5,-6, \quad 4), \quad \lambda=144817
\end{aligned}
$$

Both also satisfy (5.17). Hence all solutions of (5.1) satisfy (5.17). At this point it seems inefficient to choose appropriate parameters $C, \gamma$, and a bound for $|\lambda|$ to repeat the procedure with. But the bounds of (5.17) are small enough to admit enumeration. Doing so, we found the result.

Remark. Theorems 5.2 and 5.5 find applications in solving other exponential diophantine equations, see Stroeker and Tijdeman [1982], Alex [1985 ${ }^{\text {a }}$ ], [1985 ${ }^{\text {b }}$, Tijdeman and Wang [1988], and Section 6.4 of this book.

Remark. The computation of the reduced basis of $\Gamma_{1}$ took 113 sec , where we applied the $L^{3}$-algorithm as we described it in Section 3.5 , in 12 steps. The direct search for the solutions of (5.17) took 228 sec . The remaining computations (computation of the $\log \mathrm{p}_{\mathrm{i}}$ up to 250 decimal digits, of the reduced basis of $\Gamma_{2}$, and of the short vectors in $\Gamma_{3}$ and $\Gamma_{4}$ ) took 8 sec. Hence in total we used 349 sec.

### 5.5. Tables.

Table I. (Theorem 5.2(a)).







| $p_{1}$ | $x_{1}$ |  |  | $p_{1}^{x_{1}}$ | $p_{2}$ | $x_{2}$ |  |  | $p_{2}^{x_{2}}$ | delta |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 3 |  |  | 8 | 3 | 2 |  |  | 9 | . 00000 |
| 3 | 3 |  |  | 27 | 5 | 2 |  |  | 25 | 0.21534 |
| 2 | 5 |  |  | 32 | 3 | 3 |  |  | 27 | 0.48832 |
|  | 5 |  |  | 32 | 6 | 2 |  |  | 36 | 0.40000 |
| 5 | 3 |  |  | 125 | 11 | 2 |  |  | 121 | 0.28906 |
| 2 | 7 |  |  | 128 | 11 | 2 |  |  | 121 | 0.40575 |
| 2 | 7 |  |  | 128 | 5 | 3 |  |  | 125 | 0.22754 |
| 6 | 3 |  |  | 216 | 15 | 2 |  |  | 225 | 0.40876 |
| 2 | 8 |  |  | 256 | 3 | 5 |  |  | 243 | 0.46694 |
| 7 | 3 |  |  | 343 | 19 | 2 |  |  | 361 | 0.49512 |
| 2 | 9 |  |  | 512 | 23 | 2 |  |  | 529 | 0.45416 |
| 2 | 10 |  |  | 1024 | 10 | 3 |  |  | 1000 | 0.46007 |
| 6 | 4 |  |  | 1296 | 11 | 3 |  |  | 1331 | 0.49607 |
| 12 | 3 |  |  | 1728 | 42 | 2 |  |  | 1764 | 0.48070 |
| 2 | 11 |  |  | 2048 | 45 | 2 |  |  | 2025 | 0.41184 |
| 3 | 7 |  |  | 2187 | 13 | 3 |  |  | 2197 | 0.29941 |
| 3 | 7 |  |  | 2187 | 47 | 2 |  |  | 2209 | 0.40194 |
| 13 | 3 |  |  | 2197 | 47 | 2 |  |  | 2209 | 0.32293 |
| 15 | 4 |  |  | 50625 | 37 | 3 |  |  | 50653 | 0.30762 |
| 6 | 7 |  |  | 79936 | 23 |  |  |  | 79841 | 0.36309 |
|  | 50 | 112589 | 99068 | 42624 | 47 | 9 | 111913 | 04731 | 02767 | 0.85259 |
| 2 | 50 |  |  | 42624 | 18 | - 12 | 115683 | 13814 | 26176 | 0.89628 |
| 24 | ${ }_{11}^{11}$ | 152168 | 11431 | ${ }^{69024}$ | 33 | 10 | 153157 | 89852 | 64449 | 0.85597 |
| 15 | ${ }_{5}^{13}$ | 194619 | 50683 | 59375 | 50 | 9 | 195312 | 50000 | 00000 | 0.83986 |
| 2 | 51 | 225179 | 98136 | 85248 | 19 | 12 | 221331 | 49190 | 66161 | 0.88532 |
| ${ }^{6}$ | 20 | 365615 | 84400 | ${ }_{152976}$ | 26 | 11 | 367034 | 44869 | 87776 | 0.84507 |
| 11 | 15 | 417724 | 81694 | 15651 | 20 | 12 | 409600 | 00000 | 00000 | 0.89095 |
| 28 | 11 | 829350 | 94674 | 71872 | 39 | 10 | 814040 | 60851 | 91601 | 0.89154 |
| 10 | 16 | 1000000 | 00000 | 00000 | 17 | 13 | 990457 | 80329 | 05937 | 0.87396 |
| 5 | 23 | 1192092 | 89550 | 78125 | 29 | 11 | 1220050 | 97657 | 05829 | 0.89862 |
| 2 | 54 | 1801439 | 85094 | 81984 | 30 | 11 |  | 00000 | 00000 | 0.89096 |
| 23 | 12 | 2191462 | 44320 | 20321 | 43 | 10 | 2161148 | 23132 | 84249 | 0.88656 |
| 5 | 21 | 2193695 | 06403 | 77856 | 23 | 12 | 2191462 | 44320 | 20321 | 0.81690 |
| ${ }^{6}$ | ${ }_{5}^{21}$ | 2193695 | 06403 | 77856 | 43 | 10 | 2161148 | 23132 | 84249 | 0.88845 |
| 2 | 55 | 3602879 | 70189 | 63968 | 24 | 12 | 3652034 | 74360 | 56576 | 0.88735 |






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| $x_{1}$ | $x_{2}$ | $x_{3}$ | $x_{4}$ | $x_{5}$ | $x_{6}$. |  |  | $x$ |  |  | $y$ |  | $x-y$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| -1 | -11 | -1 | . 0 | 6 | 0 |  |  | 71561 |  |  | 71470 |  | 91 |
| 0 | 4 | 5 | 1 | -6 | 0 |  |  | 71875 |  |  | 71561 |  | 314 |
| 21 | -2 | -2 | -1 | -3 | 0 |  |  | 97152 |  | 20 | 96325 |  | 827 |
| 1 | 13 | -1 | -3 | -1 | -2 |  |  | 88646 |  | 31 | 88185 |  | 461 |
| 19 | 0 | 0 | -8 | 1 | 0 |  |  | 67168 |  | 57 | 64801 |  | 2367 |
| 6 | 2 | -1 | 1 | -6 | 3 |  |  | 58304 |  | 88 | 57805 |  | 499 |
| -2 | 15 | -1 | -2 | -4 | 0 |  | 143 | 48907 |  | 143 | 48180 |  | 727 |
| 11 | -15 | 0 | 2 | 1 | 1 |  | 143 | 50336 |  | 143 | 48907 |  | 1429 |
| 1 | 8 | -1 | -8 | 0 | 3 |  | 288 | 29034 |  | 288 | 24005 |  | 5029 |
| -22 | 5 | 1 | -1 | 1 | 3 |  | 293 | 62905 |  | 293 | 60128 |  | 2777 |
| 13 | 1 | 3 | -1 | 1 | -6 |  |  | 92000 |  | 337 | 87663 |  | 4337 |
| 1 | 2 | 9 | -4 | -4 | 0 |  | 351 | 56250 |  | 351 | 53041 |  | 3209 |
| 3 | 3 | 0 | 4 | 2 | -7 |  | 627 | 52536 |  | 627 | 48517 |  | 4019 |
| -26 | 1 | 0 | 5 | 3 | 0 |  | 671 | 10351 |  | 671 | 08864 |  | 1487 |
| 3 | -13 | 10 | -2 | 0 | 0 |  | 781 | 25000 |  | 781 | 21827 |  | 3173 |
| 8 | -2 | $-10$ | 4 | 1 | 1 |  | 878 | 95808 |  | 878 | 90625 |  | 5183 |
| 25 | 1 | -4 | 0 | -5 | 0 |  | 1006 | 63296 |  | 1006 | 56875 |  | 6421 |
| -6 | 1 | -2 | -6 | 0 | 7 |  | 1882 | 45551 |  | 1882 | 38400 |  | 7151 |
| 8 | -13 | 0 | 3 | -2 | 3 |  | 1929 | 14176 |  | 1929 | 13083 |  | 1093 |
| 1 | -13 | -3 | 7 | 2 | 0 |  | 1992 | 97406 |  | 1992 | 90375 |  | 7031 |
| -4 | -1 | -4 | 1 | -4 | 7 |  | 4392 | 39619 |  | 4392 | 30000 |  | 9619 |
| -4 | 2 | -11 | 2 | 6 | 0 |  | 7812 | 58401 |  | 7812 | 50000 |  | 8401 |
| 16 | -3 | 5 | 1 | -1 | -6 |  | 14336 | 00000 |  | 14335 | 62273 |  | 37727 |
| -8 | 8 | 0 | -8 | 3 | 2 |  | 14758 | 24779 |  | 14757 | 89056 |  | 35723 |
| -5 | -2 | -5 | 11 | 0 | -3 |  | 19773 | 26743 |  | 19773 | 00000 |  | 26743 |
| -25 | 7 | 1 | 0 | -2 | 5 |  | 40600 | 88955 |  | 40600 | 86272 |  | 2683 |
| 2 | 0 | 13 | -9 | -2 | 0 |  | 48828 | 12500 |  | 48827 | 86447 |  | 26053 |
| -14 | 19 | -2 | -4 | 1 | -1 |  | 27848 | 76137 |  | 27848 | 44800 |  | 31337 |
| -24 | -1 | -2 | 12 | -1 | 0 |  | 38412 | 87201 |  | 38412 | 03200 |  | 84001 |
| -5 | 5 | 10 | 0 | 1 | -8 |  | 61035 | 15625 |  | 61033 | 83072 | 1 | 32553 |
| 2 | -4 | -9 | 3 | 7 | -2 |  | 67363 | 98612 |  | 67363 | 28125 |  | 70487 |
| 18 | 7 | 0 | -13 | 0 | 2 | 9 | 68892 | 08832 | 9 | 68890 | 10407 |  | 98425 |
| 7 | -5 | 3 | -9 | -3 | 8 | 1305 | 16915 | 36000 | 1305 | 16881 | 72831 | 33 | 63169 |
| -10 | 10 | -6 | 5 | -6 | 4 | 2834 | 49801 | 04623 | 2834 | 49760 | 00000 | 41 | 04623 |

## Chapter 6. The equation $x+y=z$ in S-integers.

The results of this chapter have been published in de Weger [1987].

### 6.1. Introduction.

Let $S$ be the set of all positive integers composed of primes from a fixed finite set $\left\{p_{1}, \ldots, p_{s}\right\}$, where $s \geqslant 3$. This chapter is devoted to the diophantine equation

$$
\begin{equation*}
x+y=z \tag{6.1}
\end{equation*}
$$

in $x, y, z \in S$. Without loss of generality we may assume that $x, y, z$ are relatively prime. For any $a \in S$ we define

$$
m(a)=\max _{1 \leqslant i \leqslant s} \text { ord }_{p_{i}}(a)
$$

It was proved by Mahler [1933] that (6.1) has only finitely many solutions, but his proof is ineffective. An effective version, i.e. an effectively computable upper bound for $m(x \cdot y \cdot z)$ for the solutions $x, y, z$ of (6.1), can be derived from the results of Coates [1969], [1970] and Sprindzuk [1969], since (6.1) can be reduced to a finite number of Thue equations. See also Chapter 1 of Shorey and Tijdeman [1986].

We derive an explicit upper bound in Section 6.2. Section 6.3 is devoted to some details of the p-adic approximation lattices on which the reduction method of Sections 6.4 and 6.5 are based. In Section 6.4 we give a method of solving (6.1) in the one-dimensional case $s=3$. This method is based on the reduction procedure given in Section 3.10, and we also use a combination of p-adic and real approximation techniques, similar to that of Section 4.8. But instead of actually performing the real reduction step, we now can simply refer to the results of Chapter 5. As an example we find all the solutions of the slightly more general equation $x \pm y=w \cdot z$, where $x, y, z$ are powers of 2,3 or 5 , and $w \in \mathbb{Z},|w| \leqslant 1000000,(w, z)=1$.

In Section 6.5 we give a procedure for solving (6.1) in the multi-dimensional case $s \geqslant 4$, based on the reduction procedure described in Section 3.11. We work out the example $\left\{p_{1}, \ldots, p_{6}\right\}=\{2,3,5,7,11,13\}$, and actually determine all the solutions. This generalizes the result of Alex [1976], who gave by elementary arguments a complete solution of (6.1) for the case $\left\{\mathrm{p}_{1}, \ldots, \mathrm{p}_{4}\right\}=\{2,3,5,7$ \} . See also Rumsey and Posner [1964] and Brenner and Foster [1982]. We conclude in Section 6.6 with some remarks on the Oesterlé-Masser conjecture, also known as the 'abc'-conjecture, which is related to equation (6.1). In particular, our method of solving (6.1) leads to a method of finding examples that are of interest with respect to the abc-conjecture. Finally, we give tables in Section 6.7.

### 6.2. Upper bounds.

We give in this section an upper bound for the solutions of (6.1), based on Lemma 2.6 (cf. Yu [1987]). Note that in de Weger [1987] we used the result of van der Poorten [1977] instead (see also the Correction to de Weger [1987]).

We introduce a lot of notation. Assume that $p_{1}<\ldots<p_{s}$. Let $q_{i}$ be the smallest prime with $q_{i} \nmid p_{i} \cdot\left(p_{i}-1\right)$ for $i=1$, ..., s. Put

$$
\begin{aligned}
& t=[2 \cdot s / 3], \quad P=\prod_{i=1}^{s} p_{i}, \quad q=\max _{i} q_{i}, \\
& C_{1}(2, t) \text { and } a_{1} \text { as in lemma } 2.6 \text { with } n=t \text {, } \\
& U=C_{1}(2, t) \cdot a_{1}^{t} \cdot t^{t+5 / 2} \cdot q^{2 \cdot t} \cdot(q-1) \cdot \log ^{2}(t \cdot q) \cdot \max _{i} \frac{\left(p_{i}-1\right) \cdot\left(2+\frac{1}{p_{i}-1}\right)^{t}}{\left(\log p_{i}\right)^{t+2}} . \\
& \cdot\left(\log p_{S}\right)^{t} \cdot\left(\log \left(4 \cdot \log p_{S}\right)+\frac{\log p_{S}}{8 \cdot t}\right), \\
& \mathrm{C}_{1}=\mathrm{U} / 6 \cdot \mathrm{t}, \quad \mathrm{C}_{2}=\mathrm{U} \cdot \log 4, \\
& V_{i}=\max \left(1, \log p_{i}\right) \quad \text { for } i=s-t+1, \ldots, s, \quad \Omega=\prod_{i=s-t+1}^{s} V_{i}, \\
& C_{3}=2^{9 \cdot t+26} \cdot t^{t+4} \cdot \Omega \cdot \log \left(e \cdot V_{S-1}\right), \\
& C_{4}=\max \left(7.4,\left(C_{1} \cdot \log \left(P / p_{1}\right)+C_{3}\right) / \log p_{1}\right), \\
& C_{5}=\left(C_{2} \cdot \log \left(P / p_{1}\right)+C_{3} \cdot \log \left(e \cdot V_{S}\right)+0.327\right) / \log p_{1},
\end{aligned}
$$

$$
\begin{aligned}
& C_{6}=\max \left(C_{5},\left(C_{2} \cdot \log \left(P / p_{1}\right)+\log 2\right) / \log p_{1}\right), \\
& C_{7}=2 \cdot\left(C_{6}+C_{4} \cdot \log C_{4}\right) \\
& C_{8}=\max \left(p_{s}, \log \left(2 \cdot\left(P / p_{1}\right)^{p_{S}}\right) / \log p_{1}, C_{2}+C_{1} \cdot \log C_{7}, C_{7}\right) .
\end{aligned}
$$

Now we state the main result.

THEOREM 6.1. The solutions of (6.1) satisfy $m(x \cdot y \cdot z) \leqslant C_{8}$.

Proof. If we consider instead of (6.1) the equivalent equation

$$
\begin{equation*}
x \pm y=z \tag{6.2}
\end{equation*}
$$

then we may assume that $x \cdot y$ has at most $t$ prime divisors, $p_{i_{1}}, \ldots, p_{i_{t}}$ say. Suppose first that $m(x \cdot y) \leqslant p_{s}$. Then

$$
p_{1}^{m(z)} \leqslant z \leqslant 2 \cdot \max (x, y)<2 \cdot\left(P / p_{1}\right)^{p_{s}},
$$

hence

$$
m(x \cdot y \cdot z)<\max \left(p_{s}, \log \left(2 \cdot\left(P / p_{1}\right)^{p_{s}}\right) / \log p_{1}\right) \leqslant C_{8} .
$$

Next suppose that $m(x \cdot y) \geqslant p_{S}$ and $m(z) \geqslant 2$. Then for some $p=p_{i}$,

$$
m(z)=\operatorname{ord}_{p}(z)=\operatorname{ord}_{p}\left|\left[ \pm \frac{x}{y}-1\right]\right|=\operatorname{ord}_{p}\left|\left[\log _{p}\left(\frac{x}{y}\right)\right]\right| .
$$

Put $x / y=\prod_{j=1}^{t} p_{i}{ }_{i}{ }_{j}$. Then $m(x \cdot y)=\max _{1 \leqslant j \leqslant t}\left|x_{i}\right|$. We apply Lemma 2.6 ( Yu ' s lemma) with $n=t, B_{0}=B_{n}=B^{\prime}=B=m(x \cdot y)$. Since $m(x \cdot y) \geqslant p_{s}$ and $t \geqslant 2$ we have

$$
\left.W=\max \|\left[\log \left(1+\frac{3}{4 \cdot t} \cdot B\right), \log B, \log p\right] \right\rvert\,=\log B .
$$

Note that $C_{1}(p, n)$ is maximal for $p=2$. We obtain

$$
\begin{equation*}
m(z)<C_{1} \cdot \log m(x \cdot y)+C_{2} . \tag{6.3}
\end{equation*}
$$

Obviously (6.3) is true if $m(z)<2$. If in (6.2) the plus sign holds, then

$$
\left(P / p_{1}\right)^{m(z)} \geqslant z>\max (x, y) \geqslant p_{1}^{m(x \cdot y)} .
$$

By (6.3) and $C_{3}>0$ it then follows that

$$
\begin{equation*}
m(x \cdot y)<C_{4} \cdot \log m(x \cdot y)+C_{6} . \tag{6.4}
\end{equation*}
$$

Next suppose that in (6.2) the minus sign holds. Then we apply Lemma 2.4 to prove (6.4) for this case, as follows. Suppose (6.4) is false. Then

$$
\left|\frac{y}{x}-1\right|=\frac{z}{x}=\frac{z}{\max (x, y)} \leqslant \frac{\left(P / p_{1}\right)^{m(z)}}{p_{1}^{m(x \cdot y)}}<\frac{\left(P / p_{1}\right)^{C_{1} \cdot \log m(x \cdot y)+C_{2}}}{p_{1}{ }^{C_{4}} \cdot \log m(x \cdot y)+C_{6}}
$$

which is less than $\frac{1}{2}$, by the definition of $C_{4}$ and $C_{6}$. Hence

$$
\left|\log \frac{y}{x}\right|<(2 \cdot \log 2) \cdot\left|\frac{y}{x}-1\right|<(2 \cdot \log 2) \cdot \frac{\left(P / p_{1}\right)^{C_{1} \cdot \log m(x \cdot y)+C_{2}}}{p_{1}^{m(x \cdot y)}}
$$

On the other hand, Lemma 2.4 yields

$$
\left|\log \frac{y}{x}\right|>\exp \left(-C_{3} \cdot\left(\log m(x \cdot y)+\log \left(e \cdot V_{s}\right)\right)\right)
$$

Thus we obtain

$$
\begin{aligned}
& m(x \cdot y) \cdot \log p_{1}<\log (2 \cdot \log 2)+\left(C_{1} \cdot \log m(x \cdot y)+C_{2}\right) \cdot \log \left(P / p_{1}\right) \\
& \quad+C_{3} \cdot\left(\log m(x \cdot y)+\log \left(e \cdot V_{S}\right)\right) \leqslant\left(\log p_{1}\right) \cdot\left(C_{4} \cdot \log m(x \cdot y)+C_{6}\right)
\end{aligned}
$$

This contradicts our assumption that (6.4) if false. Consequently (6.4) is true in all cases. Now, by $C_{4}>e^{2}$, Lemma 2.1 yields $m(x \cdot y)<C_{7}$, and (6.3) then yields $m(x \cdot y \cdot z)<C_{8}$.

Examples. If $s=3,\left\{p_{1}, p_{2}, p_{3}\right\}=\{2,3,5\}$ then $C_{8}<3.98 \times 10^{17}$. If $s=6,\left\{p_{1}, \ldots, p_{6}\right\}=\{2,3,5,7,11,13\}$ then $C_{8}<5.60 \times 10^{27}$.

### 6.3. The p-adic approximation lattices.

As in the proof of Theorem 6.1 we consider (6.2) instead of (6.1). Let $p$ be any of the primes $p_{1}, \ldots, p_{s}$. We may assume that $p \not x x \cdot y$. Rename the other primes as $p_{0}, \ldots, p_{s-2}$, such that ord $\left.\log _{p}\left(p_{0}\right)\right)$ is minimal. For i $=1, \ldots, s-2$ put (cf. Section 3.11)

$$
\vartheta_{i}=-\log _{p}\left(p_{i}\right) / \log _{p}\left(p_{0}\right)=\sum_{\ell=0}^{\infty} u_{i, \ell} \cdot p^{\ell}
$$

where $u_{i, \ell} \in\{0,1, \ldots, p-1\}$. The $\vartheta_{i}$ take the place of the $\vartheta_{i}$ of Section 3.11. Then it is clear from Section 3.11 how to define the p-adic approximation lattices $\Gamma_{\mu}$ for $\mu \in \mathbb{N}_{0}$. Put

$$
\Lambda=\sum_{i=1}^{s-2} x_{i} \cdot \vartheta_{i}-x_{0}
$$

Then Lemma 3.13 yields

$$
\begin{aligned}
\Gamma_{\mu}= & \left.\left\langle\left(x_{1}, \ldots, x_{s-2}, x_{0}\right)\right||\Lambda|_{p} \leqslant p^{-\mu}\right\rangle \\
& \left.=\left.\left\langle\left(x_{1}, \ldots, x_{s-2}, x_{0}\right)\right|\left|\log _{p}\left(\prod_{i=0}^{s-2} p_{i}\right)\right|\right|_{p} \leqslant p^{-\left(\mu+\mu_{0}\right)}\right\rangle
\end{aligned}
$$

where $\mu_{0}=\operatorname{ord}_{p}\left(\log _{p}\left(p_{0}\right)\right)$. In Section 3.13 we studied the set

$$
\left.\Gamma_{\mu}^{*}=\left.\left\langle\left(x_{1}, \ldots, x_{s-2}, x_{0}\right)\right|\right|_{i=0} ^{s-2} p_{i}^{x_{i}} \pm\left. 1\right|_{p} \leqslant p^{-\left(\mu+\mu_{0}\right)}\right\rangle
$$

which is a sublattice of $\Gamma_{\mu}$. In Lemma 3.17 we showed how a basis of $\Gamma_{\mu}^{*}$ can be found from a basis of $\Gamma_{\mu}$. In practice this is very easy, especially if for $p \geqslant 5$ it happens to be possible to choose $p_{0}$ such that not only $\operatorname{ord}_{p}\left(\log _{p}\left(p_{0}\right)\right)$ is minimal, but also $p_{0}$ is a primitive root (mod p). Then, using the notation of Lemma 3.17 (with $\underline{b}_{0}$ as the last element of the basis), choose $\zeta \equiv \mathrm{p}_{0}(\bmod \mathrm{p})$. Then $\mathrm{k}\left(\underline{b}_{0}\right)=1$, and it follows that $\underline{b}_{i}^{\prime}=\underline{b}_{i}$ for $i=1, \ldots, s-2$. By $\underline{b}_{i}=\left(0, \ldots, 1, \ldots, 0, \vartheta_{i}^{(\mu)}\right)^{T}$ we have

$$
p_{i} \cdot p_{0}^{\vartheta} i_{i}^{(\mu)} \equiv \zeta^{k\left(\underline{b}_{i}\right)}\left(\bmod p^{\mu+\mu_{0}}\right)
$$

If $p_{i} \equiv p_{0}^{\alpha}(\bmod p)$, then it follows that

$$
\begin{aligned}
& \gamma_{i}^{*} \equiv \alpha_{i}+\vartheta_{i}^{(\mu)} \equiv \alpha_{i}+\sum_{\ell=0}^{\mu-1} u_{i, \ell}(\bmod (p-1) / 2) \text { for } i=1, \ldots, s-2, \\
& \gamma_{0}^{*}=(p-1) / 2 .
\end{aligned}
$$

Lemma 3.14 (with $c_{1}=0, c_{2}=1$ ) now yields: if

$$
\begin{equation*}
\ell\left(\Gamma_{\mu}^{*}\right)>V(s-1) \cdot X_{1} \tag{6.5}
\end{equation*}
$$

then (6.2) has no solutions with

$$
\begin{equation*}
\mu+\mu_{0} \leqslant \operatorname{ord}_{p}(z) \leqslant m(x \cdot y \cdot z) \leqslant X_{1} . \tag{6.6}
\end{equation*}
$$

### 6.4. Reducing the upper bounds in the one-dimensional case.

In Section 3.10 we have described how an upper bound for the solutions of (6.1) in the case $s=3$ can be reduced. We shall apply that method in this section to the following problem.

THEOREM 6.2. The diophantine equation

$$
\begin{equation*}
x \pm y=w \cdot z \tag{6.7}
\end{equation*}
$$

where $x=p_{0}^{x_{0}}, y=p_{1}^{x_{1}}, \quad z=p^{u}, \quad\left(p, p_{0}, p_{1}\right)=(2,3,5),(3,2,5),(5,2,3)$, $x_{0}, x_{1}, u \in \mathbb{N}_{0}, w \in \mathbb{Z},|w| \leqslant 10^{6}$, and $p \nmid w$, has exactly 291 solutions for $p=2$, 412 solutions for $p=3$, and 570 solutions for $p=5$. In Table $I$ all solutions with $u \geqslant 3$ are given. The solutions with $u \leqslant 2$ satisfy $x_{0} \leqslant 14, x_{1} \leqslant 9$ for $p=2, x_{0} \leqslant 23, x_{1} \leqslant 10$ for $p=3$, and $x_{0} \leqslant 25, x_{1} \leqslant 15$ for $p=5$.

Remark. It is easy to find all solutions of (6.7) with $u \leqslant 2$. The Tables are presented in Section 6.7.

Proof. Put $X=\max _{p=2,3,5} \operatorname{ord}_{p}(x \cdot y \cdot z)$. The example at the end of Section 6.2 shows that in the case $|w|=1$ we have $X<3.98 \times 10^{17}$. It can be checked without difficulties that the effect of the $w$ with $|w| \leqslant 10^{6}$ in the proof of Theorem 6.1 can be neclected (it disappears in the rounding off), so that for the solutions of (6.7) also $X<X_{0}=3.98 \times 10^{17}$ holds. Put

$$
x / y=p_{0}^{y_{0}} \cdot p_{1}^{y_{1}}, \quad \vartheta=-\log _{p}\left(p_{1}\right) / \log _{p}\left(p_{0}\right)
$$

Note that $\vartheta$ is a p-adic integer. Define the lattices $\Gamma_{\mu}, \Gamma_{\mu}^{*}$ as in Section 6.3 , so $\Gamma_{\mu}$ is generated by

$$
\underline{b}_{1}=\binom{1}{\vartheta^{(\mu)}} \quad, \quad \underline{b}_{0}=\binom{0}{p^{\mu}} .
$$

For $p=2,3$ we have $\Gamma_{\mu}^{*}=\Gamma_{\mu}$, and for $p=5$ a basis of $\Gamma_{\mu}^{*}$ is

$$
\underline{b}_{1}^{*}=\underline{b}_{1}-\gamma \cdot \underline{b}_{0}, \quad \underline{b}_{0}^{*}=2 \cdot \underline{b}_{0},
$$

where $\gamma=0$ if $\vartheta^{(\mu)}$ is odd, $\gamma=1$ if $\vartheta^{(\mu)}$ is even. Using the algorithm given in Section 3.10, Fig. 3, we can compute a basis $\underline{c}_{1}, \underline{c}_{2}$ of $\Gamma_{\mu}^{*}$ that is reduced in the sense that $\left|\underline{c}_{1}\right|=\ell\left(\Gamma_{\mu}^{*}\right)$. We did so, with $\mu$ as
in the following table.

| p | $\mathrm{p}_{0}$ | $\mathrm{p}_{1}$ | $\mu_{0}$ | $\mu$ | $\gamma$ | $\left\|\underline{c}_{1}\right\|>$ | $\mathrm{u} \leqslant$ | W | $\left\|y_{0}\right\| \leqslant$ | $\left\|y_{1}\right\| \leqslant$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 3 | 5 | 2 | 143 | - | $2.68 \times 10^{21}$ | 144 | $10^{6} \cdot 2^{144}$ | 114 | 78 |
| 3 | 2 | 5 | 1 | 91 | - | $2.32 \times 10^{21}$ | 91 | $10^{6} \cdot 3^{91}$ | 182 | 78 |
| 5 | 2 | 3 | 1 | 65 | 0 | $5.28 \times 10^{22}$ | 65 | $10^{6} \cdot 5^{65}$ | 189 | 119 |

The values of $\vartheta^{(\mu)}$ can be found in Table III. Making an exception to our policy, we give the reduced bases of the $\Gamma_{\mu}^{*}$ below (in base $p$ notation):
$p=2$
$\varsigma_{1}=\left(\begin{array}{rrrrrrr}10 & 00000 & 00100 & 10001 & 10110 & 01110 & 01101 \\ 00001 & 11101 & 00101 & 00100 & 11100 & 01111 & 11010 \\ 00011 \\ & -1 & 00010 & 00110 & 01000 & 01011 & 01110 \\ 00010 \\ 00101 & 11000 & 00000 & 11100 & 01111 & 01011 & 10111 \\ 00001\end{array}\right)$,

$$
\underline{c}_{2}=\left(\begin{array}{rrrrrrr}
10 & 11011 & 10000 & 01011 & 01101 & 11000 & 00111 \\
11001 & 10100 & 11011 & 00000 & 11111 & 10110 & 10110
\end{array} 00001\right.
$$

$p=3: \quad \underline{c}_{1}=\left(\begin{array}{llllllll}-102 & 01121 & 02221 & 00210 & 12120 & 20020 & 22222 & 10212 \\ 20222 \\ 21002 & 00122 & 21100 & 11102 & 22102 & 20001 & 11222 & 02212\end{array} 21011\right)$, $\underline{c}_{2}=\left(\begin{array}{lllllllll}-10 & 12210 & 12111 & 01102 & 02010 & 12112 & 12210 & 21122 & 21011 \\ 20102 \\ -2 & 22021 & 11012 & 01000 & 12021 & 00211 & 12221 & 22121 & 21220 \\ 12122\end{array}\right)$,
$p=5: \quad c_{1}=\left(\begin{array}{rrrrrr}-211 & 32230 & 21042 & 22023 & 30141 & 33034 \\ 21420 \\ - & 22104 & 43102 & 43111 & 03114 & 30134 \\ 23410\end{array}\right)$, $\underline{c}_{2}=\left(\begin{array}{rrrrrrr}340 & 34003 & 02404 & 12120 & 03412 & 22030 & 32211 \\ -414 & 20001 & 42202 & 42210 & 34043 & 20120 & 00432\end{array}\right)$.

From this we found the lower bounds for $\left|\underline{c}_{1}\right|$ given above. They are all larger than $\sqrt{ } 2 \cdot 3.98 \times 10^{17}$. Hence (6.5) holds for $X_{1}=X_{0}$, and then we infer from (6.6) that $u \leqslant \mu+\mu_{0}-1$, and $|w| \cdot z \leqslant W$ as shown in the table above. We now find the new upper bounds for $\left|y_{0}\right|,\left|y_{1}\right|$ as follows. If in (6.7) the minus sign holds, supposing that $\min (x, y)>W^{10 / 9}$, we infer

$$
|x-y|=|w| \cdot z \leqslant W<\min (x, y)^{0.9}
$$

By Theorem 5.2(a), the inequality $|x-y|<\min (x, y)^{0.9}$ has no solutions with $\min (x, y)>W$, since $W>10^{49}$. Hence $\min (x, y) \leqslant W^{10 / 9}$, and thus

$$
\max (x, y) \leqslant \min (x, y)+|w| \cdot z \leqslant w^{10 / 9}+w
$$

If in (6.7) the plussign holds, then this inequality follows at once. So now the bounds given in the above table for $\left|y_{0}\right|,\left|y_{1}\right|$ follow from

$$
\left|y_{i}\right| \cdot \log p_{i} \leqslant \log \max (x, y) \leqslant \log \left(W^{10 / 9}+W\right)
$$

We repeat the procedure with $\mu$ as in the following table.

| p | $\mu$ | $\gamma$ | $\left\|\underline{c}_{1}\right\|>$ | $\sqrt{ } 2 \cdot \mathrm{X}_{0}<$ | $u \leqslant$ | w | $\left\|y_{0}\right\| \leqslant$ | $\left\|y_{1}\right\| \leqslant$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 16 | - | 167.7 | 161.3 | 17 | $10^{6} \cdot 2^{17}$ | 31 | 21 |
| 3 | 13 | - | 535.8 | 257.4 | 13 | $10^{6} \cdot 3^{13}$ | 49 | 21 |
| 5 | 7 | 1 | 276.1 | 267.3 | 7 | $10^{6} \cdot 5^{7}$ | 49 | 31 |

The numbers are now so small that the computations can be performed by hand. For example, for $p=5$, the lattice $\Gamma_{7}^{*}$ is generated by

$$
\underline{b}_{1}^{*}=\binom{1}{-45607}, \quad \underline{b}_{0}^{*}=\binom{0}{156250}
$$

and a reduced basis is

$$
\underline{c}_{1}=\binom{185}{205}, \quad \underline{c}_{0}=\binom{-394}{408}
$$

We find upper bounds for $u$ and $W$ as given in the above table. In all three cases, $W^{10 / 9}<10^{15}$. On supposing $\min (x, y)>10^{15}$ we infer

$$
|x-y|=|w| \cdot z \leqslant w<10^{15 \cdot 0.9} \leqslant \min (x, y)^{0.9}
$$

By Theorem 5.2(a) we see that the inequality $|x-y|<\min (x, y)^{0.9}$ has only two solutions: $(x, y)=\left(2^{65}, 5^{28}\right),\left(2^{84}, 3^{53}\right)$. However, both have $|x-y|>10^{15 \cdot 0.9}$. So we infer $\min (x, y) \leqslant 10^{15}$, hence by $\max (x, y) \leqslant 10^{15}+W$ we obtain the bounds for $\left|y_{0}\right|,\left|y_{1}\right|$ as given above. These bounds are small enough to admit enumeration of the remaining cases. a

Remark. The computer calculations for the above proof took less than 1 sec.

### 6.5. Reducing the upper bounds in the multi-dimensional case.

In Section 3.11 we have described how an upper bound for the solutions of (6.1) in the case $s \geqslant 3$ can be reduced. We shall apply that method in this section to the following problem.

THEOREM 6.3. The diophantine equation

$$
\begin{equation*}
x+y=z \tag{6.8}
\end{equation*}
$$

$$
\begin{gathered}
\text { in } x, y, \quad z \in S=\left\langle 2^{x_{1}} \cdot \ldots \cdot 13^{x_{6}}\right| x_{i} \in \mathbb{N}_{0} \text { for } i=1, \ldots, 6>\text { with } \\
(x, y)=1 \text { and } x \leqslant y \text { has exactly } 545 \text { solutions. of them, } 514 \text { satisfy } \\
\operatorname{ord}_{2}(x \cdot y \cdot z) \leqslant 12, \operatorname{ord}_{3}(x \cdot y \cdot z) \leqslant 7, \operatorname{ord}_{5}(x \cdot y \cdot z) \leqslant 5, \\
\operatorname{ord}_{7}(x \cdot y \cdot z) \leqslant 4, \operatorname{ord}_{11}(x \cdot y \cdot z) \leqslant 3, \operatorname{ord}_{13}(x \cdot y \cdot z) \leqslant 3 .
\end{gathered}
$$

The remaining 31 solutions are given in Table II.

Remark. From Theorem 6.3 it is easy to compute all 545 solutions of (6.8).

Proof. In the example at the end of Section 6.2 we have seen that $m(x \cdot y \cdot z)<X_{0}=5.60 \times 10^{27}$. With the notation of Section 6.3 we choose the following parameters.

| p | $\mathrm{p}_{0}$ | $\mathrm{p}_{1}$ | $\mathrm{p}_{2}$ | $\mathrm{p}_{3}$ | $\mathrm{p}_{4}$ | $\mu_{0}$ | $\mu$ | $\gamma_{0}^{*}$ | $\gamma_{1}^{*}$ | $\gamma_{2}^{*}$ | $\gamma_{3}^{*}$ | $\gamma_{4}^{*}$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| 2 | 3 | 5 | 7 | 11 | 13 | 2 | 605 | - | - | - | - | - |
| 3 | 2 | 5 | 7 | 11 | 13 | 1 | 385 | - | - | - | - | - |
| 5 | 2 | 3 | 7 | 11 | 13 | 1 | 275 | 2 | 0 | 1 | 1 | 1 |
| 7 | 3 | 2 | 5 | 11 | 13 | 1 | 220 | 3 | 0 | -1 | -1 | 0 |
| 11 | 2 | 3 | 5 | 7 | 13 | 1 | 165 | 5 | 2 | 0 | -1 | -1 |
| 13 | 2 | 3 | 5 | 7 | 11 | 1 | 165 | 6 | -2 | -1 | -2 | 3 |

We computed the six values of the $\vartheta_{i}^{(\mu)}$ for $i=1,2,3,4$ (and give them in Table III), and the reduced bases of the six lattices $\Gamma_{\mu}^{*}$, by the $L^{3}$-algorithm. Thus we obtained lower bounds for $\ell\left(\Gamma_{\mu}^{*}\right)$ as in the following table. They are all larger than $\sqrt{ } 5 \cdot 5.60 \times 10^{27}$ (note that we have a very large margin here, we could have taken the $\mu$ 's probably about $20 \%$ smaller). So we apply Lemma 3.14 for $X_{1}=X_{0}=5.60 \times 10^{27}$. For every $p$ we thus find $\operatorname{ord}_{p}(z) \leqslant \mu+\mu_{0}-1$. Since (6.2) is invariant under permutations of $x, y$, $z$, we even have $\operatorname{ord}_{p}(x \cdot y \cdot z) \leqslant \mu+\mu_{0}-1$, as shown in the next table.

| $p$ | $\ell\left(\Gamma_{\mu}^{*}\right) \geqslant I \underline{c}_{1} \mid / 4>$ | $\operatorname{ord}_{p}(x \cdot y \cdot z) \leqslant$ |
| :---: | :---: | :---: |
| 2 | $4.70 \times 10^{35}$ | 606 |
| 3 | $1.15 \times 10^{36}$ | 385 |
| 5 | $6.27 \times 10^{37}$ | 275 |
| 7 | $3.17 \times 10^{36}$ | 220 |
| 11 | $5.74 \times 10^{33}$ | 165 |
| 13 | $1.73 \times 10^{36}$ | 165 |

Hence $m(x \cdot y \cdot z) \leqslant 606$.
We repeated the procedure with $X_{0}=606$ and $\mu$ as in the following table. After computing the reduced bases of the six lattices $\Gamma_{\mu}^{*}$ we found the following data. Note that in all cases $\ell\left(\Gamma_{\mu}^{*}\right) \geqslant \sqrt{5} \cdot 606$.

| p | $\mu$ | $\gamma_{0}^{*}$ | $\gamma_{1}^{*}$ | $\gamma_{2}^{*}$ | $\gamma_{3}^{*}$ | $\gamma_{4}^{*}$ | $\ell\left(\Gamma_{\mu}^{*}\right)>$ | $\operatorname{ord}_{\mathrm{p}}(\mathrm{x} \cdot \mathrm{y} \cdot \mathrm{z}) \leqslant$ |
| ---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 66 | - | - | - | - | - | 1909 | 67 |
| 3 | 42 | - | - | - | - | - | 2304 | 42 |
| 5 | 30 | 2 | 0 | 0 | 1 | 1 | 3417 | 30 |
| 7 | 24 | 3 | -1 | 0 | 1 | -1 | 2391 | 24 |
| 11 | 18 | 5 | 0 | -2 | 2 | -1 | 1443 | 18 |
| 13 | 18 | 6 | 0 | 1 | 1 | -2 | 3196 | 18 |

Hence $m(x \cdot y \cdot z) \leqslant 67$. Next, we repeated the procedure with $X_{0}=67$, and $\mu$ as in the following table. We found

| p | $\mu$ | $\gamma_{0}^{*}$ | $\gamma_{1}^{*}$ | $\gamma_{2}^{*}$ | $\gamma_{3}^{*}$ | $\gamma_{4}^{*}$ | $\ell\left(\Gamma_{\mu}^{*}\right)>$ | $\operatorname{ord}_{p}(x \cdot y \cdot z) \leqslant$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 55 | - | - | - | - | - | 364 | 56 |
| 3 | 35 | - | - | - | - | - | 301 | 35 |
| 5 | 25 | 2 | 1 | 1 | 1 | 0 | 622 | 25 |
| 7 | 20 | 3 | -1 | 1 | -1 | 0 | 693 | 20 |
| 11 | 15 | 5 | -1 | -2 | 2 | 2 | 192 | 15 |
| 13 | 15 | 6 | -1 | 0 | 3 | -2 | 658 | 15 |

Hence $m(x \cdot y \cdot z) \leqslant 56$.

To find the solutions of (6.2) with ord $(x \cdot y \cdot z)$ below the bounds given in the above table, we followed the following procedure. Suppose that we are at a certain moment interested in finding the solutions with ord $(x \cdot y \cdot z) \leqslant f(p)$ where $f(p)$ is given for $p=2, \ldots, 13$. Choose $p$, and $\mu<f(p)-\mu_{0}$,
and consider the lattice $\Gamma_{\mu}^{*}$ for these values of $p, \mu$. If a solution $\mathrm{x}, \mathrm{y}, \mathrm{z}$ of (6.2) exists with $\operatorname{ord}_{\mathrm{p}}(\mathrm{z}) \geqslant \mu+\mu_{0}$, then the vector $\left(x_{1}, \ldots, x_{4}, x_{0}\right)^{T}$ with $x_{i}=\operatorname{ord}_{p_{i}}(x / y)$ for $i=0, \ldots, 4$, is in the lattice. Its length is bounded by $V\left(\mathrm{f}\left(\mathrm{p}_{0}\right)^{2}+\ldots+\mathrm{f}\left(\mathrm{p}_{4}\right)^{2}\right)$. All vectors in $\Gamma_{\mu}^{*}$ with length below this bound can be computed by the algorithm of Fincke and Pohst, as given in Section 3.6. Then all solutions of (6.2) corresponding to lattice points can be selected. Then we replace $f(p)$ by $\mu+\mu_{0}-1$, and we repeat the procedure for newly chosen $\mathrm{p}, \mu$.
We performed this procedure, starting with the bounds for $\operatorname{ord}_{p}(x \cdot y \cdot z)$ given in the above table for $f(p)$, and with $p, m$ as in Table IV (where \# stands for the number of solutions of (5.2) found at that stage). At the end we have $f(2)=4, f(p)=1$ for $p=3, \ldots, 13$. The remaining solutions can be found by hand.

Remarks. 1. Theorems 6.2 and 6.3 have applications in group theory (cf. Alex [1976]). We use Theorem 6.3 in Section 7.2.
2. The computer calculations for the proof of Theorem 6.3 took 438 sec ., of which 412 were used for the first reduction step. In this first step we applied the $L^{3}$-algorithm in 11 steps (cf. Section 3.5), which cost on average about 60 sec . per lattice. The remaining 50 sec . were mainly used for the computation of the $24 \vartheta_{i}^{(\mu)}$,s.

### 6.6. Examples related to the abc-conjecture.

Let $x, y, z$ be positive integers. Put

$$
G=\prod_{\substack{p l x y z \\ p \text { prime }}} p .
$$

For all $x, y, z$ with $(x, y)=1$ and $x+y=z$ we define

$$
c(x, y, z)=\log z / \log G
$$

(called the Masser-ratio, according to Tijdeman [1989]). Recently, Oesterlé posed the problem to decide whether there exists an absolute constant $C$ such that $c(x, y, z)<C$ for all $x, y, z$. Masser [1985] conjectured the stronger assertion that $c(x, y, z)<1+\varepsilon$, when $z$ exceeds some bound depending on $\varepsilon$ only, for all $\varepsilon>0$. For a survey of related results and conjectures, see Stewart and Tijdeman [1986], Vojta [1987], Tijdeman [1989].

It might be interesting to have some empirical results on $c(x, y, z)$, and to search for $x, y, z$ for which it is large. From the preceding sections it may be clear that such $x, y, z$ correspond to relatively short vectors in appropriate p-adic approximation lattices.

As a byproduct of the proofs of Theorems 5.5 and 6.3 we computed the value of $c(x, y, z)$, corresponding to many short vectors that we came across in performing the algorithm of Fincke and Pohst. All examples that we found with $c(x, y, z) \geqslant 1.4$ are listed below. Our search was rather unsystematic, so we do not guarantee that this list is complete in any sense.

| $x$ | $y$ | $z$ | $c(x, y, z)$ |
| :--- | :--- | :--- | :--- |
| $11^{2}$ | $3^{2} \cdot 5^{6} \cdot 7^{3}$ | $2^{21} \cdot 23$ | 1.62599 |
| 1 | $2 \cdot 3^{7}$ | $5^{4} \cdot 7$ | 1.56789 |
| $7^{3}$ | $3^{10}$ | $2^{11} \cdot 29$ | 1.54708 |
| $5^{2} \cdot 7937$ | $7^{13}$ | $2^{18} \cdot 3^{7} \cdot 13^{2}$ | 1.49762 |
| $11^{2}$ | $3^{9} \cdot 13$ | $2^{11} \cdot 5^{3}$ | 1.48887 |
| 37 | $2^{15}$ | $3^{8} \cdot 5$ | 1.48291 |
| $2^{7} \cdot 5^{2}$ | $7^{6} \cdot 41$ | $13^{6}$ | 1.46192 |
| 1 | $2^{5} \cdot 3^{2} \cdot 5^{2}$ | $7^{4}$ | 1.45567 |
| $2^{19} \cdot 13 \cdot 103$ | $7^{11}$ | $3^{11} \cdot 5^{3} \cdot 11^{2}$ | 1.45261 |
| 1 | $2^{12} \cdot 5^{3}$ | $3^{5} \cdot 7^{2} \cdot 43$ | 1.44331 |
| 1 | $2^{4} \cdot 3^{7} \cdot 547$ | $5^{8} \cdot 7^{2}$ | 1.43906 |
| $2^{10} \cdot 7$ | $5^{7}$ | $3^{8} \cdot 13$ | 1.43501 |
| 3 | $5^{3}$ | $2^{7}$ | 1.42657 |
| 5 | $3^{11}$ | $2^{10} \cdot 173$ | 1.41268 |

Two more examples with $c(x, y, z) \geqslant 1.4$ are known:

$$
x=1, \quad y=3 \cdot 5 \cdot 47^{2}, \quad z=2^{18} \cdot 79, \quad c(x, y, z)=1.44965
$$

found by G. Frey (communicated to us by Prof. F. Oort), and

$$
x=2, \quad y=109 \cdot 3^{10}, \quad z=23^{5}, \quad c(x, y, z)=1.62991
$$

found by E. Reyssat (communicated to us by Prof. M. Waldschmidt), which wins the race. Note that these two examples show large primes at two places.

These results do not seem to yield any heuristical evidence for the truth or falsity of the abc-conjecture.
6.7. Tables.

Table I. (Theorem 6.2.)

$$
p=2, p_{0}=3, p_{1}=5
$$

| $x_{0}$ | $p_{0}^{x_{0}}$ | $x_{1}$ | $p_{1}^{x_{1}}$ | sign | $u$ | $w$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| 2 | 9 | 10 | 9765625 | -1 | 4 | -610351 |
| 10 | 59049 | 10 | 9765625 | -1 | 4 | -606661 |
| 4 | 81 | 12 | 244140625 | -1 | 9 | -476837 |
| 6 | 729 | 10 | 9765625 | -1 | 5 | -305153 |
| 2 | 9 | 8 | 390625 | -1 | 3 | -48827 |
| 6 | 729 | 8 | 390625 | -1 | 3 | -48737 |
| 10 | 59049 | 8 | 390625 | -1 | 3 | -41447 |
| 14 | 4782969 | 10 | 9765625 | -1 | 7 | -38927 |
| 4 | 81 | 8 | 390625 | -1 | 4 | -24409 |
| 0 | 1 | 8 | 390625 | -1 | 5 | -12207 |
| 8 | 6561 | 8 | 390625 | -1 | 6 | -6001 |
| 0 | 1 | 6 | 15625 | -1 | 3 | -1953 |
| 4 | 81 | 6 | 15625 | -1 | 3 | -1943 |
| 8 | 6561 | 6 | 15625 | -1 | 3 | -1133 |
| 6 | 729 | 6 | 15625 | -1 | 4 | -931 |
| 2 | 9 | 4 | 625 | -1 | 3 | -77 |
| 2 | 9 | 6 | 15625 | -1 | 8 | -61 |
| 0 | 1 | 4 | 625 | -1 | 4 | -39 |
| 4 | 81 | 4 | 625 | -1 | 5 | -17 |
| 0 | 1 | 2 | 25 | -1 | 3 | -3 |
| 2 | 9 | 2 | 25 | -1 | 4 | -1 |
| 1 | 3 | 1 | 5 | 1 | 3 | -1 |
| 1 | 3 | 3 | 125 | 1 | 7 | 1 |
| 2 | 9 | 0 | 1 | -1 | 3 | 1 |
| 3 | 27 | 1 | 5 | 1 | 5 | 1 |
| 4 | 81 | 0 | 1 | -1 | 4 | 1 |
| 4 | 81 | 2 | 25 | -1 | 3 | 1 |
| 6 | 729 | 2 | 25 | -1 | 6 | 5 |
| 6 | 729 | 4 | 625 | -1 | 3 | 7 |
| 3 | 27 | 3 | 125 | 1 | 3 | 11 |
| 5 | 243 | 3 | 125 | 1 | 4 | 19 |
| 5 | 243 | 1 | 5 | 1 | 3 | 23 |
| 7 | 2187 | 5 | 1 | -1 | 6 | 3 |

Table I. (cont.)

| $x_{0}$ | $p_{0}^{\chi_{0}}$ | $x_{1}$ |  | $p_{1}^{x_{1}}$ | sign | $u$ | $w$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 3 | 5 |  | 3125 | 1 | 3 | 391 |
| 5 | 243 | 5 | , | 3125 | 1 | 3 | 421 |
| 9 | 19683. | 3 |  | 125 | 1 | 5 | 619 |
| 8 | $6561{ }^{\circ}$ | 2 |  | 25 | -1 | 3 | 817 |
| 10 | 59049 | 6 |  | 15625 | -1 | 5 | 1357 |
| 5 | 243 | 7 |  | 78125 | 1 | 5 | 2449 |
| 9 | 19683 | 1 |  | 5 | 1 | 3 | 2461 |
| 9 | 19683 | 5 |  | 3125 | 1 | 3 | 2851 |
| 10 | 59049 | 2 |  | 25 | -1 | 4 | 3689 |
| 12 | 531441 | 4 |  | 625 | -1 | 7 | 4147 |
| 1 | 3 | 7 |  | 78125 | 1 | 4 | 4883 |
| 9 | 19683 | 7 |  | 78125 | 1 | 4 | 6113 |
| 13 | 1594323 | 7 |  | 78125 | 1 | 8 | 6533 |
| 10 | 59049 | 4 |  | 625 | -1 | 3 | 7303 |
| 10 | 59049 | 0 |  | 1 | -1 | 3 | 7381 |
| 12 | 531441 | 8 |  | 390625 | -1 | 4 | 8801 |
| 3 | 27 | 7 |  | 78125 | 1 | 3 | 9769 |
| 7 | 2187 | 7 |  | 78125 | 1 | 3 | 10039 |
| 11 | 177147 | 5 |  | 3125 | 1 | 4 | 11267 |
| 3 | 27 | 9 |  | 1953125 | 1 | 7 | 15259 |
| 11 | 177147 | 3 |  |  | 1 | 3 | 22159 |
| 11 | 177147 | 7 |  | 78125 | 1 | 3 | 31909 |
| 12 | 531441 | 0 |  | 1 | -1 | 4 | 33215 |
| 12 | 531441 | 6 |  | 15625 | -1 | 3 | 64477 |
| 12 | 531441 | 2 |  | 25 | -1 | 3 | 66427 |
| 11 | 177147 | 9 |  | 1953125 | 1 | 5 | 66571 |
| 13 | 1594323 | 3 |  | 125 | 1 | 4 | 99653 |
| 7 | 2187 | 9 |  | 1953125 | 1 | 4 | 122207 |
| 14 | 4782969 | 2 |  | 25 | - 1 | 5 | 149467 |
| 13 | 1594323 | 1 |  | 5 | 1 | 3 | 199291 |
|  | 1594323 | 5 |  | 3125 | 1 | 3 | 199681 |
| 1 | 3 | 9 |  | 1953125 | 1 | 3 | 244141 |
| 5 | 243 | 9 |  | 1953125 | 1 | 3 | 244171 |
| 9 | 19683 | 9 |  | 1953125 | 1 | 3 | 246601 |
| 14 | 4782969 | 6 |  | 15625 | -1 | 4 | 297959 |
| 13 | 1594323 | 9 |  | 1953125 | 1 | 3 | 443431 |
| 15 | 14348907 | 5 |  | 3125 | 1 | 5 | 448501 |
| 14 | 4782969 | 8 |  | 390625 | -1 | 3 | 549043 |
| 14 | 4782969 | 4 |  | 625 | -1 | 3 | 597793 |
| 14 | 4782969 | 0 |  | 1 | $-1$ | 3 | 597871 |
| 16 | 43046721 | 0 |  | 1 | -1 | 6 | 672605 |
| 9 | 19683 | 11 |  | 48828125 | 1 | 6 | 763247 |
| 15 | 14348907 | 1 |  | 5 | 1 | 4 | 896807 |

Table continued

Table I. (cont.)

$$
p=3, p_{0}=2, p_{1}=5
$$

| $x_{0}$ | $p_{0}^{x_{0}}$ | $x_{1}$ | $p_{1}^{x_{1}}$ | sign | $u$ | $w$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 14 | 16384 | 10 | 9765625 | -1 | 4 | -120361 |
| 9 | 512 | , | 1953125 | -1 | 3 | -72319 |
| 4 | 16 | 8 | 390625 | -1. | 3 | -14467 |
| 12 | 4096 | 6 | 15625 | -1 | 3 | -427 |
| 7 | 128 | 5 | 3125 | -1 | 4 | -37 |
| 2 | 4 | 4 | 625 | -1 | 3 | -23 |
| 1 | 2 | 2 | 25 | 1 | 3 | 1 |
| 5 | 32 | 1 | 5 | -1 | 3 | 1 |
| 6 | 64 | 3 | 125 | 1 | 3 | 7 |
| 11 | 2048 | 4 | 625 | 1 | 5 | 11 |
| 9 | 512 | 0 | 1 | 1 | 3 | 19 |
| 10 | 1024 | 2 | 25 | -1 | 3 | 37 |
| 3 | 8 | 6 | 15625 | 1 | 4 | 193 |
| 15 | 32768 | 3 | 125 | -1 | 4 | 403 |
| 14 | 16384 | 1 | 5 | 1 | 3 | 607 |
| 17 | 131072 | 7 | 78125 | -1 | 3 | 1961 |
| 16 | 65536 | 5 | 3125 | 1 | 3 | 2543 |
| 8 | 256 | 7 | 78125 | 1 | 3 | 2903 |
| 19 | 524288 | 2 | 25 | 1 | 4 | 6473 |
| 18 | 262144 | 0 | 1 | -1 | 3 | 9709 |
| 23 | 8388608 | 1 | 5 | -1 | 6 | 11507 |
| 13 | 8192 | 8 | 390625 | 1 | 3 | 14771 |
| 22 | 4194304 | 8 | 390625 | -1 | 5 | 15653 |
| 10 | 1024 | 11 | 48828125 | 1 | 7 | 22327 |
| 18 | 262144 | 9 | 1953125 | 1 | 4 | 27349 |
| 20 | 1048576 | 4 | 625 | -1 | 3 | 38813 |
| 0 | 1 | 9 | 1953125 | 1 | 3 | 72338 |
| 21 | 2097152 | 6 | 15625 | 1 | 3 | 78251 |
| 5 | 32 | 10 | 9765625 | 1 | 3 | 361691 |
| 24 | 16777216 | 3 | 125 | 1 | 3 | 621383 |
| 23 | 8388608 | 10 | 9765625 | 1 | 3 | 672379 |
| 26 | 67108864 | 7 | 78125 | 1 | 4 | 829469 |
| $p=5, p_{0}=2, p_{1}=3$ |  |  |  |  |  |  |
| $x_{0}$ | $p_{0}^{x_{0}}$ | $x_{1}$ | $p_{1}^{x_{1}}$ | sign | $u$ | $w$ |
| 12 | 4096 | 16 | 43046721 | -1 | 3 | -344341 |
| 5 | 32 | 15 | 14348907 | -1 | 3 | -114791 |
| 7 | 128 | 1 | 3 | -1 | 3 | 1 |
| 6 | 64 | 8 | 6561 | 1 | 3 | 53 |
| 14 | 16384 | 2 | 9 | -1 | 3 | 131 |
| 13 | 8192 | 9 | 19683 | 1 | 3 | 223 |
| 20 | 1048576 | 10 | 59049 | 1 | 3 | 8861 |
| 21 | 2097152 | 3 | 27 | -1 | 3 | 16777 |



Table III.



Table III. (cont.)

| $-\log _{5} 11 / \log _{5} 2=$ |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0.44032 | 21012 | 13124 | 21134 | 03320 | 33422 | 21041 | 12112 | 42420 | 00220 | 41143 | 12040 | 32144 | 21100 | 01304 | 24013 | 43401 | 23313 | 12022 | 34404 |
| - 12413 | 10214 | 30123 | 11014 | 24110 | 42444 | 42030 | 02413 | 20241 | 22304 | 23423 | 13414 | 03234 | 30000 | 10334 | 44322 | 00330 | 01104 | 44410 | 44113 |
| 31022 | 33142 | 14441 | 44113 | 21413 | 23132 | 31413 | 32032 | 01221 | 40210 | 24101 | 30133 | 13110 | 13400 | 22110 | 2334. |  |  |  |  |
|  | $-10$ | ${ }_{5}^{13}$ | $\log _{5} 2$ |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |
| 0.12423 | 02224 | 01323 | 24314 | 23021 | 32420 | 14134 | 41224 | 04403 | 11334 | 43213 | 33303 | 03130 | 32244 | 11133 | 43243 | 23422 | 11320 | 41041 | 31134 |
| 41320 | 34110 | 03024 | 40012 | 23213 | 10014 | 41441 | 04420 | 40114 | 00021 | 33224 | 30103 | 03243 | 32031 | 22021 | 20234 | 32441 | 00013 | 04203 | 43134 |
| 22012 | 30332 | 22422 | 43110 | 13302 | 34431 | 13241 | 13230 | 44204 | 14432 | 33210 | 24121 | 13144 | 03230 | 14301 | 3040. |  |  |  |  |

$-\log _{7} 2 / \log _{7} 3=$
 $-\log _{7} 5 / \log _{7} 3=$
$7 \begin{array}{ll}7 & 7\end{array}$



0.621 B3 15581 0A077 3B5C8

$-\log _{13} 5 / \log _{13} 2$
$0.4457079 \mathrm{C} 5173665 \quad 3796 \mathrm{C}$
 $0 A 414$ 32B3C A67A8 $48 \ldots$
$\log 7 / \log 2=$
$\log _{13} \log _{13} 2$



Table IV.

| nr . | p | m | \# | nr . | p | m | \# |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 2 | 44 | - | 27 | 2 | 13 | 1 |
| 2 | 3 | 28 | - | 28 | 2 | 12 | 2 |
| 3 | 5 | 20 | - | 29 | 2 | 11 | 2 |
| 4 | 7 | 16 | - | 30 | 3 | 13 | - |
| 5 | 11 | 12 | - | 31 | 3 | 12 | - |
| 6 | 13 | 12 | - | 32 | 3 | 11 | - |
| 7 | 2 | 33 | - | 33 | 3 | 10 | 1 |
| 8 | 3 | 21 | - | 34 | 3 | 9 | 1 |
| 9 | 5 | 15 | - | 35 | 3 | 8 | 1 |
| 10 | 7 | 12 | - | 36 | 3 | 7 | 6 |
| 11 | 11 | 9 | - | 37 | 5 | 9 | - |
| 12 | 3 | 9 | - | 38 | 5 | 8 | - |
| 13 | 2 | 22 | - | 39 | 5 | 7 | - |
| 14 | 3 | 14 | - | 40 | 5 | 6 | - |
| 15 | 5 | 10 | - | 41 | 5 | 5 | 6 |
| 16 | 7 | 8 | - | 42 | 7 | 7 | - |
| 17 | 11 | 6 | - | 43 | 7 | 6 | - |
| 18 | 13 | 6 | - | 44 | 7 | 5 | 1 |
| 19 | 2 | 21 | - | 45 | 7 | 4 | 4 |
| 20 | 2 | 20 | - | 46 | 11 | 5 | - |
| 21 | 2 | 19 | - | 47 | 11 | 4 | 1 |
| 22 | 2 | 18 | - | 48 | 11 | 3 | 4 |
| 23 | 2 | 17 | - | 49 | 13 | 5 | - |
| 24 | 2 | 16 | - | 50 | 13 | 4 | - |
| 25 | 2 | 15 | - | 51 | 13 | 3 | 1 |
| 26 | 2 | 14 | - |  |  |  |  |



## Chapter 7. The sum of two $S$-units being a square.

### 7.1. Introduction.

Let $p_{1}, \ldots, p_{s}(s \geqslant 1)$ be distinct primes, and let $S$ be the set of positive rational integers which have no prime divisors different from the $p_{i}$. A rational number is called an $S$-unit if its absolute value is a quotient of elements of $S$. Thus the set of $S$-units is

$$
\left.\left\langle \pm p_{1}^{x_{1}} \cdot \ldots \cdot p_{s}^{x_{s}}\right| x_{i} \in \mathbb{Z} \text { for } i=1, \ldots, s\right\rangle
$$

We study the diophantine equation

$$
x+y=z^{2}
$$

in S-units $x, y$, and $z \in \mathbb{Q}$, where the set of primes $p_{1}, \ldots, p_{s}$ is given. We show how to find all solutions of this equation, using the theory of p-adic linear forms in logarithms, and a computational p-adic diophantine approximation method. We actually perform all the necessary computations for solving the equation completely for $\left\{p_{1}, \ldots, p_{s}\right\}=\{2,3,5,7\}$. This type of equations has applications in arithmetic algebraic geometry (cf. Setzer [1975], Pinch [1984]).

We start with getting rid of the denominators. Let $x, y, z$ be a solution. There is a $d \in S$ such that $|d \cdot x|,|d \cdot y| \in S$. Put $d=d_{1} \cdot d_{2}^{2}$, where $d_{1}, d_{2} \in S$ and $d_{1}$ squarefree. Then

$$
\mathrm{d}_{1} \cdot \mathrm{~d} \cdot \mathrm{x}+\mathrm{d}_{1} \cdot \mathrm{~d} \cdot \mathrm{y}=\left(\mathrm{d}_{1} \cdot \mathrm{~d}_{2} \cdot \mathrm{z}\right)^{2}
$$

which has the same form as $x+y=z^{2}$, but now $\left|d_{1} \cdot d \cdot x\right|,\left|d_{1} \cdot d \cdot y\right| \in S \subset \mathbb{Z}$ and $d_{1} \cdot d_{2} \cdot z \in \mathbb{Z}$. Without loss of generality we may therefore study

$$
\begin{equation*}
x+y=z^{2} \tag{7.1}
\end{equation*}
$$

where

$$
\left\{\begin{array}{l}
x \in S, \quad \pm y \in S, \quad z \in \mathbb{Z}  \tag{7.2}\\
x \geqslant y, \quad z>0 \\
(x, y) \text { is squarefree }
\end{array}\right.
$$

We shall prove the following results.

THEOREM 7.1. Let $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{s}}$ be given. There exists an effectively computable constant C , depending on $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{s}}$ only, such that any solution $\mathrm{x}, \mathrm{y}, \mathrm{z}$ of equation (7.1) with conditions (7.2) satisfies $\max (x,|y|, z)<C$.

THEOREM 7.2. Let $\left\{p_{1}, \ldots, p_{s}\right\}=\{2,3,5,7\}$. Equation (7.1) with conditions (7.2) has exactly the 388 solutions given in Table I.

Remarks. 1. The Tables are given in Section 7.9. We stress that the aim of this chapter is not only to prove these theorems, but to show as well that for any given set of primes $\left\{p_{1}, \ldots, p_{s}\right\}$ a result similar to Theorem 7.2 can be proved along the same lines, in a more or less algorithmic way.
2. Equation (7.1) with conditions (7.2) can be seen as a further generalization of the generalized Ramanujan-Nagell equation

$$
\begin{equation*}
x^{2}+k=p_{1}^{n_{1}} \cdot \ldots \cdot p_{s}^{n_{s}} \tag{7.3}
\end{equation*}
$$

(cf. Chapter 4), namely by taking $|k| \in S$ arbitrary instead of $k \in \mathbb{Z}$ fixed. The method of this chapter to solve (7.1) is also a generalization of the method of Chapter 4 to solve (7.3).

Equation (7.1) can be transformed into a number of Pell-like equations. Put

$$
\mathrm{x}=\mathrm{D} \cdot \mathrm{u}^{2}
$$

where $D, u \in S$, and $D$ is squarefree. There are only $2^{S}$ possibilities for D. Now, (7.1) is equivalent to a finite number of equations

$$
\begin{equation*}
z^{2}-D \cdot u^{2}=y \tag{7.4}
\end{equation*}
$$

in $u \in S, \quad \pm y \in S, \quad z \in \mathbb{Z}$, with $z>0$ and $(u, y)=1$. We treat equation (7.4) by factorizing its both sides in the field $K=\mathbb{Q}(\sqrt{D})$. When dealing with equation (7.4) we allow $z$ and $u$ to be negative.

### 7.2. The case $D=1$.

First we consider the special case $D=1$. Then (7.4) is equivalent to

$$
\left\{\begin{array}{l}
z+u=y_{1} \\
z-u=y_{2}
\end{array}\right.
$$

where $y=y_{1} \cdot y_{2}, \quad y_{1} \in S, \quad \pm y_{2} \in S$, and $y_{1}>\left|y_{2}\right|$. Subtraction yields

$$
\begin{equation*}
2 \cdot u=y_{1}-y_{2}, \tag{7.5}
\end{equation*}
$$

where now all variables $u, y_{1}, y_{2}$ (apart from the sign) are in $S$, hence in $\mathbb{Z}$. By $\left(u, y_{1}\right)=\left(u, y_{2}\right)=1$, equation (7.5) is of the form $a+b=c$, or $2 \cdot a+2 \cdot b=2 \cdot c$, where $a, b, c$ are composed of primes $2, p_{1}, \ldots, p_{s}$ only, and $(a, b)=1, a \geqslant b>0$. In Chapter 6 it was shown how to solve $a+b=c$. For our standard example $\left\{p_{1}, \ldots, p_{s}\right\}=\{2,3,5,7\}$ we have the following result.

LEMMA 7.3. Let $\left\{p_{1}, \ldots, p_{s}\right\}=\{2,3,5,7\}$. Equation (7.1) with conditions (7.2) and $\mathrm{D}=1$ has exactly the 95 solutions given in Table I with $\mathrm{D}=1$.

Proof. From Theorem 6.3 it follows that $a+b=c$ with $a, b, c \in S$, $(a, b)=1$, $a \geqslant b$ has exactly 63 solutions. They are easy to compute. Each of these gives rise to three possibilities for (7.5):

$$
\begin{aligned}
& \text { if } 2 \mid a \text { then }\left(u, y_{1}, y_{2}\right)=\left(\frac{1}{2} a, b, c\right),(b, 2 c, 2 a),(c, 2 a,-2 b), \\
& \text { if } 2 \mid b \text { then }\left(u, y_{1}, y_{2}\right)=(a, 2 b, 2 c),\left(\frac{1}{2} b, c, a\right),(c, 2 a,-2 b), \\
& \text { if } 2 \mid c \text { then }\left(u, y_{1}, y_{2}\right)=(a, 2 b, 2 c),(b, 2 c, 2 a),\left(\frac{1}{2} c, a,-b\right) .
\end{aligned}
$$

Of the thus found 189 possibilities, the 95 ones given in Table I with $\mathrm{D}=1$ satisfy $x \geqslant y$ and $z>0$, whereas the others don't.

This completes our treatment of the case $\mathrm{D}=1$.

### 7.3. Towards generalized recurrences.

From now on, let $D>1$. Put $K=\mathbb{Q}(\sqrt{D})$. Let $\sigma: K \rightarrow K$ be the automorphism of $K$ with $\sigma(\sqrt{D})=-\sqrt{D}$. For any number or ideal $X$ in $K$ we write $X$ for $\sigma(X)$, for convenience. Let $p_{i}$ for $i=1, \ldots$, $s$ be the prime ideal in $K$ such that $\operatorname{ord}_{p_{i}}\left(p_{i}\right)>0$. If $p_{i}$ splits in $\mathcal{O}_{K}$, this is well defined if a choice has been made from the two possibilities for $V D(\bmod p)$. Put for a solution $z, u, y$ of (7.4)

$$
\chi=z+u \cdot \sqrt{D} .
$$

Then $y=\chi \cdot \chi^{\prime}$, and by $(u, y)=1$ we have

$$
\begin{equation*}
\min \left(\operatorname{ord}_{p_{i}}(u), \operatorname{ord}_{p_{i}}(y)\right)=0 \tag{7.6}
\end{equation*}
$$

Equation (7.4) leads to the conjugated ideal equations

$$
\left\{\begin{array}{l}
(\chi)=\prod_{i=1}^{s} \mathfrak{p}_{i}^{a_{i}} \cdot \mathfrak{p}_{i}^{\prime}{ }^{b_{i}}  \tag{7.7}\\
\left(\chi^{\prime}\right)=\prod_{i=1}^{s} \mathfrak{p}_{i}^{\prime}{ }^{i} \cdot \mathfrak{p}_{i}
\end{array}\right.
$$

where $a_{i}, b_{i} \in \mathbb{N}_{0}$, and $b_{i}=0$ if $p_{i}=\mathfrak{p}_{i}$. We need the following auxiliary lemma.

LEMMA 7.4. If $\xi \in K$ and $\operatorname{ord}_{p}(\xi)=\operatorname{ord}_{p}\left(\xi^{\prime}\right)$ for a prime $p$, then

$$
\operatorname{ord}_{p}(\xi) \leqslant \operatorname{ord}_{p}\left(\xi-\xi^{\prime}\right)
$$

Moreover, if $\mathrm{p}=2$ and $\mathrm{D} \equiv 1(\bmod 8)$, then

$$
\operatorname{ord}_{2}(\xi) \leqslant \operatorname{ord}_{2}\left(\left(\xi-\xi^{\prime}\right) / 2\right),
$$

and, if $\mathrm{p}=2$ and $\mathrm{D} \equiv 2,3(\bmod 4)$, then

$$
\operatorname{ord}_{2}(\xi) \leqslant \operatorname{ord}_{2}\left(\left(\xi-\xi^{\prime}\right) / 2 \sqrt{ } D\right)+\frac{1}{2}
$$

Proof. This is an easy exercise, which we leave to the reader.

We distinguish, as usual, three cases for the factorization of the prime $p_{i}$ in $K$ : it may split, ramify or remain prime. See Borevich and Shafarevich [1966], section III. 8.
$\rightarrow p_{i}$ remains prime in $K$. Then $p_{i} \nmid D$, and if $p_{i}=2$ then $D \equiv 5(\bmod 8)$. We have $\left(p_{i}\right)=p_{i}=p_{i}^{\prime}$, and from $\operatorname{ord}_{p_{i}}(\chi)=\operatorname{ord}_{p_{i}}\left(\chi^{\prime}\right)$ and Lemma 7.4 we obtain

$$
\operatorname{ord}_{p_{i}}(y)=2 \cdot \operatorname{ord}_{p_{i}}(\chi) \leqslant 2 \cdot \operatorname{ord}_{p_{i}}\left(\chi-\chi^{\prime}\right)=2 \cdot \operatorname{ord}_{p_{i}}(2 \cdot u \cdot \sqrt{D})
$$

It follows, using (7.6), that

$$
\begin{aligned}
& \text { if } p_{i} \neq 2 \text { then } \operatorname{ord}_{p_{i}}(y)=2 \cdot a_{i}=0, \\
& \text { if } \quad p_{i}=2 \text { then } \operatorname{ord}_{2}(y)=2 \cdot a_{i}=0,2, \text { and if } a_{i}=1 \text { then } \\
& \quad \operatorname{ord}_{2}(u)=0 .
\end{aligned}
$$

$\rightarrow p_{i}$ ramifies in $K$. Then $p_{i} \mid D$ if $p_{i} \neq 2$, and $D \equiv 2,3(\bmod 4)$ if $p_{i}=2$. We have $\left(p_{i}\right)=p_{i}^{2}, p_{i}=p_{i}^{\prime}$, and $\operatorname{ord}_{p_{i}}(\chi)=\operatorname{ord}_{p_{i}}\left(\chi^{\prime}\right)=\frac{1}{2} \cdot a_{i}$.
From Lemma 7.4 we find

$$
\operatorname{ord}_{p_{i}}(y)=2 \cdot \operatorname{ord}_{p_{i}}(\chi) \leqslant 1+2 \cdot \operatorname{ord}_{p_{i}}\left(\left(\chi-\chi^{\prime}\right) / 2 \cdot \sqrt{ }\right)=1+2 \cdot \operatorname{ord}_{p_{i}}(u)
$$

By (7.6) we obtain

$$
\operatorname{ord}_{p_{i}}(y)=a_{i}=0,1, \text { and } i f \quad a_{i}=1 \text { then } \operatorname{ord}_{p_{i}}(u)=0
$$

$\rightarrow p_{i}$ splits in $K$. Then $p_{i} \nmid D$, and if $p_{i}=2$ then $D \equiv 1(\bmod 8)$.
We have $\left(p_{i}\right)=p_{i} \cdot p_{i}^{\prime}, p_{i} \neq p_{i}^{\prime}$. Further, $\quad \operatorname{ord}_{p_{i}}\left(p_{i}\right)=1, \quad$ ord $p_{i}\left(p_{i}^{\prime}\right)=0$.
Hence $\operatorname{ord}_{p_{i}}(\chi)=a_{i}, \quad \operatorname{ord}_{p_{i}}\left(\chi^{\prime}\right)=b_{i}$. If $a_{i}=b_{i}$ then from

$$
\operatorname{ord}_{p_{i}}(y)=2 \cdot \operatorname{ord}_{p_{i}}(\chi) \leqslant 2 \cdot \operatorname{ord}_{p_{i}}\left(\left(\chi-\chi^{\prime}\right) / 2\right)=2 \cdot \operatorname{ord}_{p_{i}}(u)
$$

we obtain by (7.6) that

$$
\operatorname{ord}_{p_{i}}(y)=a_{i}=b_{i}=0
$$

If $a_{i} \neq b_{i}$ then ord $p_{i}(y)=a_{i}+b_{i}>0$, hence ord $p_{i}(u)=0$, by (7.6).
We infer in this case

$$
\begin{aligned}
& \operatorname{ord}_{p_{i}}(y)=a_{i}+b_{i} \geqslant 1+2 \cdot \min \left(a_{i}, b_{i}\right)=1+2 \cdot \operatorname{ord}_{p_{i}}\left(\chi-\chi^{\prime}\right) \\
& \quad=1+2 \cdot \operatorname{ord}_{p_{i}}(2) .
\end{aligned}
$$

It follows that

$$
\begin{aligned}
& \operatorname{ord}_{p_{i}}(y)=\max \left(a_{i}, b_{i}\right), \quad \min \left(a_{i}, b_{i}\right)=0 \quad \text { if } \quad p_{i} \neq 2 \\
& \operatorname{ord}_{p_{i}}(y)=\max \left(a_{i}, b_{i}\right)+1, \quad \min \left(a_{i}, b_{i}\right)=1 \quad \text { if } \quad p_{i}=2
\end{aligned}
$$

Put $b_{0}=\min \left(a_{i}, b_{i}\right)$ if $p_{i}=2$ occurs, and $b_{0}=0$ otherwise. (Note that
$\min \left(a_{i}, b_{i}\right)=1$ may occur only if $p_{i} \neq p_{i}^{\prime}$, hence only if $p_{i}=2$ splits). Let us assume that the splitting primes of $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{s}}$ are $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{t}}$ for some $0 \leqslant t \leqslant s$. Put

$$
\begin{aligned}
& I=\left\{i \mid 1 \leqslant i \leqslant t, \quad a_{i}>b_{i}\right\} \\
& I^{\prime}=\left\{i \mid 1 \leqslant i \leqslant t, \quad a_{i}<b_{i}\right\}
\end{aligned}
$$

For $i=1, \ldots, t$, let $h_{i}$ be the smallest positive integer such that $p_{i}{ }_{i}$ is a principal ideal, say

$$
\mathfrak{p}_{i}^{h_{i}}=\left(\pi_{i}\right)
$$

If $h$ denotes the class number of $K$, then $h_{i} l h . N o w, \pi_{i} \in K$ is determined up to multiplication by a unit. Thus we may choose $\pi_{i}$ such that

$$
\begin{aligned}
& \left|\pi_{i}\right|>\left|\pi_{i}^{\prime}\right| \quad \text { if } \quad i \in I \\
& \left|\pi_{i}\right|<\left|\pi_{i}^{\prime}\right| \quad \text { if } \quad i \in I
\end{aligned}
$$

For $\mathrm{i}=1, \ldots, \mathrm{t}$, put

$$
\left|a_{i}-b_{i}\right|=c_{i} \cdot h_{i}+d_{i}
$$

with $c_{i}, d_{i} \in \mathbb{N}_{0}$, and $0 \leqslant d_{i} \leqslant h_{i}-1$. Consider the ideal

$$
a=(2)^{b_{0}} \cdot \prod_{i \in I} p_{i}^{d_{i}} \cdot \prod_{i \in I} p_{i}^{\prime}{ }^{d_{i}} \cdot \prod_{i=t+1}^{s} p_{i}^{a_{i}}
$$

From the above considerations it follows that, for given $K, p_{1}, \ldots, p_{s}$, there are only finitely many possibilities for $a$. By (7.7) it follows that

$$
(\chi)=a \cdot \prod_{i \in I}\left(\pi_{i}\right)^{c} i \quad \prod_{i \in I}\left(\pi_{i}^{\prime}\right)^{c} i
$$

(namely, $\quad\left|a_{i}-b_{i}\right|=\max \left(a_{i}, b_{i}\right) \quad$ if $p_{i} \neq 2$, since then $\min \left(a_{i}, b_{i}\right)=0$; and $\left|a_{i}-b_{i}\right|=\max \left(a_{i}, b_{i}\right)-1$ if $p_{i}=2$ and $b_{0}=1$ ). Hence $a$ is a principal ideal, say

$$
a=(\alpha)
$$

for an $\alpha \in \mathcal{O}_{\mathrm{K}}$. Up to multiplication by a unit, there are only finitely many possibilities for $\alpha$. Let $\varepsilon$ be the fundamental unit of $K$ with $\varepsilon>1$.

Now, (7.7) leads to the system of equations

$$
\left\{\begin{array}{l}
\chi=z+u \sqrt{ }= \pm \alpha \cdot \varepsilon^{n} \cdot \prod_{i \in I} \pi_{i}^{c_{i}} \cdot \prod_{i \in I} \pi^{\prime}{ }^{C_{i}}  \tag{7.8}\\
\chi^{\prime}=z-u \sqrt{ }= \pm \alpha^{\prime} \cdot \varepsilon^{\prime}{ }^{n} \cdot \prod_{i \in I} \pi_{i}^{\prime}{ }^{c} i \cdot \prod_{i \in I},^{c_{i}}{ }_{i}
\end{array}\right.
$$

where $n \in \mathbb{Z}$. Put for $n \in \mathbb{Z}, m_{1}, \ldots, m_{t} \in \mathbb{N}_{0}$, and for each possible $\alpha$

$$
\begin{aligned}
& G_{\alpha}\left(n, m_{1}, \ldots, m_{t}\right)=\frac{\alpha}{2 \sqrt{D}} \cdot \varepsilon^{n} \cdot \prod_{i \in I} \pi_{i}^{m_{i}} \cdot \prod_{i \in I} \pi^{\prime}{ }^{m_{i}}-\frac{\alpha^{\prime}}{2 \sqrt{D}} \cdot \varepsilon^{, n} \cdot \prod_{i \in I} \pi_{i}^{\prime}{ }^{m_{i}} \cdot \prod_{i \in I},{ }^{m_{i}}{ }^{i}, \\
& H_{\alpha}\left(n, m_{1}, \ldots, m_{t}\right)=\frac{\alpha}{2} \cdot \varepsilon^{n} \cdot \prod_{i \in I} \pi_{i}^{m_{i}} \cdot \prod_{i \in I} \pi_{i}^{\prime}{ }^{m_{i}}+\frac{\alpha}{2} \cdot{ }^{\prime},{ }^{\prime} n \cdot \prod_{i \in I} \pi_{i}^{\prime}{ }^{m_{i}} \cdot \prod_{i \in I},^{m_{i}}{ }^{i} .
\end{aligned}
$$

Then (7.8) is equivalent to

$$
\left\{\begin{array}{l} 
\pm u=G_{\alpha}\left(n, c_{1}, \ldots, c_{t}\right)  \tag{7.9}\\
\pm z=H_{\alpha}\left(n, c_{1}, \ldots, c_{t}\right)
\end{array}\right.
$$

The functions $G_{\alpha}$ and $H_{\alpha}$ are generalized recurrences in the sense that if all variables but one are fixed, then they become integral binary recurrence sequences. We show an example in Fig. 8.


Figure 8. $G_{\alpha}(n, m)=\frac{\alpha}{2 \sqrt{D}} \cdot \varepsilon^{n} \cdot \pi^{m}-\frac{\alpha^{\prime}}{2 \sqrt{D}} \cdot \varepsilon^{, n} \cdot \pi^{, m}$ for $D=30, \alpha=5+\sqrt{ } 30$, $\varepsilon=11+2 \cdot \sqrt{ } 30, \pi=13+2 \cdot \sqrt{ } 30$, with $-10 \leqslant n \leqslant 10$ (vertically) and $0 \leqslant m \leqslant 10$ (horizontally). Numbers $\geqslant 10^{12}$ are denoted by asterisks.

### 7.4. Towards linear forms in logarithms.

Let us write $u_{i}=\operatorname{ord}_{p_{i}}(u)$ for $i=1, \ldots, s$. Put for each $\alpha$

$$
\begin{aligned}
I_{U}= & \left\langle i \quad 1 \leqslant i \leqslant s, \operatorname{ord}_{p_{i}}\left(G_{\alpha}\left(n, m_{1}, \ldots, m_{t}\right)\right)\right\rangle 0 \text { occurs } \\
& \text { for at least one } \left.\left(n, m_{1}, \ldots, m_{t}\right) \in \mathbb{Z} \times \mathbb{N}_{0}^{t}\right\rangle
\end{aligned}
$$

Note that since $(u, y)=1$ the sets $I_{U}$, I, I, are disjunct. We proceed with the first equation of system (7.9). Written out in full detail it reads

$$
\begin{equation*}
\frac{\alpha}{2 \sqrt{D}} \cdot \varepsilon^{n} \cdot \prod_{i \in I} \pi_{i}^{c}{ }_{i} \cdot \prod_{i \in I} \pi_{i}^{\prime}{ }^{c} i \quad-\frac{\alpha}{2 \sqrt{D}} \cdot \varepsilon^{, n} \cdot \prod_{i \in I} \pi_{i}^{\prime}{ }^{c} \cdot \prod_{i \in I} \pi_{i}^{c}{ }_{i}= \pm \prod_{i \in I} p_{i}^{u_{i}} \tag{7.10}
\end{equation*}
$$

Now, I, I', $I_{U}$ depend on $\alpha$, which depends on the particular solution of equation (7.4) that we presupposed. However, we know that $\alpha$ belongs to a finite set, which can be computed explicitly. So if we can solve (7.10) completely for each $\alpha$ of this set, then we can find all solutions of (7.9), hence of (7.1).

The set of the $\alpha$ 's may be reduced, without loss of generality, as follows. If $D \equiv 1(\bmod 8)$ then $b_{0}=0,1$ may both occur, with $\alpha=\alpha_{0}, 2 \cdot \alpha_{0}$ respectively. We only have to consider $2 \cdot \alpha_{0}$, because if $u=u_{0}, z=z_{0}$ is a solution of (7.9) for $\alpha=\alpha_{0}$, then $u=2 \cdot u_{0}, z=2 \cdot z_{0}$ is a solution of (7.9) for $\alpha=2 \cdot \alpha_{0}$. Hence it is not necessary to consider $\alpha=\alpha_{0}$ if also $\alpha=2 \cdot \alpha_{0}$ is already being considered. By the same argument, if $D \equiv 5(\bmod 8)$ then with $\alpha=\alpha_{0}$ such that $\operatorname{ord}_{2}\left(\alpha_{0}\right)=0$ also $\alpha=2 \cdot \alpha_{0}$ may occur, so that we only have to consider the latter. Note that it may now occur that $(u, y)=2$. The condition $(u, y)=1$ is used only to ensure that $I_{U}$ and I U I, are disjunct. This remains true in the above cases with $(u, y)=2$. Further, if $\left(\alpha_{0}\right) \neq\left(\alpha_{0}^{\prime}\right)$ for some $\alpha_{0}$, then we only have to consider one $\alpha$ of the pair $\alpha_{0}, \alpha_{0}$. Namely, if the I, I' belonging to $\alpha_{0}$ are $I_{0}, I_{0}^{\prime}$, then the $I, ~ I$, belonging to $\alpha_{0}^{\prime}$ are $I_{0}^{\prime}, I_{0}$, and then

$$
\begin{aligned}
& G_{\alpha_{0}}\left(n, m_{1}, \ldots, m_{t}\right)=\frac{\alpha_{0}^{\prime}}{2 \sqrt{D}} \cdot \varepsilon^{n} \cdot \prod_{I_{0}} \pi_{i}{ }^{c}{ }^{i} \cdot \prod_{I_{0}} \pi_{i}{ }^{c_{i}}-\frac{\alpha_{0}}{2 \sqrt{D}} \cdot \varepsilon^{, n} \cdot \prod_{I_{0}},{ }^{\prime}{ }_{i}{ }^{c_{i}} \cdot \prod_{I_{0}} \pi_{i}{ }_{i}
\end{aligned}
$$

$$
\begin{aligned}
& =\mp G_{\alpha_{0}}\left(-n, m_{1}, \ldots, m_{t}\right),
\end{aligned}
$$

(by using $\varepsilon \cdot \varepsilon^{\prime}= \pm 1$ ), and analogously

$$
H_{\alpha_{0}^{\prime}}\left(n, m_{1}, \ldots, m_{t}\right)= \pm H_{\alpha_{0}}\left(-n, m_{1}, \ldots, m_{t}\right)
$$

From equation (7.10) we now derive $p_{i}$-adic linear forms in logarithms, in three different ways, according to $i \in I$, $I$, or $I_{U}$. Put

$$
\gamma_{i}=\frac{3}{2} \text { if } p_{i}=2, \quad \gamma_{i}=1 \text { if } p_{i}=3, \quad \gamma_{i}=\frac{1}{2} \text { if } p_{i} \geqslant 5 .
$$

Then $\gamma_{i}>1 /\left(p_{i}-1\right)$, hence if $\operatorname{ord}_{p_{i}}(\xi) \geqslant \gamma_{i}$ for a $\xi \in K$ then

$$
\begin{equation*}
\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}(1 \pm \xi)\right)=\operatorname{ord}_{p_{i}}(\xi) \tag{7.11}
\end{equation*}
$$

We now have the following result.

LEMMA 7.5. Let $n, c_{i}\left(i \in I U I\right.$ ) , $u_{i}\left(i \in I_{U}\right)$ satisfy (7.10).
(i). For $\mathrm{i} \in \mathrm{I}_{\mathrm{U}}$ put

$$
\begin{aligned}
\lambda_{i}= & \operatorname{ord}_{p_{i}}\left(2 \sqrt{ } / \alpha^{\prime}\right) \\
\Lambda_{i}= & \log _{p_{i}}\left(\frac{\alpha}{\alpha^{\prime}}\right)+n \cdot \log _{p_{i}}\left(\frac{\varepsilon}{\varepsilon^{\prime}}\right)+\sum_{j \in I} c_{j} \cdot \log _{p_{i}}\left(\frac{\pi}{\pi_{j}^{\prime}}\right) \\
& -\sum_{j \in I},{ }^{c}{ }_{j} \cdot \log _{p_{i}}\left(\frac{\pi}{\pi_{j}^{\prime}}\right)
\end{aligned}
$$

If $u_{i}+\lambda_{i} \geqslant \gamma_{i}$ then

$$
u_{i}+\lambda_{i}=\operatorname{ord}_{p_{i}}\left(\Lambda_{i}\right)
$$

(ii). For $\mathrm{i} \in \mathrm{I}$ put

$$
\begin{aligned}
\kappa_{i}= & \operatorname{ord}_{p_{i}}\left(\frac{\alpha}{\alpha^{\prime}}\right), \\
K_{i}= & \log _{p_{i}}\left(\frac{\alpha^{\prime}}{2 \sqrt{D}}\right)+n \cdot \log _{p_{i}}\left(\varepsilon^{\prime}\right)-\sum_{j \in I_{U}} u_{j} \cdot \log _{p_{i}}\left(p_{j}\right) \\
& +\sum_{j \in I} c_{j} \cdot \log _{p_{i}}\left(\pi_{j}^{\prime}\right)+\sum_{j \in I},{ }^{c} \cdot{ }_{j} \log _{p_{i}}\left(\pi{ }_{j}\right)
\end{aligned}
$$

If $\mathrm{h}_{\mathrm{i}} \cdot \mathrm{C}_{\mathrm{i}}+\kappa_{\mathrm{i}} \geqslant \gamma_{\mathrm{i}}$ then

$$
h_{i} \cdot c_{i}+\kappa_{i}=\operatorname{ord}_{p_{i}}\left(K_{i}\right)
$$

(ii'). For i $\in$ I' put

$$
\kappa_{i}^{\prime}=\operatorname{ord}_{p_{i}}\left(\frac{\alpha^{\prime}}{\alpha}\right)
$$

$$
\begin{aligned}
K_{i}^{\prime}= & \log _{p_{i}}\left(\frac{\alpha}{2 \sqrt{D}}\right)+n \cdot \log _{p_{i}}(\varepsilon)-\sum_{j \in I_{U}} u_{j} \cdot \log _{p_{i}}\left(p_{j}\right) \\
& +\sum_{j \in I} c_{j} \cdot \log _{p_{i}}\left(\pi \pi_{j}\right)+\sum_{j \in I},{ }^{c}{ }_{j} \cdot \log _{p_{i}}\left(\pi_{j}^{\prime}\right)
\end{aligned}
$$

If $h_{i} \cdot c_{i}+\kappa_{i}^{\prime} \geqslant \gamma_{i}$ then

$$
h_{i} \cdot c_{i}+\kappa_{i}^{\prime}=\operatorname{ord}_{p_{i}}\left(K_{i}^{\prime}\right)
$$

Remark. Note that all the above $p_{i}$-adic logarithms are well-defined, since their arguments have $p_{i}$-adic order zero. This follows from the fact that $I_{U}$, I and $I$, are disjunct, and if $D \equiv 1(\bmod 8)$ from the choice $\alpha=2 \cdot \alpha_{0}$.

Proof. For (i), divide (7.10) by its second term. For (ii), divide (7.10) by its second term, and add 1. For (ii'), divide (7.10) by its first term, and add -1 . Then in all three cases take the $p_{i}$-adic order, and apply (7.11). a

The linear forms in logarithms $\Lambda_{i}, K_{i}, K_{i}^{\prime}$, as they appear in Lemma 7.5, seem to be inhomogeneous, since the first term has coefficient 1. However, it can be made homogeneous by incorporating this first term in the other ones, as follows. Put

$$
h^{*}=\operatorname{lcm}\left(2, h_{1}, \ldots, h_{s}\right)
$$

Note that, by the definition of $\alpha$,

$$
\begin{equation*}
\alpha^{h^{*}}= \pm \varepsilon^{n_{0}} \cdot \prod_{i \in I} \pi_{i}^{n} \cdot \prod_{i \in I} \pi_{i}^{n}{ }^{n} \cdot \prod_{i=t+1}^{s} p_{i}^{n_{i}} \cdot 2^{h^{*} \cdot b_{0}} \tag{7.12}
\end{equation*}
$$

where the exponents $n_{i}$ for $0 \leqslant i \leqslant s$ are integral. It follows that

$$
\left(\frac{\alpha}{\alpha^{\prime}}\right)^{h^{*}}= \pm\left(\frac{\varepsilon}{\varepsilon^{\prime}}\right)^{n_{0}} \cdot \prod_{i \in I}\left(\frac{\pi}{\pi^{\prime}}\right)^{n} \mathrm{i} \cdot \prod_{i \in I}\left(\frac{\pi^{\prime}}{\pi}\right)^{n_{i}}
$$

Put

$$
\Lambda_{i}^{*}=h^{*} \cdot \Lambda_{i}, \quad n^{*}=h^{*} \cdot n+n_{0}, \quad c_{j}^{*}=h^{*} \cdot c_{j}+n_{j}
$$

Then it follows that

$$
\Lambda_{i}^{*}=n^{*} \cdot \log _{p_{i}}\left(\frac{\varepsilon}{\varepsilon^{\prime}}\right)+\sum_{j \in I} c_{j}^{*} \cdot \log _{p_{i}}\left(\frac{\pi}{\pi_{j}^{\prime}}\right)-\sum_{j \in I} c^{*} \cdot \log _{p_{i}}\left(\frac{\pi_{j}}{\pi_{j}^{\prime}}\right)
$$

Note that the prime divisors of $D$ are just the ramifying primes. By (7.12),

$$
\left(\frac{\alpha}{2 \sqrt{D}}\right)^{h^{*}}= \pm \varepsilon^{n_{0}} \cdot \prod_{i \in I} \pi_{i}^{n} \cdot \prod_{i \in I} \pi_{i}^{\prime}{ }^{n}{ }_{i} \cdot \prod_{i=t+1}^{s} p_{i}^{n_{i}}{ }^{-v} \cdot 2^{h^{*} \cdot\left(b_{0}-v_{0}\right)}
$$

where $v_{i}=\frac{1}{2} \cdot h^{*} \cdot \operatorname{ord}_{p_{i}}(4 D) \in \mathbb{Z}$ for $i=t+1, \ldots, s$, and $v_{0}=1$ if 2 splits, $v_{0}=0$ otherwise. If $p_{i}=2$ splits we have assumed that $b_{0}=1$. Hence the last factor vanishes. So put

$$
\begin{aligned}
& K_{i}^{*}=h^{*} \cdot K_{i}, K_{i}^{\prime *}=h^{*} \cdot K_{i}^{\prime}, \quad u_{j}^{*}=h^{*} \cdot u_{j}-\left(n_{j}-v_{j}\right), \\
& I_{U}^{*}=I_{U} U\left\langle i \mid t+1 \leqslant i \leqslant s, \quad v_{i} \neq 0\right\rangle .
\end{aligned}
$$

Then it follows that

$$
\begin{aligned}
K_{i}^{*}= & n^{*} \cdot \log _{p_{i}}\left(\varepsilon^{\prime}\right)-\sum_{j \in I_{U}^{*}} u_{j}^{*} \cdot \log _{p_{i}}\left(p_{j}\right)+\sum_{j \in I} c_{j}^{*} \cdot \log _{p_{i}}\left(\pi_{j}^{\prime}\right)+ \\
& +\sum_{j \in I},^{*} \cdot{ }_{j}^{*} \cdot \log _{p_{i}}\left(\pi{ }_{j}\right), \\
K_{i}^{*}= & n^{*} \cdot \log _{p_{i}}(\varepsilon)-\sum_{j \in I_{U}} u^{*}{ }_{j}^{*} \cdot \log _{p_{i}}\left(p_{j}\right)+\sum_{j \in I} c_{j}^{*} \cdot \log _{p_{i}}\left(\pi{ }_{j}\right)+ \\
& +\sum_{j \in I},{ }^{*}{ }_{j}^{*} \cdot \log _{p_{i}}\left(\pi_{j}^{\prime}\right) .
\end{aligned}
$$

This leads to the following reformulation of Lemma 7.5.

LEMMA 7.6. Let $n, c_{i}$ for $i \in I \cup I, u_{i}$ for $i \in I_{U}$ be a solution of (7.10), let $\lambda_{i}, \kappa_{i}, \kappa_{i}^{\prime}$ be as in Lemma 7.5, and let $h^{*}, \Lambda_{i}^{*}, K_{i}^{*}, K_{i}^{\prime}, n_{i}^{*}$, $c_{i}^{*}, u_{i}^{*}, I_{U}^{*}$ be as above.
(i). Let $i \in I_{U}$. If $u_{i}+\lambda_{i} \geqslant \gamma_{i}$ then

$$
u_{i}+\lambda_{i}+\operatorname{ord}_{p_{i}}\left(h^{*}\right)=\operatorname{ord}_{p_{i}}\left(\Lambda_{i}^{*}\right)
$$

(ii). Let $\mathrm{i} \in \mathrm{I}$. If $\mathrm{h}_{\mathrm{i}} \cdot \mathrm{c}_{\mathrm{i}}+\kappa_{\mathrm{i}} \geqslant \gamma_{\mathrm{i}}$ then

$$
h_{i} \cdot c_{i}+\kappa_{i}+\operatorname{ord}_{p_{i}}\left(h^{*}\right)=\operatorname{ord}_{p_{i}}\left(K_{i}^{*}\right)
$$

(ii'). Let $\mathrm{i} \in \mathrm{I}$, If $\mathrm{h}_{\mathrm{i}} \cdot \mathrm{c}_{\mathrm{i}}+\kappa_{\mathrm{i}}^{\prime} \geqslant \gamma_{\mathrm{i}}$ then

$$
h_{i} \cdot c_{i}+\kappa_{i}^{\prime}+\operatorname{ord}_{p_{i}}\left(h^{*}\right)=\operatorname{ord}_{p_{i}}\left(K_{i}^{*}\right)
$$

Remark. We will study the linear forms in logarithms $\Lambda_{i}^{*}, K_{i}^{*}, K_{i}^{*}$ for arbitrary integral values of the variables $n^{*}, c_{i}^{*}, u_{i}^{*}$. Notice that the parameter $\alpha$ has disappeared completely from these linear forms. This means that we have to consider the linear forms for each $D$ only, instead of for each $\alpha$.

### 7.5. Upper bounds for the solutions: outline.

Let us first give a global explanation of our application of the theory of p-adic linear forms in logarithms, that gives explicit upper bounds for the variables occurring in the linear forms $\Lambda_{i}^{*}, K_{i}^{*}, K_{i}^{*}$. Then we give arguments why we choose this way to apply the theory, and not other possible ways. In the next section we give full details of the derivation of the upper bounds. In the sequel, by the 'constants' $C_{1}, \ldots, C_{12}$ we mean numbers that depend only on the parameters of (7.10), not on the unknowns $n, c_{i}, u_{i}$.

Put

$$
\begin{aligned}
& M=\max _{i \in I U I}\left(c_{i}\right), \quad U=\max _{i \in I_{U}}\left(u_{i}\right), B=\max (M, U,|n|), \\
& M^{*}=\max _{i \in I U I}\left(c_{i}^{*}\right), \quad U^{*}=\max _{i \in I_{U}}\left(u_{i}^{*}\right), B^{*}=\max \left(M^{*}, U^{*},\left|n^{*}\right|\right), \\
& N=\max \left(\left|n_{0}\right|, \ldots,\left|n_{t}\right|,\left|n_{t+1}-v_{t+1}\right|, \ldots,\left|n_{s}-v_{s}\right|\right) .
\end{aligned}
$$

Then it follows that

$$
\begin{equation*}
X^{*} \leqslant h^{*} \cdot X+N, \quad X \leqslant \frac{X^{*}+N}{h^{*}} \tag{7.13}
\end{equation*}
$$

for $X=M, U, B$. We apply Lemma 2.6 to the p-adic linear forms in logarithms. For $\Lambda_{i}^{*}$ we find, in view of Lemma 7.6(i),

$$
\begin{equation*}
\mathrm{U}<\mathrm{C}_{1}+\mathrm{C}_{2} \cdot \log \left(\mathrm{~B}^{*}\right) \tag{7.14}
\end{equation*}
$$

and for $K_{i}^{*}, K_{i}^{*}$ we find, in view of Lemma 7.6(ii), (ii'),

$$
\begin{equation*}
M<C_{3}+C_{4} \cdot \log \left(B^{*}\right) \tag{7.15}
\end{equation*}
$$

Here, $C_{1}, C_{2}, C_{3}, C_{4}$ are constants that can be written down explicitly. In order to find an upper bound for $B$ we try to find $C_{10}, C_{11}$ such that

$$
\begin{equation*}
B<C_{10}+C_{11} \cdot \log \left(B^{*}\right) \tag{7.16}
\end{equation*}
$$

In view of (7.13) we may insert and delete asterisks any time we like, as long as we don't specify the constants. In order to prove (7.16) it remains, in view of (7.14) and (7.15), to bound $|n|$ by a constant times $\log B$. We will introduce certain constants $C_{5}, C_{6}, C_{7}$, and distinguish three cases:
(a). $-\left(C_{6}+C_{7} \cdot M\right) \leqslant n \leqslant C_{5}$,
(b). $n>C_{5}$,
(c). $n<-\left(C_{6}+C_{7} \cdot M\right)$.

In case (a) it is, by (7.15), obvious that (7.16) holds. In cases (b) and (c) one of the two terms of $G_{\alpha}$ dominates. We shall show that there exist constants $\mathrm{C}_{8}, \mathrm{C}_{9}$ such that

$$
\begin{equation*}
|n|<C_{8}+C_{9} \cdot U \tag{7.18}
\end{equation*}
$$

Then (7.16) follows from (7.14).

From (7.16) we derive immediately an explicit upper bound $C_{12}$ for $B$, hence for all the variables involved. Since the constants $C_{1}, \ldots, C_{4}$ will be very large, also $C_{12}$ will be very large. To find all solutions we proceed by reducing this upper bound, by applying the computational p-adic diophantine approximation technique described in Section 3.11, to the p-adic linear forms in logarithms $\Lambda_{i}^{*}, K_{i}^{*}, K_{i}^{*}$. Crucial in that line of argument is that the constants $C_{5}, \ldots, C_{9}$ are very small compared to $C_{1}, \ldots, C_{4}$. This method leads to reduced bounds for the $p$-adic orders of the linear forms. Then we can replace (7.14) and (7.15) by much sharper inequalities, and repeat the above argument, to find a much sharper inequality for (7.16). In general we expect that it is in this way possible to reduce in one step the upper bound $C_{12}$ for $B$ to a reduced bound of size $\log C_{12}$.

Before going into detail we explain briefly that it is possible to treat (7.10) partly by the theory of real (instead of p-adic) linear forms in logarithms, and subsequently by a real computational diophantine approximation technique (cf. Section 3.7), and why we prefer not to do so. First, note that $K_{i}$ and $K_{i}^{\prime}$ have generically more terms than $\Lambda_{i}$, and are therefore more complicated to handle. Since $K_{i}, K_{i}^{\prime}$ occur only in case (a), this is the most difficult case. Equation (7.10) consist of three terms, each of which is purely exponential, i.e. the bases are fixed and the exponents are variable. If one of these three terms is essentially smaller than the
other two (more specifically, smaller than the other terms raised to the power $\delta$, for a fixed $\delta \in(0,1)$ ), then we can apply the real method. There are two ways of doing this. Write (7.10) as

$$
\chi-\chi^{\prime}=2 \cdot u \cdot \sqrt{ } D .
$$

First, suppose that $\left|\chi-\chi^{\prime}\right|<\left|\chi^{\prime}\right|^{\delta}$. Then $|n|$ cannot be very large, and we are essentially (i.e. apart from a finite domain) in case (a). Unfortunately, the region for $|n|$ that we can cover in this way becomes smaller as $M \rightarrow \infty$ (see the example below). Second, suppose that $|\chi|>\left|\chi^{\prime}\right|^{1 / \delta}$, or $|\chi|<\left|\chi^{\prime}\right|^{\delta}$. Then we are essentially in case (b) or (c). But this area can be dealt with easier p-adically, since here we use the linear forms $\Lambda_{i}$, whereas the real linear forms in logarithms used in this case will generically have more terms. The areas sketched above, in which we can apply the real theory, will not cover the whole domain corresponding to case (a) (cf. the white regions in Fig. 9 below). Hence we cannot avoid working with the $p$-adic linear forms $K_{i}, K_{i}$. But then it is more convenient to avoid the use of real linear forms.


Figure 9.

Let us illustrate the above reasoning with an example. Let $\alpha=\alpha^{\prime}=1$, $\varepsilon=1+\sqrt{2}, \pi_{1}=1+2 \cdot \sqrt{2}, s=1, I=\{1\}, p_{1}=7, I \prime=\varnothing$, and $\delta=\frac{1}{2}$. Then we have $\chi=(1+\sqrt{2})^{n} \cdot(1+2 \cdot \sqrt{2})^{M}$. Fig. 9 above gives in the $(n, M)-p l a n e$ the curves $\chi=\chi^{\prime 2}, 2 \cdot\left|\chi^{\prime}\right|,\left|\chi^{\prime}\right|+\sqrt{ }\left|\chi^{\prime}\right|,\left|\chi^{\prime}\right|,\left|\chi^{\prime}\right|-\sqrt{ }\left|\chi^{\prime}\right|, \frac{1}{2} \cdot\left|\chi^{\prime}\right|, \sqrt{ }\left|\chi^{\prime}\right|$, which are boundaries of the four regions $A, B, C, D$. We have the following possibilities.

|  | case | number of terms in linear form |  |
| :---: | :---: | :---: | :---: |
| region | (essentially) | p-adic method | real method |
| A | (b), (c) | 2 | 3 |
| B | (b), (c) | 2 | - |
| C | (a) | 3 | - |
| D | (a) | 3 | 2 |

The hardest part is $C$. It can be reduced to $\frac{1}{c} \cdot\left|\chi^{\prime}\right|<\chi<\left|\chi^{\prime}\right|-\left|\chi^{\prime}\right|^{\delta}$ and $\left|\chi^{\prime}\right|+\left|\chi^{\prime}\right|^{\delta}<\chi<c \cdot\left|\chi^{\prime}\right|$ for any $c>1, \delta \in(0,1)$, but will never disappear. So we cannot avoid the p-adic linear form in case (a), which then works in regions C and D together.

### 7.6. Upper bounds for the solutions: details.

We now proceed with filling in the details of the procedure outlined in the previous section.

We apply Yu's lemma (Lemma 2.6) as follows. We have $L=K=\mathbb{Q}(\sqrt{ } D)$, so $\mathrm{d}=2$. For the $\alpha_{i}$ we have $\varepsilon / \varepsilon^{\prime}, \pi_{j} / \pi_{j}^{\prime}$, or $\varepsilon, \varepsilon^{\prime}, p_{j}, \pi_{j}, \pi_{j}$. We have to compute the heights of these numbers. We have at once

$$
\begin{aligned}
& h\left(p_{j}\right)=\log \left(p_{j}\right) \text { if } p_{j} \geqslant 3, h(2)=1, \\
& h(\varepsilon)=h\left(\varepsilon^{\prime}\right)=\frac{1}{2} \cdot \log (\varepsilon), \\
& h\left(\pi_{j}\right)=h\left(\pi_{j}^{\prime}\right)=\frac{1}{2} \cdot \log \left(\max \left(1,\left|\pi_{j}\right|\right) \cdot \max \left(1,\left|\pi_{j}^{\prime}\right|\right)\right) .
\end{aligned}
$$

Further, let $\beta=\varepsilon$ or $\beta=\pi_{j}$. Then the leading coefficient of $\beta / \beta$, is $a_{0}=\left|\beta \cdot \beta^{\prime}\right|$, and we infer

$$
h\left(\frac{\beta}{\beta^{\prime}}\right)=\frac{1}{2} \log \left(\left|\beta \cdot \beta^{\prime}\right| \cdot \max \left(1,\left|\frac{\beta}{\beta^{\prime}}\right|\right) \cdot \max \left(1,\left|\frac{\beta^{\prime}}{\beta}\right|\right)\right)=\log \left(\max \left(|\beta|,\left|\beta^{\prime}\right|\right)\right)
$$

Hence

$$
h\left(\frac{\varepsilon}{\varepsilon^{\prime}}\right)=\log (\varepsilon), \quad h\left(\frac{\pi_{j}}{\pi_{j}^{\prime}}\right)=\log \left(\max \left(\left|\pi_{j}\right|,\left|\pi_{j}^{\prime}\right|\right)\right)
$$

The order of the $\alpha_{i}$ is important in two respects: it is required that the $V_{i}$ for $i=1, \ldots, n-1$ are in increasing order, and that ord $\left(b_{n}\right)$ is minimal among the $\operatorname{ord}_{p}\left(b_{i}\right)$. Since the $b_{i}$ are the unknowns, we should assume that $V_{n} \leqslant V_{1} \leqslant \ldots \leqslant V_{n-1}$. In the final bound however, only the product $V_{1} \cdot \ldots \cdot V_{n}$ and $V_{n-1}^{+}$appear. So the ordering of the $V_{i}$ only matters for defining $V_{n-1}^{+}$. It follows that we can take

$$
V_{i}=\max \left(h\left(\alpha_{i}\right), f_{p} \cdot(\log p) / d\right)
$$

with the $\alpha_{i}$ in any order, if we define

$$
\mathrm{V}_{\mathrm{n}-1}^{+}=\max \left(1, \mathrm{~V}_{1}, \ldots, \mathrm{~V}_{\mathrm{n}}\right)
$$

Further, we take

$$
B=B_{0}=B_{n}=B^{\prime}=\max \left(\left|b_{1}\right|, \ldots,\left|b_{n}\right|, 2, \frac{4}{3} \cdot n \cdot\left(p^{f_{p} / d}-1\right)\right)
$$

Then $\log \left(1+\frac{3}{4 n} \cdot B\right) \geqslant f_{p} \cdot(\log p) / d$. By $B \geqslant 2$ it follows that $1+\frac{3}{4 n} \cdot B<B$. Hence we can take

$$
\mathrm{W}=\log \mathrm{B}
$$

There are two more conditions to be checked. The first one is that $\alpha_{1}{ }^{b_{1}} \ldots \cdot \alpha_{n}{ }_{n} \neq 1$. This is immediate, if we assume the obvious condition that not all $b_{i}$ are zero. The second one is $\left[K\left(\alpha_{1}^{1 / q}, \ldots, \alpha_{n}^{1 / q}\right): K\right]=q^{n}$, which is less obvious. For our situation it follows from the following lemma. Application of Yu's newest results avoids such a condition (cf. Yu [1989]). Nevertheless we include the lemma here, to show that it is often possible to prove such a condition, which may in some cases lead to lower constants.

LEMMA 7.7. Let $K=\mathbb{Q}(\sqrt{ } D)$, with $\varepsilon$ as fundamental unit, and $h$ as class number. Let $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{s}}$ be distinct prime numbers, and let $\mathfrak{p}_{\mathrm{i}}$ be for $\mathrm{i}=1$, ..., s a prime ideal in K lying above $\mathrm{p}_{\mathrm{i}}$. Let $\mathrm{h}_{\mathrm{i}}$ be a divisor of $h$ such that $p_{i}$ is principal, and denote a generator by $\pi_{i}$. Let either: (1) all $\mathrm{p}_{\mathrm{i}}$ split, and then

$$
\xi_{0}=\frac{\varepsilon}{\varepsilon^{\prime}}, \quad \xi_{j}=\frac{\pi_{j}}{\pi_{j}^{\prime}} \text { for } i=1, \ldots, s
$$

or: (2)

$$
\xi_{0}=\varepsilon \text { or } \varepsilon^{\prime}, \quad \xi_{j}=\pi_{j} \text { or } \pi_{j}^{\prime} \text { for } j=1, \ldots, s
$$

Let q be an odd prime, not dividing h . Then

$$
\left[K\left(\xi_{0}^{1 / q}, \ldots, \xi_{S}^{1 / q}\right): K\right]=q^{s+1}
$$

Proof. Let $K_{0}=K\left(\xi_{0}^{1 / q}\right)$, and $K_{i}=K_{i-1}\left(\xi_{i}^{1 / q}\right)$ for $i=1$, .., $s$. We use induction on $i$ to prove that $\left[K_{S}: K\right]=q^{s+1}$. Note that $\left[K_{0}: K\right]=q$. Suppose that $\left[K_{i}: K\right]=q^{i+1}$. It remains to prove that $\left[K_{i+1}: K_{i}\right]=q$, hence it suffices to prove that $\xi_{i+1} \not \not \mathrm{~K}_{\mathrm{i}}$, since $q$ is prime. Suppose the contrary is true. $K_{i}$ is a K-vector space of dimension $q^{i+1}$, with as basis all the elements

$$
\tau_{k_{0}, \ldots, k_{i}}=\prod_{j=0}^{i} \xi_{j} j_{j} / q
$$

for $k_{j} \in\{0,1, \ldots, q-1\}$ for $j=0, \ldots, i$. It follows that there exist $a_{k_{0}}, \ldots, k_{i} \in K$ such that

$$
\begin{equation*}
\xi_{i+1}^{1 / q}=\sum_{k_{0}, \ldots, k_{i}} a_{k_{0}}, \ldots, k_{i} \cdot \tau_{k_{0}}, \ldots, k_{i} \tag{7.19}
\end{equation*}
$$

The group of $K$-embeddings of $K_{i}$ into $\mathbb{C}$ is generated by the $\sigma_{j}$ for $j=0, \ldots, i \quad$ defined by

$$
\begin{aligned}
& \sigma_{j}\left(\xi_{\ell}^{1 / q}\right)=\xi_{\ell}^{1 / q} \text { for } \ell=0, \ldots, i, \quad \ell \neq j, \\
& \sigma_{j}\left(\xi_{j}^{1 / q}\right)=\rho \cdot \xi_{j}^{1 / q},
\end{aligned}
$$

where $\rho$ is a primitive $q$ th root of unity. Hence all the embeddings are given by

$$
\varphi_{\ell_{0}}, \ldots \ell_{i}=\prod_{j=0}^{i} \sigma_{j}^{\ell}
$$

for $\ell_{j} \in\{0,1, \ldots, q-1\}$. It follows that

$$
\varphi_{\ell_{0}}, \ldots, \ell_{i}\left(\tau_{k_{0}}, \ldots, k_{i}\right)=\prod_{j=0}^{i} \sigma_{j}^{\ell}\left(\prod_{m=0}^{i} \xi_{m}^{k_{m}^{\prime q}}\right)=\prod_{j=0}^{i} \rho^{\ell}{ }^{k}{ }^{j} \cdot \tau_{k_{0}}, \ldots, k_{i}
$$

$$
=\rho^{\sum_{j=0}^{i} \ell j^{k} j} \cdot \tau_{k_{0}}, \ldots, k_{i}
$$

The minimal polynomial of $\xi_{i+1}^{1 / q}$ over $K$ is $X^{q}-\xi_{i+1}$. Hence the conjugates of $\xi_{i+1}^{1 / q}$ are $\rho^{j} \cdot \xi_{i+1}^{1 / q}$ for $j=0,1, \ldots, q-1$, all with equal multiplicity. There exist numbers $m_{j} \in\{0,1, \ldots, q-1$ such that for $j=0,1, \ldots, q-1$ we have

$$
\sigma_{j}\left(\xi_{i+1}^{1 / q}\right)=\rho^{m} \cdot \xi_{i+1}^{1 / q}
$$

Hence

$$
\varphi_{\ell_{0}}, \ldots, \ell_{i}^{\left(\xi_{i+1}^{1 / q}\right)}=\rho^{\sum_{j=0}^{i} \ell_{j}^{m} j} \cdot \xi_{i+1}^{1 / q}
$$

Now apply $\varphi_{\ell_{0}}, \ldots, \ell_{i}$ to (7.19). Then for each tuple $\left(\ell_{0}, \ldots, \ell_{i}\right)$ we find

$$
\rho^{\sum_{j=0}^{i} \ell_{j} m} j \xi_{i+1}^{1 / q}=\sum_{k_{0}, \ldots, k_{i}}^{a_{k_{0}}, \ldots, k_{i} \cdot \rho^{\sum_{j=0}^{i} \ell j^{k} j} \cdot \tau_{k_{0}}, \ldots, k_{i} .}
$$

Here we have a system of $q^{i+1}$ linear equations in the $q^{i+1}$ unknowns $a_{k_{0}}, \ldots, k_{i}$. The determinant of this system is exactly the square root of the discriminant of $K_{i}$ over $K$, hence nonzero. Consequently there is in $\mathbb{C}^{q^{i+1}}$ just one solution of the system. But we know that solution:

$$
\begin{aligned}
& a_{k_{0}}, \ldots, k_{i}=0 \quad \text { if } \quad\left(k_{0}, \ldots, k_{i}\right) \neq\left(m_{0}, \ldots, m_{i}\right), \\
& a_{m_{0}}, \ldots, m_{i}=\xi_{i+1}^{1 / q} \cdot \tau_{m_{0}}^{-1}, \ldots, m_{i} .
\end{aligned}
$$

The latter equation now yields an equation over $K$ :

$$
\xi_{i+1}=a_{m_{0}}^{q}, \ldots, m_{i} \cdot \prod_{j=0}^{i} \xi_{j}^{m}
$$

In case (1) this leads to the ideal equation

$$
\left(\frac{p_{i+1}}{\mathfrak{p}_{i+1}^{\prime}}\right)^{h_{i+1}}=a^{q} \cdot \prod_{j=1}^{i}\left(\frac{p_{j}}{\mathfrak{p}_{j}^{\prime}}\right)^{m}{ }^{m}{ }^{h},
$$

and in case (2) to

$$
\mathfrak{p}_{i+1}^{(\prime)^{h}} \mathbf{i + 1}=a^{q} \cdot \prod_{j=1}^{i} p_{j}^{(\prime)^{m}}{ }^{\cdot h_{j}},
$$

(where $\mathfrak{p}^{(')}$ stands for $\mathfrak{p}$ or $\mathfrak{p}^{\prime}$ ) for some fractional ideal $\mathfrak{a}$ (note that $\left(\xi_{0}\right)=(1)$ ). Because of unique factorization for ideals it follows in both cases that $q$ divides all $m_{j} \cdot h_{j}$ for $j=1, \ldots, i$ and $h_{i+1}$. This contradicts the assumption $q \nmid h$.

Remarks. 1. If $\operatorname{ord}_{p}\left(\alpha_{1}^{b} \cdot \ldots \cdot \alpha_{n}^{b_{n}}-1\right)>1 /(p-1)$ then

$$
\operatorname{ord}_{p}\left(\alpha_{1}^{b_{1}} \cdot \ldots \cdot \alpha_{n}^{b_{n}}-1\right)=\operatorname{ord}_{p}\left(b_{1} \cdot \log _{p}\left(\alpha_{1}\right)+\ldots+b_{n} \cdot \log _{p}\left(\alpha_{n}\right)\right)
$$

We prefer to work with the logarithmic version, since that is the one we use in the computational method of reducing the upper bounds.
2. In order to apply Yu's lemma we can take for $q$ the smallest odd prime that does not divide $h \cdot p \cdot\left(p^{f} p_{-1}\right)$.
3. The author is grateful to M.A.J.G. van der Vlugt (Leiden) for discussions on the above lemma.

We now proceed to compute the constants $C_{1}$ to $C_{12}$. To find $C_{1}$ and $C_{2}$ we apply Lemma 2.6 to $\Lambda_{i}^{*}$, for all $i \in I_{U}$. Then we find for each such $i$ constants $\mathrm{C}_{1, i}, \mathrm{C}_{2, i}$ such that, under the conditions

$$
u_{i}+\lambda_{i} \geqslant \gamma_{i}, \quad B^{*} \geqslant \max \left(2, \frac{4}{3} \cdot t_{i} \cdot\left(p_{i}^{f_{p_{i}}^{/ 2}}-1\right)\right)
$$

(where $t_{i}$ denotes the number of terms in $\Lambda_{i}^{*}$ ), we obtain

$$
\operatorname{ord}_{p_{i}}\left(\Lambda_{i}^{*}\right)<C_{1, i}+C_{2, i} \cdot \log B^{*}
$$

By Lemma $7.6(i)$ and the relation $\operatorname{ord}_{\mathfrak{p}}=e_{\mathfrak{p}} \cdot \operatorname{ord}_{p}$, and assuming that

$$
\begin{equation*}
U \geqslant \max _{i \in I_{U}}\left(\gamma_{i}-\lambda_{i}\right), \quad B^{*} \geqslant \max _{i \in I_{U}}\left(2, \frac{4}{3} \cdot t_{i} \cdot\left(p_{i}^{f_{i}}-1\right)\right) \tag{7.20}
\end{equation*}
$$

we see that it suffices to take

$$
C_{1}=\max _{i \in I_{U}}\left(-\left(\lambda_{i}+{ }^{\text {ord }} p_{p_{i}}\left(h^{*}\right)\right)+C_{1, i} / e_{p_{i}}\right), \quad C_{2}=\max _{i \in I_{U}}\left(C_{2, i} / e_{p_{i}}\right)
$$

Then (7.14) holds.

Next we apply Lemma 2.6 to $K_{i}^{*}$ and $K_{i}^{*}$, for all $i \in I$ and $I$, respectively, to obtain $C_{3}$ and $C_{4}$. By $X^{(')}$ we denote $X$ if $i \in I$, and $X$ ' if $i \in I$, There exist by Lemma 2.6 constants $C_{3, i}$ and $C_{4, i}$ such that under the conditions

$$
h_{i} \cdot c_{i}+\kappa_{i}^{(\prime)} \geqslant \gamma_{i}, \quad B^{*} \geqslant \max \left(2, \frac{4}{3} \cdot t_{i} \cdot\left(p_{i}^{f} p_{i}^{\prime 2}-1\right)\right)
$$

(where again $t_{i}$ denotes the number of terms of $K_{i}^{(\prime) *}$ ), it follows that

$$
\operatorname{ord}_{\mathfrak{p}_{i}}\left(\mathrm{~K}_{\mathrm{i}}^{(\prime)}{ }^{*}\right)<\mathrm{C}_{3, \mathrm{i}}+\mathrm{C}_{4, \mathrm{i}} \cdot \log \mathrm{~B}^{*}
$$

Again, by Lemma 7.6(ii), (ii') it follows that, under the conditions

$$
\begin{equation*}
M \geqslant \max _{i \in I U I}\left(\frac{\gamma_{i}-\kappa_{i}^{(\prime)}}{h_{i}}\right), \quad B^{*} \geqslant \max _{i \in I U I}\left(2, \frac{4}{3} \cdot t_{i} \cdot\left(p_{i}^{f_{i}}-1\right)\right) \tag{7.21}
\end{equation*}
$$

it suffices to take

$$
C_{3}=\max _{i \in I U I},\left(-\frac{\kappa_{i}^{(\prime)} \operatorname{rord}_{p_{i}}\left(h^{*}\right)}{h_{i}}+\frac{C_{3, i}}{h_{i} \cdot e_{p_{i}}}\right), \quad C_{4}=\max _{i \in I U I},\left(\frac{C_{4, i}}{h_{i} \cdot e_{p_{i}}}\right)
$$

Then (7.15) holds.

We take $\mathrm{C}_{5}$ to $\mathrm{C}_{7}$ as follows:

$$
\begin{aligned}
& C_{5}=\log \left(2 \cdot\left|\frac{\alpha^{\prime}}{\alpha}\right|\right) / 2 \cdot \log \varepsilon \quad, \quad C_{6}=\log \left(2 \cdot\left|\frac{\alpha}{\alpha^{\prime}}\right|\right) / 2 \cdot \log \varepsilon \\
& C_{7}=\left(\sum_{i \in I} \log \left|\frac{\pi_{i}}{\pi_{i}^{\prime}}\right|+\sum_{i \in I} \log \left|\frac{\pi_{i}^{\prime}}{\pi_{i}}\right|\right) / 2 \cdot \log \varepsilon
\end{aligned}
$$

Note that $\mathrm{C}_{5}$ or $\mathrm{C}_{6}$ may be negative, but that always $-\mathrm{C}_{6}<\mathrm{C}_{5}$. Further, $\mathrm{C}_{7}$ is always strictly positive, unless $\mathrm{I}=\mathrm{I}$ ' $=\varnothing$. Next we show how to take $\mathrm{C}_{8}$ and $\mathrm{C}_{9}$. Suppose first that

$$
n>\max \left(C_{5}, 0\right)
$$

Then, from $\varepsilon \cdot \varepsilon^{\prime}= \pm 1$ and the choice of $\pi_{i}$ we find by (7.8) that

$$
\left|\frac{\chi}{\chi^{\prime}}\right|=\left|\frac{\alpha}{\alpha^{\prime}}\right| \cdot\left|\frac{\varepsilon}{\varepsilon^{\prime}}\right|^{n} \cdot \prod_{i \in I}\left|\frac{\pi_{i}}{\pi_{i}^{\prime}}\right|^{c_{i}} \cdot \prod_{i \in I},\left|\frac{\pi_{i}^{\prime}}{\pi_{i}}\right|^{c_{i}} \geqslant\left|\frac{\alpha}{\alpha^{\prime}}\right| \cdot \varepsilon^{2 \cdot n}>2,
$$

which expresses that the first term of $G_{\alpha}$ dominates. Put

$$
P=\prod_{i \in I_{U}} p_{i}
$$

Then we infer

$$
\begin{aligned}
P^{U} & \geqslant \prod_{i \in I} p_{U} p_{i}^{u_{i}}=|\chi-\chi,|/ 2 \cdot V D>|\chi| / 4 \cdot V D \\
& =\frac{|\alpha|}{4 \sqrt{D}} \cdot \varepsilon^{n} \cdot \prod_{i \in I}\left|\pi_{i}\right|^{c_{i}} \cdot \prod_{i \in I},\left|\pi_{i}^{\prime}\right|^{c_{i}}>\frac{|\alpha|}{4 \sqrt{D}} \cdot \varepsilon^{n},
\end{aligned}
$$

hence

$$
\mathrm{n}<\left(\log \left(\frac{4 \sqrt{ } \mathrm{D}}{|\alpha|}\right)+U \cdot \log (\mathrm{P})\right) / \log \varepsilon .
$$

Next suppose that

$$
n<\min \left(-\left(C_{6}+C_{7} \cdot M\right), 0\right)
$$

Then we find that the second term of $G_{\alpha}$ dominates, namely

$$
\begin{aligned}
& \left|\frac{\chi^{\prime}}{\chi}\right|=\left|\frac{\alpha^{\prime}}{\alpha}\right| \cdot\left|\frac{\varepsilon^{\prime}}{\varepsilon}\right|^{n} \cdot \prod_{i \in I}\left|\frac{\pi_{i}^{\prime}}{\pi_{i}}\right|^{c} i \quad \prod_{i \in I},\left|\frac{\pi_{i}}{\pi_{i}^{\prime}}\right|^{c_{i}} \\
& \geqslant\left|\frac{\alpha^{\prime}}{\alpha}\right| \cdot \varepsilon^{-2 \cdot n} \cdot\left(\prod_{i \in I}\left|\frac{\pi_{i}^{\prime}}{\pi_{i}}\right| \cdot \prod_{i \in I},\left|\frac{\pi_{i}}{\pi_{i}^{\prime}}\right|\right)^{M}=\left|\frac{\alpha^{\prime}}{\alpha}\right| \cdot \varepsilon^{-2 \cdot\left(n+C_{7} \cdot M\right)} \\
& >\left|\frac{\alpha^{\prime}}{\alpha}\right| \cdot \varepsilon^{2 \cdot \mathrm{C}_{6}}=2 .
\end{aligned}
$$

Put

$$
\Gamma=\prod_{i \in I} \min \left(1,\left|\pi_{i}^{\prime}\right|\right) \cdot \prod_{i \in I} \min \left(1,\left|\pi_{i}\right|\right)
$$

Then we infer

$$
\begin{aligned}
& P^{U} \geqslant\left|\chi-\chi^{\prime}\right| / 2 \cdot \sqrt{ } D>\left|\chi^{\prime}\right| / 4 \cdot \sqrt{ } D=\frac{\left|\alpha^{\prime}\right|}{4 \sqrt{ } D} \cdot \varepsilon|\ln | \cdot \prod_{i \in I}\left|\pi_{i}^{\prime}\right|^{C_{i}} \cdot \prod_{i \in I},\left|\pi_{i}\right|^{{ }^{C}} \\
& \geqslant \frac{\left|\alpha^{\prime}\right|}{4 \sqrt{D}} \cdot \varepsilon^{|n|} \cdot \prod_{i \in I} \min \left(1,\left|\pi_{i}^{\prime}\right|\right)^{\mathrm{C}} \mathrm{i} \cdot \prod_{i \in I}, \min \left(1,\left|\pi_{i}\right|\right)^{{ }^{\mathrm{C}} \mathrm{i}} \\
& \geqslant \frac{\left|\alpha^{\prime}\right|}{4 \sqrt{D}} \cdot \varepsilon^{|n|} \cdot \Gamma^{M}>\frac{\left|\alpha^{\prime}\right|}{4 \sqrt{D}} \cdot \varepsilon^{|n|} \cdot \Gamma^{-\left(\ln \mid-C_{6}\right) / C_{7}}
\end{aligned}
$$

Hence

$$
\ln \left\lvert\,<\left(\log \left(\frac{4 \sqrt{D}}{\left|\alpha^{\prime}\right|} \cdot \Gamma^{-C_{6} / C_{7}}\right)+U \cdot \log (P)\right) / \log \left(\varepsilon \cdot \Gamma^{1 / C_{7}}\right) .\right.
$$

The remaining possibilities in cases (b) and (c) are $C_{5}<n \leqslant 0$ and $0 \leqslant n<-\left(\mathrm{C}_{6}+\mathrm{C}_{7} \cdot \mathrm{M}\right)<-\mathrm{C}_{6}$. So we may take, noting that $\Gamma \leqslant 1$,

$$
\begin{aligned}
& C_{8}= \max \left(\log \left(\frac{4 V D}{|\alpha|}\right) / \log \varepsilon, \log \left(\frac{4 \sqrt{ } D}{\left|\alpha^{\prime}\right|} \cdot \Gamma^{-C_{6} / C_{7}}\right) / \log \left(\varepsilon \cdot \Gamma^{1 / C_{7}}\right),-C_{5},-C_{6}\right), \\
& C_{9}=(\log P) / \log \left(\varepsilon \cdot \Gamma^{1 / C_{7}}\right) .
\end{aligned}
$$

Then (7.18) holds in the cases (b) and (c). Now take

$$
\begin{aligned}
& C_{10}=\max \left(C_{1}, C_{3}, \quad\left|C_{5}\right|, \quad\left|C_{6}\right|+C_{3} \cdot C_{7}, C_{8}+C_{1} \cdot C_{9}\right), \\
& C_{11}=\max \left(C_{2}, C_{4}, C_{4} \cdot C_{7}, C_{2} \cdot C_{9}\right) .
\end{aligned}
$$

Then it follows that (7.16) is true, if conditions (7.20) and (7.21) hold. Hence, by Lemma 2.1, we infer the following result.

LEMMA 7.8. In the above notation,

$$
\mathrm{B}^{*}<\mathrm{C}_{12}^{*}, \quad \mathrm{~B}<\mathrm{C}_{12}
$$

hold unconditionally, where

$$
\begin{aligned}
C_{12}^{*}= & \max \left(2 \cdot\left(N+h^{*} \cdot C_{10}+h^{*} \cdot C_{11} \cdot \log \left(h^{*} \cdot C_{11}\right)\right), \max _{i \in I_{U}}\left(h^{*} \cdot\left(\gamma_{i}-\lambda_{i}\right)+N\right),\right. \\
& \left.\max _{i \in I U I}\left(h^{*} \cdot \frac{\gamma_{i}-\kappa_{i}^{(\prime)}}{h_{i}}+N\right), 2, \max _{i \in I U I}^{\prime} \cup I_{U}\left(\frac{4}{3} \cdot t_{i} \cdot\left(p_{i}^{f}-1\right)\right)\right), \\
C_{12}= & \frac{1}{h^{*}} \cdot\left(C_{12}^{*}+N\right) .
\end{aligned}
$$

Proof. Clear.

Remarks. 1. Theorem 7.1 is an immediate corollary of Lemma 7.8.
2. In practice, almost always the first term in the max-definition of $C_{12}^{*}$ dominates. Moreover, the term N will in practice disappear in the rounding off. Similarly, in the definitions of $C_{10}$ and $C_{11}$, the dominating factors are in practice $C_{1}$ to $C_{4}$.

### 7.7. The reduction technique.

We now want to reduce the upper bound $C_{12}$ for $B$ (or $C_{12}^{*}$ for $B^{*}$, which is equivalent), to a much smaller upper bound. We do so using the p-adic computational diophantine approximation technique described in Section 3.11.

We perform this procedure for $\Lambda=\Lambda_{i}^{*}, K_{i}^{*}, K_{i}^{*}$, for the relevant $i$. We work in the p-adic approximation lattices $\Gamma_{\mu}$ themselves, and not in the sublattices described in Section 3.13. The computational bottlenecks are the computation of the p-adic logarithms to the desired precision, and the application of the $L^{3}$-Algorithm. We refer to Chapter 3 for details. Once we have found reduced bounds for ord $(\Lambda)$ for the above mentioned $\Lambda$, we combine these bounds with Lemma 7.6 and with estimates (7.13), (7.17) and (7.18) to find reduced bounds for $B$ and $B^{*}$.

When reduced upper bounds for $B, B^{*}$ are found in this way, we may try the above procedure again, with $C_{12}, C_{12}^{*}$ replaced by their reduced analogons. We may repeat the argument as long as improvement is still being made. But at a certain stage, usually near to the actual largest solution, the procedure will not yield any further improvement. Then we have to find all solutions by some other method. One technique that may be useful is the algorithm of Fincke and Pohst, described in Section 3.6. Another way is to search directly for solutions of the original diophantine equation below the reduced bounds. In our present equation this may well be done by employing congruence arguments for finding all solutions of the second equation of system (7.9) below the obtained bounds.

### 7.8. The standard example.

In this section we shall work out the procedure outlined above for our standard example $\left\{p_{1}, \ldots, p_{s}\right\}=\{2,3,5,7\}$, thus proving Theorem 7.2. In Tables II and III we give the necessary data on the fields $K=\mathbb{Q}(\sqrt{D})$ for the 15 values of $D$, and on the factorization of $2,3,5,7$ in $K$.

Explanation of Tables II and III. For $p_{i}=2,3,5,7$ we give in Table II a generator of the ideal $\mathfrak{p}_{i}$ with $\operatorname{ord}_{p_{i}}\left(p_{i}\right)>0$ if $\mathfrak{p}_{i}$ is a principal ideal, and we give "pi" if it is not principal. In all the latter cases, $h_{i}=2$, so $p_{i}^{2}=\left(\pi_{i}\right)$ is principal. An asterisk (*) denotes a splitting
prime. Note that for each $D$ at most one of the primes 2, 3, 5, 7 splits, so $t \leqslant 1$. In the final column of Table II we give for the splitting prime $p_{i}$ a generator $\pi_{i}$ of the ideal $p_{i}^{h_{i}}$. In Table III, when $p_{i}$ and $p_{j}$ are not principal, but $\mathfrak{p}_{i} \cdot p_{j}$ is, we give a generator of it. The autor is grateful to R.J. Kooman (Leiden) for checking these tables.

From Tables II and III it is easy to find all possibilities for I, I' and $\alpha$. We may assume $I^{\prime}=\varnothing$. In Table IV we give all possible $I, I_{U}, \alpha$ (we give primes $p_{i}$ instead of indices $i \quad$ ). An asterisk (*) appears when $(\alpha) \neq\left(\alpha^{\prime}\right)$. The set $I_{U}$ is found by checking $G_{\alpha}\left(\bmod p_{i}\right)$ for all $p_{i}$.

There are 54 cases with $I=\varnothing$ (the "symmetric" cases), and 54 cases with I $\neq \varnothing$ (the "asymmetric" cases). We start with the symmetric cases. This incorporates all cases with $\mathrm{D}=3,5,35,42,210$, when none of the primes $2,3,5,7$ splits in $\mathbb{Q}(\sqrt{D})$. Now, $t=0$, hence equation (7.10) becomes

$$
\begin{equation*}
G_{\alpha}(n)=\frac{\alpha}{2 \sqrt{D}} \cdot \varepsilon^{n}-\frac{\alpha^{\prime}}{2 \sqrt{D}} \cdot \varepsilon^{, n}= \pm \prod_{i \in I_{U}} p_{i}^{u_{i}} \tag{7.22}
\end{equation*}
$$

With $A=\varepsilon+\varepsilon^{\prime} \in \mathbb{Z}, \quad B=N \varepsilon=\varepsilon \cdot \varepsilon^{\prime}= \pm 1$, we have for all $n \in \mathbb{Z}$

$$
G_{\alpha}(n+2)=A \cdot G_{\alpha}(n+1)-B \cdot G_{\alpha}(n)
$$

Since $(\alpha)=\left(\alpha^{\prime}\right)$, there is an $n_{0} \in \mathbb{Z}$ such that $\alpha^{\prime}= \pm^{n^{n}} \cdot \alpha$. Hence

$$
\left|G_{\alpha}\left(n_{0}-n\right)\right|=\left|G_{\alpha}(n)\right|
$$

for all $n \in \mathbb{Z}$, which explains why we call these cases "symmetric". In this situation we can apply elementary congruence arguments, as explained in Section 4.5. We have the following result.

LEMMA 7.9. Let $\left\{\mathrm{p}_{1}, \ldots, \mathrm{p}_{4}\right\}=\{2,3,5,7$. Equation (7.1) with conditions (7.2) and $\mathrm{I}=\varnothing$ has exactly 91 solutions, that appear in Table I marked with an asterisk (*).

Sketch of proof. In Table $V$ we give the necessary data for these 54 cases. We explain this table, and leave many details to the reader to check. For
 $\ell_{1}$ is given, then $p^{\ell_{1}+1} \nprec G_{\alpha}(n)$ for all $n \in \mathbb{Z}$, and $p^{\ell_{1}} \mid G_{\alpha}(n)$ for at least one $n \in \mathbb{Z}$. If $n_{1}, a_{1}$ are given, then

$$
p^{\ell_{1}+1} \mid G_{\alpha}(n) \Leftrightarrow n \equiv n_{1}\left(\bmod a_{1}\right)
$$

Define $n_{2}=a_{1}$ if $n_{1}=0$, and $n_{2}=n_{1}$ if $n_{1} \neq 0$. Then $n_{2}$ is the smallest positive index such that $p^{\ell_{1}+1} \mid G_{\alpha}\left(n_{2}\right)$. Now it is true that

$$
G_{\alpha}\left(n_{2}\right) \mid G_{\alpha}(n) \quad \text { whenever } \quad n \equiv n_{1}\left(\bmod a_{1}\right)
$$

This is related to symmetry properties of the recurrence sequence $\left\{G_{\alpha}(n)\right\}_{n=-\infty}^{\infty}$. For $q=2,3,5,7$ we have defined

$$
h_{q}=\operatorname{ord}_{q}\left(G_{\alpha}\left(n_{2}\right)\right)
$$

Hence $2^{h_{2}} \cdot 3^{h_{3}} \cdot 5^{h_{5}} \cdot 7^{h_{7}}$ I $G_{\alpha}(n)$ whenever $p^{\ell_{1}+1} \mid G_{\alpha}(n)$. We have taken $\ell_{1}$ so large that always

$$
\begin{equation*}
G_{\alpha}\left(n_{2}\right)>2^{h_{2}} \cdot 3^{h_{3}} \cdot 5^{h_{5}} \cdot 7^{h_{7}} \tag{7.23}
\end{equation*}
$$

Consequently, there exists some prime $r \geqslant 11$ that divides $G_{\alpha}\left(n_{2}\right)$, hence $r$ divides all $G_{\alpha}(n)$ with $p^{\ell_{1}+1} \mid G_{\alpha}(n)$. It follows that for a solution of equation (7.22) we must have

$$
\operatorname{ord}_{p}\left(G_{\alpha}(n)\right) \leqslant \ell_{1}
$$

In this way we find with ease all solutions of (7.22).

Let us illustrate this with the example $D=3, \alpha=\sqrt{ } 3$. Then

$$
G_{\alpha}(n)=\frac{1}{2} \cdot(2+\sqrt{ } 3)^{n}+\frac{1}{2} \cdot(2-\sqrt{ } 3)^{n}
$$

and $G_{\alpha}(-n)=G_{\alpha}(n)$. We have for $G_{\alpha}(n)$ :

| n | 0 | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 | 9 | 10 | 11 | 12 | 13 | 14 | 15 |
| :---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| $\mathrm{G}_{\alpha}(\mathrm{n})$ | 1 | 2 | 7 | 26 | 97 | 362 | $\ldots$ |  |  |  |  | $G_{\alpha}(14)$ | $=50843527$ |  |  |  |
| $\bmod$ | 4 | 1 | 2 | -1 | 2 | 1 | 2 | -1 | 2 | 1 | 2 | -1 | 2 | 1 | 2 | -1 |
| $\bmod$ | 3 | 1 | -1 | 1 | -1 | 1 | -1 | 1 | -1 | 1 | -1 | 1 | -1 | 1 | -1 | 1 |
| $\bmod$ | 5 | 1 | 2 | 2 | 1 | 2 | 2 | 1 | 2 | 2 | 1 | 2 | 2 | 1 | 2 | 2 |
| $\bmod 49$ | 1 | 2 | 7 | -23 | -1 | 19 | -21 | -5 | 1 | 9 | -14 | -16 | -1 | 12 | 0 | -12 |

We see that $2^{2}, 3,5 \nmid G_{\alpha}(n)$ for all $n \in \mathbb{Z}$, and $2 \mid G_{\alpha}(n)$ if and only if $n$ odd. So $p=7$ is the only interesting case. We have $7 \mid G_{\alpha}(n)$ if
and only if $n \equiv 2(\bmod 4), 7^{2} \mid G_{\alpha}(n) \quad$ if and only if $n \equiv 14(\bmod 28)$, (and in general

$$
7^{k} \mid G_{\alpha}(n) \Leftrightarrow n \equiv 2 \cdot 7^{k-1}\left(\bmod 4 \cdot 7^{k-1}\right)
$$

for $k \geqslant 1$, and a similar relation holds for any symmetric recurrence and any prime $p$ for which arbitrary high powers of $p$ occur in $G_{\alpha}(n)$, cf. Lemma 4.10). Now, $\ell_{1}=0$ does not lead to (7.23), since then $n_{2}=2$, and $G_{\alpha}(2)=7$, so that no suitable $r$ exists. But with $\ell_{1}=1$ we have $n_{2}=14$, and $h_{2}=h_{3}=h_{5}=0, h_{7}=2$, and (7.23) holds, since $G_{\alpha}(14)>7^{2}$. Hence there exists a prime $r \geqslant 11$ such that $r \mid G_{\alpha}(14)$, and thus $r \mid G_{\alpha}(n)$ whenever $7^{2} \mid G_{\alpha}(n)$. It follows that for solutions of (7.22) we have $G_{\alpha}(n) \leqslant 2^{1} \cdot 3^{0} \cdot 5^{0} \cdot 7^{1}=14$, so that all solutions can be read from the above table. Note that it is not necessary that $r$ is known explicitly, only that $G_{\alpha}\left(n_{2}\right)$ is large enough. In our example, $r=337$ or $r=3079$ satisfy.

Finally we treat the remaining 54 cases, where $I \neq \varnothing$. Then we need the non-elementary reduction technique described in Sections 7.5 to 7.7.

In all our instances, the set $I$ contains only one element, since there is only one splitting prime. We denote by $\pi$ the $\pi_{i}$ belonging to this prime, and we write $m$ for $c_{i}$. Equation (7.10) now reads

$$
\frac{\alpha}{2 \sqrt{D}} \cdot \varepsilon^{n} \cdot \pi^{m}-\frac{\alpha^{\prime}}{2 \sqrt{D}} \cdot \varepsilon^{, n} \cdot \pi,^{m}= \pm \prod_{j \in I_{U}} p_{j}^{u}
$$

We computed the constants $C_{1}$ to $C_{12}, C_{12}^{*}$, according to Section 7.6 , for each of the 54 cases. We omit the details of these computations, and simply give the data in Table VI. In this table we give for each $D$ the $p_{i} \in I_{U}$ together with the $v_{i}$ and $\lambda_{i}$ (it turns out that the $\lambda_{i}$ do not depend on the $\alpha$, only on the $p_{i}$ ). The values " $n_{\varepsilon}, n_{\pi}, n_{2}, n_{3}, n_{5}, n_{7}$ " are the integers such that

$$
\alpha^{2}= \pm \varepsilon^{{ }^{n}}{ }^{\varepsilon} \cdot \pi^{n^{n}} \pi \cdot 2^{n_{2}} \cdot \ldots \cdot 7^{n^{7}} .
$$

It follows that in all cases we have $C_{12}^{*}<3.23 \times 10^{30}$.

The next step is to define the lattices, and find lower bounds for the shortest nonzero vectors in the lattices. We start with treating the $\Lambda_{i}^{*}$, of which there are 3 for each of the 10 D 's. We have computed the 30 values of

$$
\vartheta=-\frac{\log _{p_{i}}\left(\frac{\pi}{\pi^{\prime}}\right)}{\log _{p_{i}}\left(\frac{\varepsilon}{\varepsilon^{\prime}}\right)} \text { or }-\frac{\log _{p_{i}}\left(\frac{\varepsilon}{\varepsilon^{\prime}}\right)}{\log _{p_{i}}\left(\frac{\pi}{\pi^{\prime}}\right)}
$$

such that it is a $p_{i}$-adic integer, to the desired precision of $\mu$ digits. We took $\mu$ as follows:

| $p_{i}$ | $\mu$ | $p_{i}^{\mu}$ |
| :---: | :---: | :---: |
| 2 | 209 | $8.22 \times 10^{62}$ |
| 3 | 133 | $2.87 \times 10^{63}$ |
| 5 | 95 | $2.52 \times 10^{66}$ |
| 7 | 76 | $1.69 \times 10^{64}$ |

in order to have $p_{i}^{\mu}$ somewhat larger than the maximal $C_{12}^{* 2}$, being $1.05 \times 10^{61}$. We computed the 30 values of the $\vartheta^{(\mu)}$, $s$, but do not give them here. The lattices $\Gamma_{\mu}$ are generated by the column vectors of the matrices

$$
\left(\begin{array}{cc}
1 & 0 \\
\vartheta^{(\mu)} & \mathrm{p}^{\mu}
\end{array}\right) .
$$

We performed the p-adic continued fraction algorithm of Section 3.10 for each of these 30 lattices. In the table below we give for each $D$ the maximal $C_{12}^{*}$ (there is one for each $\alpha$ ), and the minimal bound for $\ell\left(\Gamma_{\mu}\right)$ (there is one for each $i \in I_{U}$ ) that we found. We omit further details.

| D | p | $\mu_{0}$ | $C_{12}^{*} \leqslant$ | $\ell\left(\Gamma_{\mu}\right)>$ | U $\leqslant$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 2, 3, 5 | $1.5,1.0,1.0$ | $3.19 \times 10^{28}$ | $8.26 \times 10^{30}$ | 210 |
| 6 | 2, 3, 7 | $1.5,1.5,1.0$ | $2.72 \times 10^{26}$ | $2.05 \times 10^{31}$ | 210 |
| 7 | 2, 5, 7 | 2.0, 1.0, 0.5 | $1.07 \times 10^{30}$ | $2.43 \times 10^{31}$ | 210 |
| 10 | 2, 5, 7 | $1.5,0.5,1.0$ | $3.22 \times 10^{29}$ | $2.22 \times 10^{31}$ | 210 |
| 14 | 2, 3, 7 | $1.5,1.0,0.5$ | $4.80 \times 10^{26}$ | 1. $48 \times 10^{31}$ | 210 |
| 15 | 2, 3, 5 | 3.5, 1.5, 0.5 | $2.15 \times 10^{28}$ | $1.55 \times 10^{31}$ | 212 |
| 21 | 2, 3, 7 | $3.0,0.5,0.5$ | $1.90 \times 10^{26}$ | $7.78 \times 10^{30}$ | 211 |
| 30 | 2, 3, 5 | $2.5,0.5,0.5$ | $4.15 \times 10^{28}$ | $1.37 \times 10^{31}$ | 211 |
| 70 | 2, 5, 7 | $2.5,0.5,0.5$ | $3.23 \times 10^{30}$ | $2.51 \times 10^{31}$ | 211 |
| 105 | 3, 5, 7 | $1.5,0.5,0.5$ | $4.54 \times 10^{29}$ | 3. $96 \times 10^{31}$ | 134 |

In all cases, $\ell\left(\Gamma_{\mu}\right)>\sqrt{ } 2 \cdot C_{12}^{*}$. Hence Lemma 3.14 with $n=2, c_{1}=0, c_{2}=1$ yields

$$
\operatorname{ord}_{p_{i}}\left(\Lambda_{i}^{*}\right)<\mu+\mu_{0}, \quad i \in I_{U}
$$

where

$$
\mu_{0}=\min \left(\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}\left(\frac{\varepsilon}{\varepsilon^{\prime}}\right)\right), \operatorname{ord}_{p_{i}}\left(\log _{p_{i}}\left(\frac{\pi}{\pi^{\prime}}\right)\right)\right)
$$

as given above. By $\lambda_{i}+\operatorname{ord}_{p_{i}}\left(h^{*}\right) \geqslant 0$ we obtain from Lemma $7.6(i)$ upper bounds for $u_{i}, i \in I_{U}$, hence the upper bounds for $U$, as given above.

Next, we treat the $K_{i}^{*}$, one for each $D$, having 5 terms, namely

$$
K_{i}^{*}=n^{*} \cdot \log _{p_{i}}\left(\varepsilon^{\prime}\right)+m^{*} \cdot \log _{p_{i}}\left(\pi^{\prime}\right)-\sum_{\substack{1 \leqslant j \leqslant 4 \\ j \neq i}} u_{j}^{*} \cdot \log _{p_{i}}\left(p_{j}\right)
$$

where $i \in I$, so $p_{i}$ is the splitting prime. We have the following data.

| D |  | $\sqrt{D}\left(\bmod \mathrm{p}_{\mathrm{i}}\right)$ | $\varepsilon$ | $\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}(\cdot)\right)$ |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
|  |  |  |  | $\pi$, | 2 | 3 | 5 | 7 |
| 2 | 7 | 3 | 1 | 2 | 1 | 1 | 1 | - |
| 6 | 5 | 4 | 1 | 1 | 1 | 1 | - | 2 |
| 7 | 3 | 1 | 1 | 1 | 1 | - | 1 | 1 |
| 10 | 3 | 2 | 1 | 1 | 1 | - | 1 | 1 |
| 14 | 5 | 2 | 1 | 1 | 1 | 1 | - | 2 |
| 15 | 7 | 6 | 1 | 1 | 1 | 1 | 1 | - |
| 21 | 5 | 4 | 1 | 1 | 1 | 1 | - | 2 |
| 30 | 7 | 4 | 1 | 1 | 1 | 1 | 1 | - |
| 70 | 3 | 2 | 1 | 1 | 1 | - | 1 | 1 |
| 105 | 2 | $1(\bmod 4)$ | 2 | 4 | - | 2 | 2 | 3 |

From this table our choice for $\sqrt{ } D\left(\bmod p_{i}\right)$ becomes clear. It follows that $\operatorname{ord}_{p_{i}}\left(\log _{p_{i}}\left(\varepsilon^{\prime}\right)\right)$ is always the least one of the five ord $p_{i}$ 's in the above table. So we define:

$$
\vartheta_{1}=-\frac{\log _{p_{i}}\left(\pi^{\prime}\right)}{\log _{p_{i}}\left(\varepsilon^{\prime}\right)}, \quad \vartheta_{2,3,4}=-\frac{\log _{p_{i}}\left(p_{j}\right)}{\log _{p_{i}}\left(\varepsilon^{\prime}\right)}, \quad(j \in\{1,2,3,4\}, \quad j \neq i)
$$

and we computed these numbers up to $\mu$ digits, with $\mu$ as follows:

| $\mathrm{p}_{\mathrm{i}}$ | $\mu$ | $\mathrm{p}_{\mathrm{i}}^{\mu}$ |
| :---: | :---: | :---: |
| 2 | 539 | $1.80 \times 10^{162}$ |
| 3 | 343 | $4.49 \times 10^{163}$ |
| 5 | 245 | $1.77 \times 10^{171}$ |
| 7 | 196 | $4.36 \times 10^{165}$ |

so that $p_{i}^{\mu}$ is somewhat larger than the maximal $C_{12}^{* 5}$. We computed the 40 values of the $\vartheta_{1,2,3,4}^{(\mu)}$, but do not give them here. The lattices $\Gamma_{\mu}$ are generated by the columns of the following matrices:

$$
\left(\begin{array}{ccccc}
1 & 0 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 & 0 \\
0 & 0 & 1 & 0 & 0 \\
0 & 0 & 0 & 1 & 0 \\
\vartheta_{1}^{(\mu)} & \vartheta_{2}^{(\mu)} & \vartheta_{3}^{(\mu)} & \vartheta_{4}^{(\mu)} & p^{\mu}
\end{array}\right)
$$

We computed the reduced bases of the 10 lattices by the $L^{3}$-algorithm. Again, we omit the computational details. We found data as follows.

| D | p in I | $\mu$ | $\mu_{0}$ | $\mathrm{C}_{12}^{*} \leqslant$ | $\ell\left(\Gamma_{\mu}\right)>$ | $\mathrm{M} \leqslant$ |
| ---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 7 | 196 | 1 | $3.19 \times 10^{28}$ | $2.25 \times 10^{32}$ | 196 |
| 6 | 5 | 245 | 1 | $2.72 \times 10^{26}$ | $2.16 \times 10^{33}$ | 245 |
| 7 | 3 | 343 | 1 | $1.07 \times 10^{30}$ | $1.14 \times 10^{32}$ | 343 |
| 10 | 3 | 343 | 1 | $3.22 \times 10^{29}$ | $1.07 \times 10^{32}$ | 343 |
| 14 | 5 | 245 | 1 | $4.80 \times 10^{26}$ | $4.92 \times 10^{33}$ | 245 |
| 15 | 7 | 196 | 1 | $2.15 \times 10^{28}$ | $2.78 \times 10^{32}$ | 196 |
| 21 | 5 | 245 | 1 | $1.90 \times 10^{26}$ | $4.37 \times 10^{33}$ | 245 |
| 30 | 7 | 196 | 1 | $4.15 \times 10^{28}$ | $2.69 \times 10^{32}$ | 196 |
| 70 | 3 | 343 | 1 | $3.23 \times 10^{30}$ | $1.03 \times 10^{32}$ | 343 |
| 105 | 2 | 539 | 2 | $4.54 \times 10^{29}$ | $6.68 \times 10^{31}$ | 540 |

In all instances, $\ell\left(\Gamma_{\mu}\right)>\sqrt{ } 5 \cdot C_{12}^{*}$, so that by Lemmas 3.14 and $7.6(\mathrm{ii})$ and $\kappa_{i}+\operatorname{ord}_{p_{i}}\left(h^{*}\right) \geqslant 0$ and $h_{i} \geqslant 1$ we have $M \leqslant \operatorname{ord}_{p_{i}}\left(K_{i}^{*}\right)<\mu+\mu_{0}$, hence an upper bound for $M$ as given in the table above.

Finally, we compute the new, reduced bounds for $|n|$, and thus for $B$, by

$$
|n|<\max \left(C_{5}, C_{6}+C_{7} \cdot M, C_{8}+C_{9} \cdot U\right)
$$

Hence we find data as in the following table.

| D | $\mathrm{C}_{5}<\mathrm{C}_{6}<\mathrm{C}_{7}<\mathrm{C}_{8}<\mathrm{C}_{9}<$ | $\mathrm{M} \leqslant$ | $\mathrm{U} \leqslant$ | $\ln \mid \leqslant$ | $\mathrm{B} \leqslant$ | $\mathrm{N} \leqslant$ | $\mathrm{B}^{*} \leqslant$ |  |  |  |  |
| ---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | ---: |
| 2 | 0.394 | 0.394 | 0.420 | 1.967 | 3.859 | 196 | 210 | 812 | 812 | 3 | 1627 |
| 6 | 0.152 | 0.652 | 0.190 | 1.345 | 1.631 | 245 | 210 | 343 | 343 | 3 | 689 |
| 7 | 0.126 | 0.626 | 0.357 | 2.702 | 2.757 | 343 | 210 | 581 | 581 | 2 | 1164 |
| 10 | 0.601 | 0.191 | 0.181 | 1.396 | 2.337 | 343 | 210 | 492 | 492 | 3 | 987 |
| 14 | 0.102 | 0.602 | 0.325 | 1.861 | 1.508 | 245 | 210 | 318 | 318 | 3 | 639 |
| 15 | 0.540 | 0.668 | 0.257 | 1.394 | 1.649 | 196 | 212 | 350 | 350 | 2 | 702 |
| 21 | 0.222 | 0.722 | 0.142 | 1.564 | 2.386 | 245 | 211 | 505 | 505 | 1 | 1011 |
| 30 | 0.414 | 0.613 | 0.399 | 1.239 | 1.102 | 196 | 211 | 233 | 233 | 3 | 469 |
| 70 | 0.362 | 0.556 | 0.390 | 2.729 | 1.505 | 343 | 211 | 320 | 343 | 3 | 689 |
| 105 | 0.390 | 0.579 | 0.379 | 3.232 | 2.545 | 540 | 134 | 344 | 540 | 1 | 1081 |

Here we used $B^{*} \leqslant h^{*} \cdot B+N$ and $h^{*}=2$. So in one step we have reduced the bound $B^{*}<3.23 \times 10^{30}$ to $B^{*} \leqslant 1627$. The total computation time was 1715 sec , on average 0.7 sec for each 2-dimensional lattice, and 170 sec for each 5-dimensional lattice.

We made a further reduction step, now using the reduced bound for $B^{*}$ as given above in stead of $C_{12}^{*}$. We give the data for the $\Lambda_{i}^{*}$ in the tables below. For $\mu$ we took $\mu_{1} \cdot \mu_{2}$, with $\mu_{1}, \mu_{2}$ as below:

| $p$ | 2 | 3 | 5 | 7 |
| :--- | ---: | ---: | ---: | ---: |
| $\mu_{2}$ | 11 | 7 | 5 | 4 |,


| D | $\mathrm{B}^{*} \leqslant$ | $\sqrt{ } 2 \cdot \mathrm{~B}^{*}<$ | $\mu_{1}$ | $\mu \leqslant$ | $\ell\left(\Gamma_{\mu}\right) \geqslant$ | $\mu_{0} \leqslant$ | $\mathrm{U} \leqslant$ |
| ---: | ---: | ---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 1627 | 2301 | 2 | 22 | $1.82 \times 10^{3}$ | 1.5 | 23 |
| 6 | 689 | 975 | 3 | 33 | $3.99 \times 10^{4}$ | 1.5 | 34 |
| 7 | 1164 | 1647 | 3 | 33 | $4.50 \times 10^{4}$ | 2 | 34 |
| 10 | 987 | 1396 | 3 | 33 | $5.91 \times 10^{4}$ | 1.5 | 34 |
| 14 | 639 | 904 | 3 | 33 | $2.58 \times 10^{4}$ | 1.5 | 34 |
| 15 | 702 | 993 | 3 | 33 | $7.36 \times 10^{4}$ | 3.5 | 36 |
| 21 | 1011 | 1430 | 3 | 33 | $2.00 \times 10^{4}$ | 3 | 35 |
| 30 | 469 | 664 | 2 | 22 | $9.98 \times 10^{2}$ | 2.5 | 24 |
| 70 | 689 | 975 | 3 | 33 | $5.76 \times 10^{4}$ | 2.5 | 35 |
| 105 | 1081 | 1529 | 3 | 21 | $3.89 \times 10^{4}$ | 1.5 | 22 |

We found $\ell\left(\Gamma_{\mu}\right)$ and bounds for $U$ as given in the above table. For the $K_{i}^{*}$ we found, with $\mu=\mu_{1} \cdot \mu_{2}$ with $\mu_{2}$ as above, and $\mu_{1}$ as in the table below, the results given in that table.

| D | $\mathrm{B}^{*} \leqslant$ | $\sqrt{ } 5 \cdot \mathrm{~B}^{*}<$ | $\mu_{1}$ | $\mu \leqslant$ | $\ell\left(\Gamma_{\mu}\right) \geqslant$ | $\mu_{0} \leqslant$ | $\mathrm{M} \leqslant$ | $\|\mathrm{n}\| \leqslant$ | $\mathrm{B} \leqslant$ | $\mathrm{B}^{*} \leqslant$ |
| ---: | ---: | ---: | ---: | ---: | :---: | ---: | ---: | ---: | ---: | ---: |
| 2 | 1627 | 3639 | 7 | 28 | $1.24 \times 10^{4}$ | 1 | 28 | 90 | 90 | 183 |
| 6 | 689 | 1541 | 6 | 30 | $4.04 \times 10^{3}$ | 1 | 30 | 145 | 145 | 293 |
| 7 | 1164 | 2603 | 7 | 49 | $1.07 \times 10^{4}$ | 1 | 49 | 96 | 96 | 194 |
| 10 | 987 | 2207 | 7 | 49 | $1.16 \times 10^{4}$ | 1 | 49 | 80 | 80 | 163 |
| 14 | 639 | 1429 | 6 | 30 | $3.07 \times 10^{3}$ | 1 | 30 | 53 | 53 | 109 |
| 15 | 702 | 1570 | 6 | 24 | $2.70 \times 10^{3}$ | 1 | 24 | 60 | 60 | 122 |
| 21 | 1011 | 2261 | 6 | 30 | $3.88 \times 10^{3}$ | 1 | 30 | 85 | 85 | 171 |
| 30 | 469 | 1049 | 6 | 24 | $2.50 \times 10^{3}$ | 1 | 24 | 27 | 27 | 57 |
| 70 | 689 | 1541 | 6 | 42 | $1.90 \times 10^{3}$ | 1 | 42 | 55 | 55 | 113 |
| 105 | 1081 | 2418 | 7 | 77 | $1.00 \times 10^{4}$ | 2 | 78 | 59 | 78 | 157 |

The computation time was 15 sec .

We made a third step, and give data like above, for $\Lambda_{i}^{*}$ :

| D | $\mathrm{B}^{*} \leqslant$ | $\sqrt{ } 2 \cdot \mathrm{~B}^{*}<\mu_{1}$ | $\mu \leqslant\left(\Gamma_{\mu}\right) \geqslant$ | $\mu_{0} \leqslant$ | $\mathrm{U} \leqslant$ |  |  |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| 2 | 183 | 258.9 | 2 | 22 | 1821 | 1.5 | 23 |
| 6 | 299 | 414.4 | 2 | 22 | 875 | 1.5 | 23 |
| 7 | 194 | 274.4 | 2 | 22 | 1285 | 2 | 23 |
| 10 | 163 | 230.6 | 2 | 22 | 634 | 1.5 | 23 |
| 14 | 109 | 154.2 | 2 | 22 | 268 | 1.5 | 23 |
| 15 | 122 | 172.6 | 2 | 22 | 873 | 3.5 | 25 |
| 21 | 171 | 241.9 | 2 | 22 | 818 | 3 | 25 |
| 30 | 57 | 80.7 | 2 | 22 | 998 | 2.5 | 24 |
| 70 | 113 | 159.9 | 2 | 22 | 585 | 2.5 | 24 |
| 105 | 157 | 222.1 | 2 | 14 | 281 | 1.5 | 15 |

and for $K_{i}^{*}$ :

| D | $B^{*} \leqslant$ | $\sqrt{5} \cdot \mathrm{~B}^{*}<$ | $\mu_{1}$ | $\mu \leqslant$ | $\ell\left(\Gamma_{\mu}\right) \geqslant$ | $\mu_{0} \leqslant$ | M $\leqslant$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 183 | 409.3 | 5 | 20 | 440 | 1 | 20 |
| 6 | 293 | 655.2 | 5 | 25 | 665 | 1 | 25 |
| 7 | 194 | 433.8 | 6 | 42 | 602 | 1 | 42 |
| 10 | 163 | 364.5 | 5 | 35 | 473 | 1 | 35 |
| 14 | 109 | 243.8 | 5 | 25 | 626 | 1 | 25 |
| 15 | 122 | 272.9 | 6 | 24 | 2700 | 1 | 24 |
| 21 | 171 | 382.4 | 5 | 25 | 645 | 1 | 25 |
| 30 | 57 | 127.5 | 4 | 16 | 129 | 1 | 16 |
| 70 | 113 | 252.7 | 5 | 35 | 366 | 1 | 35 |
| 105 | 157 | 351.1 | 5 | 55 | 354 | 2 | 56 |

and finally for $|n|$, and in more detail for $\operatorname{ord}_{p_{i}}(u)$ for $i \in I_{U}$

| D | $\mathrm{M} \leqslant$ | $u_{2} \leqslant$ | $u_{3} \leqslant$ | $u_{5} \leqslant$ | $u_{7} \leqslant$ | $\|n\| \leqslant$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| 2 | 20 | 23 | 14 | 10 | 0 | 90 |
| 6 | 25 | 23 | 15 | 0 | 8 | 38 |
| 7 | 42 | 23 | 0 | 10 | 8 | 66 |
| 10 | 35 | 23 | 0 | 10 | 8 | 55 |
| 14 | 25 | 23 | 14 | 0 | 8 | 36 |
| 15 | 24 | 25 | 15 | 10 | 0 | 42 |
| 21 | 25 | 24 | 14 | 0 | 8 | 61 |
| 30 | 16 | 24 | 14 | 10 | 0 | 27 |
| 70 | 35 | 24 | 0 | 10 | 8 | 65 |
| 105 | 56 | 0 | 14 | 10 | 8 | 41 |

Now we will not find any further improvement if we proceed in the same way. But the upper bounds are now small enough to admit enumeration of the remaining possibilities, making use of mod $p$ arithmetic for $p=2,3,5,7$. We did so, and found the remaining solutions, presented in Table I. We used only 3 sec computer time for this last step.

This completes the proof of Theorem 7.2.
7.9. Tables.

Table I. (Theorem 7.2.)

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Table I. (cont.)
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Table I. (cont.)









Table I. (cont.)


$\Rightarrow$



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Table II.

| D | h | $\varepsilon$ | $N \varepsilon$ | $p_{1}$ | $p_{2}$ | $p_{3}$ | $p_{4}$ | $\pi_{i}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 1 | $1+\sqrt{2}$ | -1 | $\sqrt{ } 2$ | 3 | 5 | $1+2 \sqrt{ } 2^{*}$ | $1+2 \sqrt{ } 2$ |
| 3 | 1 | $2+\sqrt{3}$ | 1 | $1+\sqrt{3}$ | $\sqrt{3}$ | 5 | 7 | - |
| 5 | 1 | $\frac{1}{2}(1+\sqrt{5})$ | -1 | 2 | 3 | $\sqrt{ } 5$ | 7 | - |
| 6 | 1 | $5+2 \sqrt{6}$ | 1 | $2+\sqrt{6}$ | $3+\sqrt{6}$ | $1+\sqrt{6}$ * | 7 | $1+\sqrt{6}$ |
| 7 | 1 | $8+3 \sqrt{7}$ | 1 | $3+\sqrt{7}$ | $2+\sqrt{7}{ }^{*}$ | 5 | $\sqrt{7}$ | $2+\sqrt{7}$ |
| 10 | 2 | $3+\sqrt{10}$ | -1 | $p_{1}$ | $\mathfrak{p}_{2}^{*}$ | $p_{3}$ | 7 | $1+\sqrt{ } 10$ |
| 14 | 1 | $15+4 \sqrt{ } 14$ | 1 | $4+\sqrt{14}$ | 3 | $3+\sqrt{14}$ * | $7+2 \sqrt{ } 14$ | $3+\sqrt{14}$ |
| 15 | 2 | $4+\sqrt{15}$ | 1 | $p_{1}$ | $p_{2}$ | $p_{3}$ | $\mathfrak{p}_{4}^{*}$ | $8+\sqrt{ } 15$ |
| 21 | 1 | $\frac{1}{2}(5+\sqrt{ } 21)$ | 1 | 2 | $\frac{1}{2}(3+\sqrt{21})$ | $\frac{1}{2}(1+\sqrt{2} 21)^{*}$ | $\frac{1}{2}(7+\stackrel{4}{\sqrt{2}} 21)$ | $\frac{1}{2}(1+\sqrt{ } 21)$ |
| 30 | 2 | $11+2 \sqrt{ } 30$ | 1 | $p_{1}$ | $\mathfrak{p}_{2}$ | $5+\sqrt{ } 30$ | $\mathfrak{p}_{4}^{*}$ | $13+2 \sqrt{ } 30$ |
| 35 | 2 | $6+\sqrt{35}$ | 1 | $p_{1}$ | 3 | $p_{3}$ | $\mathfrak{p}_{4}$ | - |
| 42 | 2 | $13+2 \sqrt{ } 42$ | 1 | $p_{1}$ |  | 5 | $7+\sqrt{42}$ | - |
| 70 | 2 | $251+30 \sqrt{ } 70$ | 1 | $p_{1}$ | $p_{2}^{*}$ | $25+3 \sqrt{70}$ | $\mathfrak{p}_{4}$ | $17+2 \sqrt{ } 70$ |
| 105 | 2 | $41+4 \sqrt{ } 105$ | 1 | $p_{1}^{*}$ | $p_{2}$ | $10+\sqrt{ } 105$ | $\mathfrak{p}_{4}$ | $\frac{1}{2}(11+\sqrt{ } 105$ |
| 210 | 4 | $29+2 \sqrt{ } 210$ | 1 | $p_{1}$ | $p_{2}$ | $\mathfrak{p}_{3}$ | $p_{4}$ | - |

## Table III.

| D | $\mathfrak{p}_{1} \cdot \mathfrak{p}_{2}$ | $\mathfrak{p}_{1} \cdot \mathfrak{p}_{3}$ | $\mathfrak{p}_{1} \cdot \mathfrak{p}_{4}$ | $\mathfrak{p}_{2} \cdot \mathfrak{p}_{3}$ | $\mathfrak{p}_{2} \cdot \mathfrak{p}_{4}$ | $\mathfrak{p}_{3} \cdot \mathfrak{p}_{4}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 10 | $-2+\sqrt{ } 10$ | $\sqrt{ } 10$ | - | $5-\sqrt{ } 10$ | - | - |
| 15 | $3+\sqrt{ } 15$ | $5+\sqrt{ } 15$ | $1+\sqrt{ } 15$ | $\sqrt{ } 15$ | $6-\sqrt{ } 15$ | $-5+2 \sqrt{ } 15$ |
| 30 | $6+\sqrt{ } 30$ | - | $-4+\sqrt{ } 30$ | - | $3+\sqrt{ } 30$ | - |
| 35 | - | $5+\sqrt{ } 35$ | $7+\sqrt{ } 35$ | - | - | $\sqrt{ } 35$ |
| 42 | $6+\sqrt{ } 42$ | - | - | - | - | - |
| 70 | $-8+\sqrt{ } 70$ | - | $42+5 \sqrt{ } 70$ | - | $7+\sqrt{ } 70$ | - |
| 105 | $\frac{1}{2}(-9+\sqrt{ } 105)$ | - | $\frac{1}{2}(7+\sqrt{ } 105)$ | - | $21+2 \sqrt{ } 105$ | - |
| 210 | - | - | $14+\sqrt{ } 210$ | $15+\sqrt{ } 210$ | - | - |

Table IV.

| D | $\alpha$ | I | $\mathrm{I}_{\mathrm{U}}$ | D | $\alpha$ | I | $\mathrm{I}_{\mathrm{U}}$ | D | $\alpha$ | I | $\mathrm{I}_{\mathrm{U}}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 1 | - | 2357 | 14 | $4+\sqrt{14}$ | - | 7 | 35 | 1 | - | 2357 |
|  | 1 | 7 | 235 |  | $4+\sqrt{14}$ | 5 | 7 |  | $\sqrt{ } 35$ | - | 23 |
|  | $\sqrt{ } 2$ | - | 37 |  | $7+2 \sqrt{14}$ | - | 2 |  | $5+\sqrt{35}$ | - | 7 |
|  | $\sqrt{ } 2$ | 7 | 35 |  | $7+2 \sqrt{ } 14$ | 5 | 2 |  | $7+\sqrt{ } 35$ | - | 5 |
| 3 | 1 | - | 2357 | 15 | 1 | - | 2357 | 42 | 1 | - | 2357 |
|  | $\sqrt{ } 3$ | - | 27 |  | 1 | 7 | 235 |  | $\sqrt{42}$ | - | - |
|  | $1+\sqrt{3}$ | - | 3 |  | $\sqrt{ } 15$ | - | 2 |  | $6+\sqrt{42}$ | - | 57 |
|  | $3+\sqrt{3}$ | - | 5 |  | $\sqrt{ } 15$ | 7 | 2 |  | $7+\sqrt{42}$ | - | 3 |
| 5 | 2 | - | 2357 |  | $3+\sqrt{15}$ | - | 57 | 70 | 1 | - | 2357 |
|  | $2 \sqrt{5}$ | - | 237 |  | $3+\sqrt{15}$ | 7 | 5 |  | 1 | 3 | 257 |
| 6 | 1 | - | 2357 |  | $5+\sqrt{15}$ | - | 3 |  | $\sqrt{70}$ | - | - |
|  | 1 | 5 | 237 |  | $5+\sqrt{15}$ | 7 | 3 |  | $\sqrt{70}$ | 3 | - |
|  | $\checkmark 6$ | - | 57 |  | $1+\sqrt{15}{ }^{*}$ | 7 | 35 |  | $25+3 \sqrt{70}$ | - | 37 |
|  | $\sqrt{6}$ | 5 | 7 |  | $15+\sqrt{15}{ }^{*}$ | 7 | - |  | $25+3 \sqrt{70}$ | 3 | 7 |
|  | $2+\sqrt{6}$ | - | 3 |  | $6-\sqrt{15}{ }^{*}$ | 7 | 25 |  | $42+5 \sqrt{ } 70$ | - | 5 |
|  | $2+\sqrt{6}$ | 5 | 3 |  | $-5+2 \sqrt{ } 15^{*}$ | 7 | 23 |  | $42+5 \sqrt{70}$ | 3 | 5 |
|  | $3+\sqrt{6}$ | - | - | 21 | 2 | - | 2357 |  | $7+\sqrt{70}{ }^{*}$ | 3 | 5 |
|  | $3+\sqrt{6}$ | 5 | 2 |  | 2 | 5 | 237 |  | $10+\sqrt{70}{ }^{*}$ | 3 | 7 |
| 7 | 1 | - | 2357 |  | $2 \sqrt{21}$ | - | 25 |  | $-8+\sqrt{70}{ }^{*}$ | 3 | 57 |
|  | 1 | 3 | 257 |  | $2 \sqrt{21}$ | 5 | 2 |  | 35-4 $70{ }^{*}$ | 3 | 2 |
|  | $\sqrt{7}$ | - | 2 |  | $3+\sqrt{21}$ | - |  | 105 | 2 | - | 2357 |
|  | $\sqrt{7}$ | 3 | 25 |  | $3+\sqrt{21}$ | 5 | 27 |  | 2 | 2 | 357 |
|  | $3+\sqrt{7}$ | - | 7 |  | $7+\sqrt{21}$ | - | 23 |  | $2 \sqrt{105}$ | - | 2 |
|  | $3+\sqrt{7}$ | 3 | 57 |  | $7+\sqrt{21}$ | 5 | 23 |  | $2 \sqrt{105}$ | 2 | - |
|  | $7+3 \sqrt{7}$ | - | 35 | 30 | 1 | - | 2357 |  | $20+2 \sqrt{ } 105$ | - | 237 |
|  | $7+3 \sqrt{7}$ | 3 | 5 |  | 1 | 7 | 235 |  | $20+2 \sqrt{105}$ | 2 | 37 |
| 10 | 1 | - | 2357 |  | $\sqrt{ } 30$ | - | - |  | $42+4 \sqrt{ } 105$ | - | 25 |
|  | 1 | 3 | 257 |  | $\sqrt{ } 30$ | 7 | - |  | $42+4 \sqrt{105}$ | 2 | 5 |
|  | $\sqrt{10}$ | - | 37 |  | $5+\sqrt{30}$ | - | 37 |  | $7+\sqrt{105 *}$ | 2 | 35 |
|  | $\sqrt{ } 10$ | 3 | 7 |  | $5+\sqrt{30}$ | 7 | 3 |  | $15+\sqrt{105} *$ | 2 | 7 |
|  | $-2+\sqrt{10}{ }^{*}$ | 3 | 57 |  | $6+\sqrt{30}$ | - | 5 |  | $-9+\sqrt{105}$ | 2 | 57 |
|  | $5-\sqrt{10}{ }^{*}$ | 3 | 27 |  | $6+\sqrt{30}$ | 7 | 5 |  | 35-3V105* | 2 | 3 |
| 14 | 1 | - | 2357 |  | $3+\sqrt{30}{ }^{*}$ | 7 | 5 | 210 | 1 | - | 2357 |
|  | 1 | 5 | 237 |  | $10+\sqrt{30}$ * | 7 | 3 |  | $\checkmark 210$ | - | - |
|  | $\checkmark 14$ | - | 35 |  | $-4+\sqrt{30} *$ | 7 | 35 |  | $14+\sqrt{210}$ | - | 35 |
|  | $\sqrt{ } 14$ | 5 | 3 |  | 15-2V30* | 7 | 2 |  | $15+\sqrt{210}$ | - | 7 |

Table V．

| $N$ | NN | $N \sim$ | NN | mm | $\cdots$ | $\cdots$ | $\cdots \rightarrow$ | － | － | N | $N$ | － | $\rightarrow$ | $\cdots$ | $N$ | $\rightarrow$ | $\cdots$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| ${ }^{\text {un }}$ | $\rightarrow 0$ | OO | －0 | $\pm 0$ | 0 | 0 | 00 | 0 | 0 | 0 | 0 |  | 0 | $\bigcirc$ | 0 |  | 0 |
| $m$ | 00 | 00 | －10 | 00 | $\cdots$ | 0 | －10 | 0 | 0 | N | 0 | 0 | 0 | $\cdots$ | $\cdots$ | 0 | 0 |
| 1 N | － 0 | Mo | 00 | Mo | 0 | 0 | mo | N | － | － | 0 | － | － | $\cdots$ | － | 0 | 0 |
| A \％ | N は | $\stackrel{\infty}{\infty}$ | ヒึ้ำ | $\stackrel{\infty}{\infty} \underset{N}{\infty}$ | $N$ | $N$ | $\infty \infty$ | $N$ | $N$ | － | $\cdots$ | N | $N$ | M | $\cdots$ | $N$ | $N$ |
| $\mathrm{a}^{-1}$ | ${ }^{O H}$ | $\underset{H}{O H}$ | $0 \underset{\mathbf{N}}{\infty}$ | $0 \underset{r}{01}$ | ${ }^{\circ}$ | $m$ | Ot | 0 | $m$ | － | $\underset{\sim}{0}$ | 0 | $m$ | － | $\underset{\sim}{0}$ | 0 | $\cdots$ |
| 5 | Hrs | Hroo | $-1+$ | NNOO |  | OO | －0 |  | OO |  | －10 |  | 00 |  | －10 |  | －0 |
| $N$ | 0 | 00 | 0 | 00 | 0 | 0 | 0 | O |  | 0 | 0 | － |  | 0 | 0 | 0 | － |
| $\sim^{\text {Ln }}$ | N | N N | N | NN | $\square$ | N | $\cdots$ | $N$ |  | $\cdots$ | $\cdots$ | N |  | $m$ | $\cdots$ | $\cdots$ | $\cdots$ |
| $\mathrm{c}^{m}$ | － | $\cdots 0$ | 0 | 00 | $N$ | $N$ | 0 | $\rightarrow$ |  | 0 | － | － |  | 0 | 0 | 0 | 0 |
| $\\| \mathrm{e}^{N}$ | O | 00 | 0 | NO | 0 | 0 | 0 | $m$ |  | 0 | 0 | － |  | $\cdots$ | 0 | － | 0 |
| ค．a | $\frac{\ln }{n}$ | $\underset{n}{n} \quad \text { n }$ | $\underset{N}{n}$ | on | M | $\stackrel{I n}{n}$ | เก | $\xrightarrow{0}$ |  |  | แn |  |  | เก | ก | เn | กา |
| ${ }^{-1}$ | 0 | －N | － | orn | 0 | N | 0 | O |  | 0 | N | － |  | 0 | N | 0 | N |
| $\omega^{-1}$ | －10 | moor | ＋0 | H－100 |  | －-1 | 00 |  | 100 |  | 00 |  | 100 |  | 000 |  | 00 |
| $N$ | －10 | 00 | －0 | 00 | 0 | 0 | －0 | － |  | $\square$ | － | － | － | $\rightarrow$ | $n$ | － |  |
| cen | －10 | $\cdots 0$ | 00 | 00 | $\sim$ | $\cdots$ | －0 | $\rightarrow$ |  | － | － | 0 | － | 0 | － | － |  |
| $m$ | NN | N N | mm | N m | N | $m$ | NN | $N$ |  | $m$ | $m$ | $N$ | N | $\rightarrow$ | $N$ | N |  |
| 1 N | NO | 00 | けーロ | $\Rightarrow 0$ | 0 | 0 | －0 | $m$ |  | 0 | 0 | $m$ | $\rightarrow$ | $\cdots$ | 0 | N |  |
| \％${ }^{-1}$ | $\begin{gathered} \mathrm{N} \\ \mathrm{HN} \end{gathered}$ | $0 \quad 0$ | $\begin{array}{ll} \infty \\ m M \end{array}$ | M 0 | m | 0 | 00 | $\checkmark$ |  | 0 | $\sigma$ | $\sigma$ | $\sigma$ | $m$ | $a$ | $\cdots$ |  |
| $\square^{-1}$ | 00 | 0 － | $\underset{\Gamma}{\infty}$ | 0 － | 0 | $\pm$ | OM | 0 |  | 0 | $\checkmark$ | 0 | け | 0 | v | 0 |  |
| $5^{-1}$ | $\cdots \mathrm{ra}$ | －10－10 | NN | OONO |  | ON | $\rightarrow-1$ |  | 100 |  | OON |  | －9m | $\bigcirc$ | －10 |  | 400 |
|  | 0 | $\cdots$ | － | N | 0 |  | 0 | － |  | － |  | 0 |  | － |  | 0 |  |
| n | 0 | $\bigcirc$ | 0 | $\cdots$ | 0 |  | 0 | $\cdots$ |  | － |  | $\square$ |  | 0 |  | 0 |  |
|  | $r$ | 0 | $N$ | 0 | － |  | $\cdots$ | $\rightarrow$ |  | 0 |  | $\rightarrow$ |  | 0 |  | － |  |
| N | m | G | In | $\checkmark$ | $\cdots$ |  | N | $\checkmark$ |  | $\checkmark$ |  | $\pm$ |  | $N$ |  | $m$ |  |
| ค ه ${ }^{-1}$ | $\infty$ | $\infty$ | N | $\infty$ | \％ |  | 5 | 5 |  | $v$ |  | 0 |  | $N$ |  | $\pm$ |  |
|  | 0 | 0 | 0 | 0 | 0 |  | 0 | － |  | － |  | 0 |  | － |  | 0 |  |
| $\underbrace{-1}$ | No | Mroo | GN | mooo |  | noo | $\rightarrow 0$ |  | OOM |  | NoO |  | HNH |  | 000 |  | 400 |
| $\mathrm{F}^{\circ}$ | －O | $00-1$ | OO | $00-1 \overrightarrow{1}$ |  | $0-1$ | －0 |  | $0 \underset{1}{ }-1$ |  | $0-1$ |  | $0 \rightarrow \overrightarrow{1}$ |  | $0-1$ |  | $\overrightarrow{1}-1$ |
| ص | O－1 | OHmm | ON | OHnm |  | －-m | Or |  |  |  | － |  | Nmm |  |  |  | $\rightarrow$－1～ |
| a | －10 | momm | No | HONM |  | OMN | ＋o |  | OUN |  | OMIn |  | OMN | － | － 10 |  | OnN |
| ゅ | $\overrightarrow{1}$ | H－Arra | $\overrightarrow{1}$ | －r－rn |  | $\rightarrow$－1～ | $\overrightarrow{1} \overrightarrow{1}$ |  | $\rightarrow-1$ |  | H－1－4 |  | $\rightarrow-1$ |  | －hror |  | $\rightarrow-1 /$ |
| 4 | NN | Gくず | $\rightarrow$－ | $\begin{aligned} & 0000 \\ & \text { 4r } \end{aligned}$ |  | $\begin{gathered} 006 \\ -1-1-1 \end{gathered}$ | －o |  | MOM |  | $\infty \infty$ |  | パレก |  | NNN NNN |  | $\begin{gathered} \text { NNN } \\ \text {-in } \end{gathered}$ |
| － | NN | MMmm | เก⿺𠃊 | 00000 |  | NNN | $00$ |  | $\begin{aligned} & \text { GT } \\ & \text { HT } \end{aligned}$ |  | $\underset{\sim}{\text { un }}$ |  | NN |  | $0 \mathrm{MO}$ |  | กーกルก <br> クMM |

Table V. (cont.)
$(\alpha-a+b / D)$

Table VI.

$\left(i \in I_{U}^{*}\right)$

| D | $\alpha$ |  | $\mathrm{n}_{\pi}$ | $\mathrm{n}_{2}$ |  |  | $\mathrm{n}_{7}$ |  | $\mathrm{I}_{\mathrm{U}}$ |  | $I_{U}^{*}$ | N | $\kappa$ | $\mathrm{C}_{12}^{*}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 2 | 35 | 2 | 35 | 3 | 0 | 3. $190 \times 10^{28}$ |
|  | $\sqrt{ } 2$ | 0 | 0 | 1 | 0 | 0 | 0 | 3 | 5 | 2 | 35 | 2 | 0 | 3. $190 \times 10^{28}$ |
| 6 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 2 | 37 | 2 | 37 | 3 | 0 | 2. $712 \times 10^{26}$ |
|  | $\sqrt{ } 6$ | 0 | 0 | 1 | 1 | 0 | 0 | 7 |  | 2 | 7 | 2 | 0 | $4.604 \times 10^{22}$ |
|  | $2+\sqrt{6}$ | 1 | 0 | 1 | 0 | 0 | 0 | 3 |  | 2 | 3 | 2 | 0 | $2.090 \times 10^{22}$ |
|  | $3+\sqrt{6}$ | 1 | 0 | 0 | 1 | 0 | 0 | 2 |  | 2 | 3 | 3 | 0 | $2.090 \times 10^{22}$ |
| 7 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 2 | 57 | 2 | 57 | 2 | 0 | $1.065 \times 10^{30}$ |
|  | $\sqrt{7}$ | 0 | 0 | 0 | 0 | 0 | 1 | 2 | 5 | 2 | 5 | 2 | 0 | 2. $146 \times 10^{28}$ |
|  | $3+\sqrt{7}$ | 1 | 0 | 1 | 0 | 0 | 0 | 5 | 7 | 2 | 57 | 1 | 0 | $1.065 \times 10^{30}$ |
|  | $7+3 \sqrt{7}$ | 1 | 0 | 1 | 0 | 0 | 1 | 5 |  | 2 | 5 | 1 | 0 | 2. $146 \times 10^{25}$ |
| 10 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 2 | 57 | 2 | 57 | 3 | 0 | $3.214 \times 10^{29}$ |
|  | $\sqrt{10}$ | 0 | 0 | 1 | 0 | 1 | 0 | 7 |  | 2 | 7 | 2 | 0 | $8.414 \times 10^{24}$ |
|  | $-2+\sqrt{ } 10$ | -1 | 1 | 1 | 0 | 0 | 0 | 5 | 7 | 2 | 57 | 2 | 1 | $3.214 \times 10^{29}$ |
|  | $5-\sqrt{10}$ | -1 | 1 | 0 | 0 | 1 | 0 | 2 | 7 | 2 | 7 | 3 | 1 | $8.414 \times 10^{24}$ |
| 14 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 2 | 37 | 2 | 37 | 3 | 0 | $4.791 \times 10^{26}$ |
|  | $\checkmark 14$ | 0 | 0 | 1 | 0 | 0 | 1 | 3 |  | 2 | 3 | 2 | 0 | $4.347 \times 10^{22}$ |
|  | $4+\sqrt{14}$ | 1 | 0 | 1 | 0 | 0 | 0 | 7 |  | 2 | 7 | 2 | 0 | 8. $143 \times 10^{22}$ |
|  | $7+2 \sqrt{ } 14$ | 1 | 0 | 0 | 0 | 0 | 1 | 2 |  | 2 |  | 3 | 0 | 8. $371 \times 10^{18}$ |

Table VI. (cont.)

| D | $\alpha$ | $\mathrm{n}_{\varepsilon} \mathrm{n}_{\pi} \mathrm{n}_{2} \mathrm{n}_{3} \mathrm{n}_{5} \mathrm{n}_{7}$ |  |  |  |  |  | $\mathrm{I}_{\mathrm{U}}$ | $I_{U}^{*}$ | N | $\kappa$ | $\mathrm{C}_{12}^{*}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 15 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 235 | 235 | 2 | 0 | $2.144 \times 10^{28}$ |
|  | $\checkmark 15$ | 0 | 0 | 0 | 1 | 1 | 0 | 2 | 2 | 2 | 0 | 9. $427 \times 10^{19}$ |
|  | $3+\sqrt{15}$ | 1 | 0 | 1 | 1 | 0 | 0 | 5 | 25 | 1 | 0 | $1.694 \times 10^{24}$ |
|  | $5+\sqrt{15}$ | 1 | 0 | 1 | 0 | 1 | 0 | 3 | 23 | 1 | 0 | $1.035 \times 10^{24}$ |
|  | $1+\sqrt{15}$ | 0 | 1 | 1 | 0 | 0 | 0 | 35 | 235 | 1 | 1 | 2. $144 \times 10^{28}$ |
|  | $15+\sqrt{15}$ | 0 | 1 | 1 | 1 | 1 | 0 |  | 2 | 1 | 1 | 9. $427 \times 10^{19}$ |
|  | 6-V15 | -1 | 1 | 0 | 1 | 0 | 0 | 25 | 25 | 2 | 1 | $1.694 \times 10^{24}$ |
|  | $-5+2 \sqrt{ } 15$ | -1 | 1 | 0 | 0 | 1 | 0 | 23 | 23 | 2 | 1 | $1.035 \times 10^{24}$ |
| 21 | 2 | 0 | 0 | 2 | 0 | 0 | 0 | 237 | 237 | 1 | 0 | $1.898 \times 10^{26}$ |
|  | $2 \sqrt{21}$ | 0 | 0 | 2 | 1 | 0 | 1 | 2 | 2 | 0 | 0 | $2.640 \times 10^{18}$ |
|  | $3+\sqrt{21}$ | 1 | 0 | 2 | 1 | 0 | 0 | 27 | 27 | 1 | 0 | 3. $220 \times 10^{22}$ |
|  | $7+\sqrt{21}$ | 1 | 0 | 2 | 0 | 0 | 1 | 23 | 23 | 1 | 0 | 1. $435 \times 10^{22}$ |
| 30 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 235 | 235 | 3 | 0 | $4.141 \times 10^{28}$ |
|  | $\sqrt{ } 30$ | 0 | 0 | 1 | 1 | 1 | 0 |  | 2 | 2 | 0 | $2.022 \times 10^{20}$ |
|  | $5+\sqrt{30}$ | 1 | 0 | 0 | 0 | 1 | 0 | 3 | 23 | 3 | 0 | $2.217 \times 10^{24}$ |
|  | $6+\sqrt{30}$ | 1 | 0 | 1 | 1 | 0 | 0 | 5 | 25 | 2 | 0 | 3. $276 \times 10^{24}$ |
|  | $3+\sqrt{30}$ | 0 | 1 | 0 | 1 | 0 | 0 | 5 | 25 | 3 | 1 | $3.276 \times 10^{24}$ |
|  | $10+\sqrt{30}$ | 0 | 1 | 1 | 0 | 1 | 0 | 3 | 23 | 2 | 1 | $2.217 \times 10^{24}$ |
|  | $-4+\sqrt{30}$ | -1 | 1 | 1 | 0 | 0 | 0 | 35 | 235 | 2 | 1 | $4.141 \times 10^{28}$ |
|  | 15-2V30 | -1 | 1 | 0 | 1 | 1 | 0 | 2 | 2 | 3 | 1 | $2.022 \times 10^{20}$ |
| 70 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 257 | 257 | 3 | 0 | $3.229 \times 10^{30}$ |
|  | $\sqrt{ } 70$ | 0 | 0 | 1 | 0 | 1 | 1 |  | 2 | 2 | 0 | $2.115 \times 10^{21}$ |
|  | $25+3 \sqrt{70}$ | 1 | 0 | 0 | 0 | 1 | 0 | 7 | 27 | 3 | 0 | 8. $482 \times 10^{25}$ |
|  | $42+5 \sqrt{70}$ | 1 | 0 | 1 | 0 | 0 | 1 | 5 | 25 | 2 | 0 | $7.003 \times 10^{25}$ |
|  | $7+\sqrt{ } 70$ | 0 | 1 | 0 | 0 | 0 | 1 | 5 | 25 | 3 | 1 | $7.003 \times 10^{25}$ |
|  | $10+\sqrt{70}$ | 0 | 1 | 1 | 0 | 1 | 0 | 7 | 27 | 2 | 1 | 8. $482 \times 10^{25}$ |
|  | $-8+\sqrt{70}$ | -1 | 1 | 1 | 0 | 0 | 0 | 57 | 257 | 2 | 1 | $3.229 \times 10^{30}$ |
|  | 35-4V70 | -1 | 1 | 0 | 0 | 1 | 1 | 2 | 2 | 3 | 1 | $2.115 \times 10^{21}$ |
| 105 | 2 | 0 | 0 | 2 | 0 | 0 | 0 | 357 | 357 | 1 | 0 | $4.533 \times 10^{29}$ |
|  | $2 \sqrt{105}$ | 0 | 0 | 2 | 1 | 1 | 1 |  |  | 0 | 0 | $4.295 \times 10^{16}$ |
|  | $20+2 \sqrt{ } 105$ | 1 | 0 | 2 | 0 | 1 | 0 | 37 | 37 | 1 | 0 | $1.690 \times 10^{25}$ |
|  | $42+4 \sqrt{105}$ | 1 | 0 | 2 | 1 | 0 | 1 | 5 | 5 | 1 | 0 | $8.655 \times 10^{20}$ |
|  | $7+\sqrt{105}$ | 0 | 1 | 2 | 0 | 0 | 1 | 35 | 35 | 1 | 1 | $1.396 \times 10^{25}$ |
|  | $15+\sqrt{105}$ | 0 | 1 | 2 | 1 | 1 | 0 | 7 | 7 | 1 | 1 | $1.049 \times 10^{21}$ |
|  | $-9+\sqrt{ } 105$ | -1 | 1 | 2 | 1 | 0 | 0 | 57 | 57 | 1 | 1 | 2. $485 \times 10^{25}$ |
|  | 35-3V105 | -1 | 1 | 2 | 0 | 1 | 1 | 3 | 3 | 1 | 1 | 5. $880 \times 10^{20}$ |

## Chapter 8. The Thue equation.

Acknowledgements. The research for this chapter has been done in cooperation with N. Tzanakis from Iraklion. The results have been published in Tzanakis and de Weger [1989 ${ }^{\text {a }}$.

### 8.1. Introduction.

Let $F(X, Y) \in \mathbb{Z}[X, Y]$ be a binary form with integral coefficients, of degree at least three, and irreducible. Let $m$ be a nonzero integer. The diophantine equation

$$
F(X, Y)=m
$$

in $X, Y \in \mathbb{Z}$ is called a Thue equation. It plays a central role in the theory of diophantine equations. In 1909 Thue proved that it has only finitely many solutions (cf. Thue [1909]). His proof was ineffective. An effective proof was given by Baker [1968]. See Chapter 5 of Shorey and Tijdeman [1986] for a survey of results on Thue equations. By using Lemma 2.4 in Baker's argument, we derive a fully explicit upper bound for the solutions of the Thue equation. Then we show how the methods developed in Chapter 3 can be used to actually find all the solutions of a Thue equation. Our method works in principle for any Thue equation, and in practice for any Thue equation of not too large degree, provided that some algebraic data on the form $F$ are available. See also Tzanakis [1989] for a short introduction.

Variants of the method we use here have been used in practice to solve Thue equations by Ellison, Ellison, Pesek, Stahl and Stall [1975], Steiner [1986], Pethö and Schulenberg [1987], and Blass, Glass, Meronk and Steiner [1987a ${ }^{\mathrm{a}}$, [1987 ${ }^{\text {b }] . ~ I n ~ a l l ~ t h e s e ~ c a s e s ~} m=1$, whereas de Weger [1989 ${ }^{\text {b }}$ ] treats an example with $m>1$, using the method described in this chapter. When determining all cubes in the Fibonacci sequence, Pethö [1983] solved a Thue equation by the Gelfond-Baker method, but with a completely different way to find all the solutions below the upper bound. And there are numerous Thue equations that have been solved by different (usually ad hoc) methods.

### 8.2. From the Thue equation to a linear form in logarithms.

In this section we show how the general Thue equation leads to an inequality involving a linear form in the logarithms of algebraic numbers with rational integral coefficients (unknowns). Let

$$
F(X, Y)=\sum_{i=0}^{n} f_{i} \cdot X^{n-i} \cdot Y^{i} \in \mathbb{Z}[X, Y]
$$

be a binary form of degree $n \geqslant 3$ and let $m$ be a nonzero integer. Consider the Thue equation

$$
\begin{equation*}
F(X, Y)=m, \tag{8.1}
\end{equation*}
$$

in the unknowns $X, Y \in \mathbb{Z}$. If $F$ is reducible over $\mathbb{Q}$, then (8.1) can be reduced to a system of finitely many equations of type (8.1) with irreducible binary forms. For such equations of degree 1 or 2 it is well known how to determine the solutions. Therefore we may assume from now on that $F$ is irreducible over $\mathbb{Q}$ and of degree $\geqslant 3$. Let $g(x)=F(x, 1)$. If $g(x)=0$ has no real roots then one can trivially find small upper bounds for $\max (|X|,|Y|)$ for the solutions (X,Y) of (8.1). Therefore, throughout this chapter we assume that the algebraic equation $g(x)=0$ has at least one real root. We number its roots as follows: $\xi^{(1)}, \ldots, \xi^{(s)}$ (with $s \geqslant 1$ ) are the real roots and $\xi^{(s+1)}=\overline{\xi^{(s+t+1)}}, \ldots, \xi^{(s+t)}=\overline{\xi^{(s+2 t)}}$ are the non-real roots, so that we have $t(\geqslant 0)$ pairs of complex-conjugate roots, and $s+2 \cdot t=n$.

Consider the field $K=\mathbb{Q}(\xi)$, where $g(\xi)=0$. We will define three positive real numbers $Y_{1}<Y_{2}<Y_{3}$, that will divide the set of possible solutions ( $\mathrm{X}, \mathrm{Y}$ ) of (8.1) into four classes:
$\rightarrow$ the 'very small' solutions, with $|Y| \leqslant Y_{1}$. They will be found by enumeration of all possibilities,
$\rightarrow$ the 'small' solutions, with $Y_{1}<|Y| \leqslant Y_{2}$. They will be found by evaluating the continued fraction expansions of the real roots $\xi^{(i)}$.
$\rightarrow$ the 'large' solutions, with $Y_{2}<|Y| \leqslant Y_{3}$. They will be proved not to exist by a computational diophantine approximation technique,
$\rightarrow$ the 'very large' solutions, with $|Y|>Y_{3}$. They will be proved not to exist by the theory of linear forms in logarithms.

The value of $Y_{3}$ follows from the Gelfond-Baker theory of linear forms in logarithms. The value of $Y_{2}$ follows from the restrictions that we use as we
try to prove that no 'large' solutions exist. The value of $Y_{1}$ follows from Lemma 8.1 below. This lemma shows that if $|\mathrm{Y}|$ is large enough then $\mathrm{X} / \mathrm{Y}$ is 'extremely close' to one of the real roots $\xi^{(i)}$. In a typical example $Y_{3}$ may be as large as $10^{10^{50}}, Y_{2}$ as $10^{10}$, and $Y_{1}$ as small as 10 .

LEMMA 8.1. Let $X, Y \in \mathbb{Z}$ satisfy (8.1). Put $\beta=X-\xi \cdot Y$,

$$
\begin{aligned}
& Y_{0}=\left\{\begin{array}{ll}
\left.\left(\frac{2^{n-1} \cdot|m|}{\min _{1 \leqslant i \leqslant t}\left|g^{\prime}\left(\xi^{(s+i)}\right)\right| \cdot \min _{1 \leqslant i \leqslant t}\left|\operatorname{Im} \xi^{(s+i)}\right|}\right)^{1 / n}\right] & \text { if } t \geqslant 1 \\
1 & \text { if } t=0
\end{array},\right. \\
& C_{1}=\frac{2^{n-1} \cdot|m|}{\min _{1 \leqslant i \leqslant s}\left|g^{\prime}\left(\xi^{(i)}\right)\right|}, \quad C_{2}=\frac{1}{2} . \min _{1 \leqslant i<j \leqslant n}\left|\xi^{(i)}-\xi^{(j)}\right|, \\
& Y_{1}=\max \left(Y_{0},\left\lceil\left(4 \cdot C_{1}\right)^{1 /(n-2)}\right\rceil\right) .
\end{aligned}
$$

(i). If $|\mathrm{Y}|>\mathrm{Y}_{0}$ then there exists an $\mathrm{i}_{0} \in\{1, \ldots, \mathrm{~s}\}$ such that

$$
\begin{aligned}
& \left|\beta^{\left(i_{0}\right)}\right| \leqslant C_{1} \cdot|Y|^{-(n-1)}, \\
& \left|\beta^{(i)}\right| \geqslant C_{2} \cdot|Y| \text { for } i \in\{1, \ldots, n\}, i \neq i_{0} .
\end{aligned}
$$

(ii). If $|\mathrm{Y}|>\mathrm{Y}_{1}$ then $\mathrm{X} / \mathrm{Y}$ is a convergent from the continued fraction expansion of $\xi^{\left(\mathrm{i}_{0}\right)}$.

Proof. Let $i_{0} \in\{1, \ldots, n\}$ be such that $\left|\beta^{\left(i_{0}\right)}\right|=\min _{1 \leqslant i \leqslant n}\left|\beta^{(i)}\right|$. We have from (8.1)

$$
\left|f_{0}\right| \cdot \prod_{i=1}^{n}\left|\beta^{(i)}\right|=|m|
$$

By the minimality of $\mid \beta^{\left(i_{0}\right)}$ | we have for all $i$

$$
|Y| \cdot\left|\xi^{(i)}-\xi^{\left(i_{0}\right)}\right|=\left|\beta^{(i)}-\beta^{\left(i_{0}\right)}\right| \leqslant\left|\beta^{(i)}\right|+\left|\beta^{\left(i_{0}\right)}\right| \leqslant 2 \cdot\left|\beta^{(i)}\right| .
$$

Hence $\left|\beta^{(i)}\right| \geqslant C_{2} \cdot|Y|$. Further,

$$
\left|\beta^{\left(i_{0}\right)}\right|=\frac{|m|}{\left|f_{0}\right|} \cdot \prod_{i \neq i_{0}}\left|\beta^{(i)}\right|^{-1} \leqslant \frac{|m|}{\left|f_{0}\right|} \cdot \prod_{i \neq i_{0}}\left(\frac{1}{2} \cdot|Y| \cdot\left|\xi^{(i)}-\xi^{\left(i_{0}\right)}\right|\right)^{-1}
$$

$$
=\frac{2^{n-1} \cdot|m|}{\left|f_{0} \cdot \prod_{i \neq i_{0}}\left(\xi^{(i)}-\xi^{\left(i_{0}\right)}\right)\right| \cdot|Y|^{n-1}}=\frac{2^{n-1} \cdot|m|}{\left|g^{\prime}\left(\xi^{\left(i_{0}\right)}\right)\right| \cdot|Y|^{n-1}} .
$$

Now, if $i_{0}>s$ (and hence $t \geqslant 1$ ) then, by the definition of $Y_{0}$,

$$
\begin{aligned}
\left|\frac{X}{Y}-\xi^{\left(i_{0}\right)}\right| & =\frac{\left|\beta^{\left(i_{0}\right)}\right|}{|Y|} \leqslant \frac{2^{n-1} \cdot|m|}{\left|g^{\prime}\left(\xi^{\left(i_{0}\right)^{\prime}}\right)\right|} \cdot|Y|^{-n} \\
& \leqslant\left(\frac{Y_{0}}{|Y|}\right)^{n} \cdot \min _{s+1 \leqslant i \leqslant s+t}\left|\operatorname{Im} \xi^{(i)}\right|,
\end{aligned}
$$

which is impossible if $|Y|>Y_{0}$. Hence $i_{0} \leqslant s$, and now (i) follows at once. Moreover, if $|Y|>Y_{1}$, then

$$
\left|\frac{X}{Y}-\xi^{\left(\mathrm{i}_{0}\right)}\right|=\left|\beta^{\left(\mathrm{i}_{0}\right)}\right| \cdot|Y|^{-1} \leqslant C_{1} \cdot|Y|^{-n} \leqslant \frac{1}{4} \cdot Y_{1}^{\mathrm{n}-2} \cdot|Y|^{-\mathrm{n}} \leqslant \frac{1}{2} \cdot|Y|^{-2},
$$

and thus $\left|\frac{X}{Y}-\xi^{\left(i_{0}\right)}\right|<\frac{1}{2} \cdot|Y|^{-2}$, since $\xi^{\left(i_{0}\right)}$ is irrational. Now follows from a well known result on continued fractions, cf. (3.6).

Now let $|Y|>Y_{1}$ and $i_{0} \in\{1, \ldots, s\}$ as in Lemma 8.1. Choose $j, k \in\{1, \ldots, n\}$ such that $i_{0}, j, k$ are pairwise distinct and either $j, k \in\{1, \ldots, s\}$ or $j+t=k$ (so that $\xi^{(k)}=\overline{\xi^{(j)}}$ ), but further the choice of $j, k$ is free. By $\beta^{(i)}=X-Y \cdot \xi^{(i)}$ for $i=i_{0}, j, k$ we get, on eliminating the $X$ and $Y$,

$$
\beta^{\left(i_{0}\right)} \cdot\left(\xi^{(j)}-\xi^{(k)}\right)+\beta^{(j)} \cdot\left(\xi^{(k)}-\xi^{\left(i_{0}\right)}\right)+\beta^{(k)} \cdot\left(\xi^{\left(i_{0}\right)}-\xi^{(j)}\right)=0
$$

or, equivalently,

$$
\begin{equation*}
\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}-\xi^{(k)}} \cdot \frac{\beta^{(k)}}{\beta^{(j)}}-1=-\frac{\xi^{(k)}-\xi^{(j)}}{\xi^{(k)}-\xi^{\left(i_{0}\right)}} \cdot \frac{\beta^{\left(i_{0}\right)}}{\beta^{(j)}} . \tag{8.2}
\end{equation*}
$$

By Lemma 8.1, the right hand side of (8.2) is 'extremely small'. Put, if $j, k \in\{1, \ldots, s\}$ (let us call it 'the real case')

$$
\Lambda=\log \left|\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}-\xi^{(k)}} \cdot \frac{\beta^{(k)}}{\beta^{(j)}}\right|
$$

and if $j, k \in\{s+1, \ldots, s+2 \cdot t\}$ (let us call it 'the complex case')

$$
\Lambda=\frac{1}{i} \cdot \log \left(\frac{\xi^{\left(\mathrm{i}_{0}\right)_{-\xi^{(j)}}^{(j)}}}{\xi^{\left(\mathrm{i}_{0}\right)_{-\xi^{(k)}}^{(k)}}} \cdot \frac{\beta^{(\mathrm{k})}}{\beta^{(\mathrm{j})}}\right),
$$

where, in general, for $z \in \mathbb{C}$, $\log (z)$ denotes the principal value of the logarithm of $z$ (hence $-\pi<\operatorname{Im} \log (z) \leqslant \pi$ ). By $\xi^{(k)}=\overline{\xi^{(j)}}$ we have $\Lambda \in \mathbb{R}$ and $|\Lambda| \leqslant \pi$.

The following lemma shows how small $|\Lambda|$ is.

LEMMA 8.2. Put

$$
\begin{aligned}
& C_{3}=\max _{i_{1} \neq i_{2} \neq i_{3} \neq i_{1}}\left|\frac{\xi^{\left(i_{1}\right)}-\xi^{\left(i_{2}\right)}}{\xi^{\left(i_{1}\right)}-\xi^{\left(i_{3}\right)}}\right| \\
& Y_{2}^{*}=\max \left(Y_{1},\left\lceil\left(2 \cdot C_{1} \cdot C_{3} / C_{2}\right)^{1 / n} \mid\right) .\right.
\end{aligned}
$$

If $|\mathrm{Y}|>\mathrm{Y}_{2}^{*}$ then

$$
|\Lambda|<\frac{1.39 \cdot C_{1} \cdot C_{3}}{C_{2}} \cdot|Y|^{-n} .
$$

Proof. Consider first the real case. From $|Y|>Y_{2}^{*}$ and Lemma 8.1 it follows that the right hand side of (8.2) is absolutely less than $\frac{1}{2}$ and, consequently,

$$
\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}-\xi^{(k)}} \cdot \frac{\beta^{(k)}}{\beta^{(j)}}>0
$$

It follows that the left hand side of (8.2) is equal to $e^{\Lambda}-1$, and now (8.2) implies, in view of Lemma 8.1 and the definition of $C_{3}$,

$$
\left|e^{\Lambda}-1\right|<C_{3} \cdot \frac{C_{1} \cdot|Y|^{-(n-1)}}{C_{2} \cdot|Y|}=\frac{C_{1} \cdot C_{3}}{C_{2}} \cdot|Y|^{-n} .
$$

On the other hand, $\left|e^{\Lambda}-1\right|<\frac{1}{2}$ implies (cf. Lemma 2.2)

$$
|\Lambda| \leqslant 2 \cdot \log 2 \cdot\left|e^{\Lambda}-1\right| \leqslant 1.39 \cdot\left|e^{\Lambda}-1\right|,
$$

which proves our claim in the real case.
In the complex case the left hand side of (8.2) is equal to $e^{i \Lambda}-1$, and, as in the real case, we derive

$$
\left|e^{\mathrm{i} \Lambda}-1\right|<\frac{\mathrm{C}_{1} \cdot \mathrm{C}_{3}}{\mathrm{C}_{2}} \cdot|Y|^{-\mathrm{n}}<\frac{1}{2}
$$

Since $\left|e^{i \Lambda}-1\right|=2 \cdot|\sin \Lambda / 2|$, it follows that $|\sin \Lambda / 2|<\frac{1}{4}$, and therefore by Lemma 2.3

$$
|\Lambda| \leqslant 2 \cdot \frac{1 / 4}{\sin 1 / 4} \cdot|\sin \Lambda / 2|=\frac{1 / 4}{\sin 1 / 4} \cdot\left|e^{i \Lambda}-1\right| \leqslant 1.02 \cdot\left|e^{i \Lambda}-1\right|
$$

which proves the lemma in the complex case.

In the ring of integers of the field $K$ (as well as in any other order $R$ of $K$ ) there exists a system of fundamental units $\varepsilon_{1}, \ldots, \varepsilon_{r}$, where $r=s+t-1$ (Dirichlet's Unit Theorem). Note that since $F$ is irreducible and we have supposed $s>0$, the only roots of unity belonging to $K$ are $\pm 1$. We shall not discuss here the problem of finding such a system (for efficient methods see e.g. Berwick [1932], Billevic [1956], [1964], Pohst and Zassenhaus [1982], Buchmann [1985], [1986]). We simply assume that a system of fundamental units is known. On the other hand, there exist only finitely many non-associates $\mu_{1}, \ldots, \mu_{\nu}$ in $K$ such that $f_{0} \cdot N\left(\mu_{i}\right)=m$ for i $=1, \ldots, v$ (we use $N(\cdot)$ to denote the norm of the extension $K / \mathbb{Q}$ ). We also assume that a complete set of such $\mu_{i}$ 's is known. Let $M$ be the set of all $\zeta \cdot \mu_{i}$, where $\zeta$ is a root of unity in $K$. (In the important case $\left|f_{0}\right|=|m|=1$, it is clear that $M=\{-1,1\}$ ). Then, for any integral solution ( $\mathrm{X}, \mathrm{Y}$ ) of (8.1) there exist some $\mu \in M$ and $a_{1}, \ldots, a_{r} \in \mathbb{Z}$, such that

$$
\beta=\mu \cdot \varepsilon_{1}{ }_{1}{ }^{a_{1}} \cdot \ldots \cdot \varepsilon_{r}{ }^{a_{r}} .
$$

Thus, the initial problem of solving (8.1) is reduced to that of finding all integral $r$-tuples $\left(a_{1}, \ldots, a_{r}\right)$ such that $\mu \cdot \varepsilon_{1}{ }_{1}{ }^{{ }_{1}} \ldots \cdot \varepsilon_{r}{ }_{r}$ for some $\mu \in M$ be of the special shape $X-Y \cdot \xi$, with $X, Y \in \mathbb{Z}$. As we have seen, $X$ and $Y$ can be eliminated, so that we obtain (8.2). Thus the problem reduces to solving finitely many equations of the type

$$
\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}-\xi^{(k)}} \cdot \frac{\mu^{(k)}}{\mu^{(j)}} \cdot \prod_{i=1}^{r}\left(\frac{\varepsilon_{i}^{(k)}}{\varepsilon_{i}^{(j)}}\right)^{a}-1=-\frac{\xi^{(k)}-\xi^{(j)}}{\xi^{(k)}-\xi^{\left(i_{0}\right)}} \cdot \frac{\mu^{\left(i_{0}\right)}}{\mu^{(j)}} \cdot \prod_{i=1}^{r}\left(\frac{\varepsilon_{i}^{\left(i_{0}\right)}}{\varepsilon_{i}^{(j)}}\right)^{a}
$$

(the so-called 'unit equation'). In the real case we have

$$
\begin{equation*}
\Lambda=\log \left|\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}-\xi^{(k)}} \cdot \frac{\mu^{(k)}}{\mu^{(j)}}\right|+\sum_{i=1}^{r} a_{i} \cdot \log \left|\frac{\varepsilon_{i}^{(k)}}{\varepsilon_{i}^{(j)}}\right| \tag{8.3}
\end{equation*}
$$

and in the complex case

$$
\begin{equation*}
\Lambda=\operatorname{Arg}\left(\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}-\xi^{(k)}} \cdot \frac{\mu^{(k)}}{\mu^{(j)}}\right)+\sum_{i=1}^{r} a_{i} \cdot \operatorname{Arg}\left(\frac{\varepsilon_{i}^{(k)}}{\varepsilon_{i}^{(j)}}\right)+a_{0} \cdot 2 \pi \tag{8.4}
\end{equation*}
$$

with $\mathrm{a}_{0} \in \mathbb{Z}$, and $-\pi<\operatorname{Arg}(z) \leqslant \pi$ for every $z \in \mathbb{C}$. Note that $\Lambda$ in the real case, and i• $\Lambda$ in the complex case, is a linear form in (principal) logarithms of algebraic numbers, where the coefficients $a_{i}$ are integers. The Gelfond-Baker theory provides an explicit lower bound for $|\Lambda|$ in terms of maxla ${ }_{i}$. Using this in combination with Lemma 8.2 we can find an explicit upper bound for maxla, ${ }_{i}$. This is what we do in the next section.

### 8.3. Upper bounds.

Let $A=\max _{1 \leqslant i \leqslant r}\left|a_{i}\right|$. First we find an upper bound for $A$ in terms of $|Y|$.

LEMMA 8.3. Let $\mathrm{I}=\left\{\mathrm{h}_{1}, \ldots, \mathrm{~h}_{\mathrm{r}}\right\} \subset\{1, \ldots, \mathrm{n}\}$. Put

$$
U_{I}=\left(\log \left|\varepsilon_{\ell}^{\left(h_{i}\right)}\right|\right)_{1 \leqslant i \leqslant r, 1 \leqslant \ell \leqslant r}
$$

(where $i$ indicates a row and $\ell$ a column of the matrix),

$$
U_{I}^{-1}=\left(u_{i \ell}\right), \quad N\left[U_{I}^{-1}\right]=\max _{1 \leqslant i \leqslant r} \sum_{\ell=1}^{r}\left|u_{i \ell}\right|
$$

Put also

$$
\begin{aligned}
& \mu_{-}=\min _{\substack{1 \leqslant i \leqslant n \\
\mu \in M}}\left|\mu^{(i)}\right|, \quad \mu_{+}=\max _{\substack{1 \leqslant i \leqslant n \\
\mu \in M}}\left|\mu^{(i)}\right|, \\
& C_{4}=\frac{\left.\frac{1}{2}+\max _{1 \leqslant i_{1}<i_{2} \leqslant n} \right\rvert\, \xi^{\left(i_{1}\right)}{ }_{\left.-\xi^{\left(i_{2}\right.}\right)}}{\mu_{-}} \\
& C_{5}=\min \left((n-1) \cdot \min _{I} N\left[U_{I}^{-1}\right], \max _{I} N\left[U_{I}^{-1}\right]\right)
\end{aligned}
$$

Then, for

$$
|Y|>\max \left(Y_{1}, 2 \cdot|m|^{1 / n}, \mu_{+} / C_{2}\right),
$$

we have

$$
A<C_{5} \cdot \log \left(C_{4} \cdot|Y|\right)
$$

Proof. By $\beta=\mu \cdot \varepsilon_{1}{ }^{1} \cdot \ldots \cdot \varepsilon_{r}{ }^{a_{r}}$ we have

$$
\left(\begin{array}{cc}
\log ^{\left(h_{1}\right)}{ }_{/ \mu}\left(h_{1}\right)^{\prime}  \tag{8.5}\\
\vdots & \\
\log \mid \beta^{\left(h_{r}\right)} / \mu & \left(h_{r}\right) \\
&
\end{array}\right)=U_{I} \cdot\left(\begin{array}{l}
a_{1} \\
\vdots \\
a_{r}
\end{array}\right) .
$$

On the other hand, for every $h \in\{1, \ldots, n\}$, using the end of the proof of Lemma 8.1,

$$
\begin{aligned}
\left|\beta^{(h)}\right|= & \left|X-Y \cdot \xi^{(h)}\right| \leqslant\left|X-Y \cdot \xi^{\left(i_{0}\right)}\right|+|Y| \cdot\left|\xi^{\left(i_{0}\right)}-\xi^{(h)}\right| \\
& \leqslant \frac{1}{2 \cdot|Y|}+|Y| \cdot\left|\xi^{\left(i_{0}\right)}-\xi^{(h)}\right| \\
& <\left(\frac{1}{2}+\max _{1 \leqslant i_{1}<i_{2} \leqslant n}\left|\xi^{\left(i_{1}\right)}-\xi^{\left(i_{2}\right)}\right|\right) \cdot|Y|
\end{aligned}
$$

and therefore

$$
\left|\frac{\beta^{(h)}}{\mu(h)}\right|<C_{4} \cdot|Y| \text { for } h=1, \ldots, n .
$$

Note that $\mathrm{C}_{4} \cdot|\mathrm{Y}|>1$. Indeed, by

$$
\prod_{i=1}^{n}\left|\mu^{(i)}\right|=\frac{|m|}{\left|f_{0}\right|} \leqslant|m|
$$



$$
C_{4} \cdot|Y| \geqslant\left(\frac{1}{2}+\max _{1 \leqslant i_{1}<i_{2} \leqslant n}\left|\xi^{\left(i_{1}\right)}-\xi^{\left(i_{2}\right)}\right|\right) \cdot|Y| \cdot|m|^{-1 / n}>\frac{|Y|}{2|m|^{1 / n}}>1 .
$$

Then,

$$
\begin{equation*}
\log \left|\frac{\beta^{(h)}}{\mu}\right|<\log \left(C_{4} \cdot|Y|\right) \text { for } h=1, \ldots, n, \quad \log \left(C_{4} \cdot|Y|\right)>0 \tag{8.6}
\end{equation*}
$$

Next we show that

$$
\begin{equation*}
|\log | \frac{\beta^{(i)}}{\mu^{(i)}}\left|\mid<(n-1) \cdot \log \left(C_{4} \cdot|Y|\right) \quad \text { for } \quad i=1, \ldots, n\right. \tag{8.7}
\end{equation*}
$$

Indeed, in view of (8.6), a stronger inequality is true if $\left|\beta^{(i)} / \mu^{(i)}\right| \geqslant 1$. Suppose now that $\left|\beta^{(i)} / \mu^{(i)}\right|<1$. By

$$
\prod_{h=1}^{n}\left|\frac{\beta^{(h)}}{\mu^{(h)}}\right|=1
$$

it follows that

$$
|\log | \frac{\beta^{(i)}}{\mu^{(i)}}\left||=-\log | \frac{\beta^{(i)}}{\mu^{(i)}}\right|=\sum_{\substack{h=1 \\ h \neq i}}^{n} \log \left|\frac{\beta^{(h)}}{\mu^{(h)}}\right|<(n-1) \cdot \log \left(C_{4} \cdot|Y|\right)
$$

in view of (8.6). Now the inequality

$$
A<(n-1) \cdot \min _{I} N\left[U_{I}^{-1}\right] \cdot \log \left(C_{4} \cdot|Y|\right)
$$

follows from (8.5), (8.7), the definition of $N\left[U_{I}^{-1}\right]$ and the fact that, as we have not put so far any restriction on $I$, this could be chosen so that $\mathrm{N}\left[\mathrm{U}_{\mathrm{I}}^{-1}\right]$ be minimal. It remains to show that

$$
A<\max _{\mathrm{I}} \mathrm{~N}\left[\mathrm{U}_{\mathrm{I}}^{-1}\right] \cdot \log \left(\mathrm{C}_{4} \cdot|\mathrm{Y}|\right)
$$

Choose I such that $i_{0} \notin \mathrm{I}$. Then, by Lemma 8.1, for every $h \in I$, $\left|\beta^{(h)} / \mu^{(h)}\right|>C_{2} \cdot|Y| / \mu_{+}>1$ and now, in view of (8.6),

$$
|\log | \frac{\beta^{(h)}}{\mu^{(h)}}\left|\mid<\log \left(C_{4} \cdot|Y|\right)\right.
$$

which implies our assertion.

Lemmas 8.2 and 8.3 immediately yield

LEMMA 8.4. Put

$$
\mathrm{C}_{6}=\frac{1.39 \cdot \mathrm{C}_{1} \cdot \mathrm{C}_{3} \cdot \mathrm{C}_{4}^{\mathrm{n}}}{\mathrm{C}_{2}}, \quad \mathrm{Y}_{2}^{\prime}=\max \left(\mathrm{Y}_{2}^{*}, 2 \cdot|\mathrm{~m}|^{1 / \mathrm{n}}, \mu_{+} / \mathrm{C}_{2}\right)
$$

If $|Y|>Y_{2}^{\prime}$ then
$|\Lambda|<C_{6} \cdot \exp \left(\frac{-n}{C_{5}} \cdot A\right)$.

Next we apply Lemma 2.4 (Waldschmidt). It yields in the real case (assuming that $\Lambda \neq 0$ )

$$
\begin{equation*}
|\Lambda|>\exp \left(-C_{7} \cdot\left(\log A+C_{8}\right)\right) \tag{8.8}
\end{equation*}
$$

and in the complex case this holds when $A$ is replaced by $A^{\prime}=\max _{0 \leqslant i \leqslant r}\left|a_{i}\right|$. The precise values for $C_{7}$ and $C_{8}$ are given in Section 2.3. It should be noted that in the complex case $a_{0}$ makes now its appearance, while it was not present in Lemmas 8.3 and 8.4. In order to obtain an upper bound for A , we must find an upper bound for $A$, in terms of $A$. Indeed, using

$$
\operatorname{Arg}\left(z_{1} \cdot z_{2}\right)=\operatorname{Arg}\left(z_{1}\right)+\operatorname{Arg}\left(z_{2}\right)+k \cdot 2 \pi, \quad k \in\{-1,0,1\}
$$

we find from (8.4) and the proof of lemma 8.2 that if $A \geqslant 2$ then

$$
\left|a_{0}\right|<\frac{1}{2}+\frac{1}{2} \cdot r \cdot A+|\Lambda| / 2 \pi<1+r \cdot A \leqslant r \cdot A
$$

Thus we may apply (8.8) in both cases with the same $A$ if we replace $C_{8}$ by

$$
\begin{array}{ll}
\mathrm{C}_{8}^{\prime}=\mathrm{C}_{8} & \text { in the real case }, \\
\mathrm{C}_{8}^{\prime}=\mathrm{C}_{8}+\log \mathrm{r} & \text { in the complex case. }
\end{array}
$$

We can now give an upper bound for A.

LEMMA 8.5. Put

$$
C_{9}=\frac{2 \cdot C_{5}}{n} \cdot\left(\log C_{6}+C_{7} \cdot C_{8}^{\prime}+C_{7} \cdot \log \frac{C_{5} \cdot C_{7}}{n}\right)
$$

If $|Y|>Y_{2}^{\prime}$, then $A<C_{9}$.

Proof. As we have seen in the proof of Lemma 8.2, $\left|e^{\Lambda}-1\right|<\frac{1}{2}$ in the real case, and $\left|e^{i \Lambda}-1\right|<\frac{1}{2}$ in the complex case. Note that $\beta^{\left(i_{0}\right)} \neq 0$. Hence (8.2) implies $\Lambda \neq 0$. Therefore Lemma 8.4 and (8.8) yield

$$
A<\frac{C_{5}}{n} \cdot\left(\log C_{6}+C_{7} \cdot C_{8}^{\prime}+C_{7} \cdot \log A\right)
$$

The result now follows from Lemma 2.1.

Remark. From this upper bound for $A$ an upper bound for $|Y|$ can be derived, thus a value for $Y_{3}$ (cf. Section 8.2). We shall not do this. Note


### 8.4. Reducing the upper bound.

We are now left with a problem of the following type. Let be given real numbers $\delta, \mu_{1}, \ldots, \mu_{\mathrm{q}}(\mathrm{q} \geqslant 2$, the case $\mathrm{q}=1$ is trivial). Write

$$
\Lambda=\delta+a_{1} \cdot \mu_{1}+\ldots+a_{\mathrm{q}} \cdot \mu_{\mathrm{q}}
$$

where the $a_{i}$ 's belong to $\mathbb{Z}$, and put $A=\max _{1 \leqslant i \leqslant q}\left|a_{i}\right|$. If $K_{1}, K_{2}, K_{3}$ are given positive numbers, then find all $q$-tuples $\left(a_{1}, \ldots, a_{q}\right) \in \mathbb{Z}^{q}$ satisfying

$$
\begin{equation*}
|\Lambda|<K_{1} \cdot \exp \left(-K_{2} \cdot A\right), \quad A<K_{3} \tag{8.9}
\end{equation*}
$$

In our case, it follows from (8.3) or (8.4) how to define $q$, $\delta$ and the $\mu_{i} ' s$, and from Lemmas 8.4 and 8.5 how to define $K_{1}, K_{2}, K_{3}$. In general, $K_{1}$ and $K_{2}$ are 'small' constants, whereas $K_{3}$ is 'very large'. Put

$$
\Lambda_{0}=\mathrm{a}_{1} \cdot \mu_{1}+\ldots+\mathrm{a}_{\mathrm{q}} \cdot \mu_{\mathrm{q}}
$$

so that $\Lambda=\delta+\Lambda_{0}$. We apply the methods of Chapter 3 to problem (8.9).

Below we distinguish three cases. In the first two we suppose that the $\mu_{i}$ 's are $\mathbb{Q}$-independent.
(i). Let $\delta=0$, so that $\Lambda=\Lambda_{0}$. Then the linear form is homogeneous, and we apply the method of Section 3.7.
(ii) Let $\delta \neq 0$. Then the linear form is inhomogeneous, and we apply the method of Section 3.8.
(iii). Suppose now that the $\mu_{i}$ 's are $\mathbb{Q}$-dependent. Let $\Gamma$ be the approximation lattice for the linear form $\Lambda$, as defined in Section 3.7. Then we expect the lower bound for $|\underline{x}|(\underline{x} \in \Gamma, \underline{x} \neq \underline{0})$ in general to be 'very small', since the vector having as coordinates the coefficients of the dependence relation will give rise to a very short vector in the lattice. So the reduction process, as applied in the two previous cases, will not work. In such a case we work as follows. Let $M$ be a maximal subset of $\left\{\mu_{1}, \ldots, \mu_{q}\right\}$ consisting of $\mathbb{Q}$-independent numbers. With an appropriate choice of subscripts we may assume that $M=\left\{\mu_{1}, \ldots, \mu_{p}\right\}, p<q$. Then we can find integers $d>0$ and $d_{i j}$ for $1 \leqslant i \leqslant p<j \leqslant q$ such that

$$
d \cdot \mu_{j}=\sum_{i=1}^{p} d_{i j} \cdot \mu_{i} \text { for } j=p+1, \ldots, q
$$

(These numbers $d, d_{i j}$ can be found as coordinates of extremely short vectors in reduced bases). On the other hand, (8.9) is equivalent to

$$
\begin{equation*}
\left|\Lambda^{\prime}\right|<K_{1}^{\prime} \cdot \exp \left(-K_{2} \cdot A\right), \quad A<K_{3}, \tag{8.10}
\end{equation*}
$$

where $\Lambda^{\prime}=\mathrm{d} \cdot \Lambda$ and $\mathrm{K}_{1}^{\prime}=\mathrm{d} \cdot \mathrm{K}_{1}$. Now, with $\delta^{\prime}=\mathrm{d} \cdot \delta$ and

$$
a_{i}^{\prime}=d \cdot a_{i}+\sum_{j=p+1}^{q} d_{i j} \cdot a_{j}
$$

we obtain

$$
\Lambda^{\prime}=\delta^{\prime}+\sum_{i=1}^{p} a_{i}^{\prime} \cdot \mu_{i}
$$

Put $D=\max \left(|d|,\left|d_{i j}\right|: 1 \leqslant i \leqslant p<j \leqslant q\right)$. Then

$$
\left|a_{i}^{\prime}\right| \leqslant(q-p+1) \cdot D \cdot A \text { for } i=1, \ldots, p
$$

Therefore, put $A^{\prime}=\max _{1 \leqslant i \leqslant p}\left|a_{i}^{\prime}\right|$, then $A^{\prime} \leqslant(q-p+1) \cdot D \cdot A$, and (8.10) implies

$$
\begin{equation*}
\left|\Lambda^{\prime}\right|<K_{1}^{\prime} \cdot \exp \left(-K_{2}^{\prime} \cdot A^{\prime}\right), \quad A^{\prime}<K_{3}^{\prime} \tag{8.11}
\end{equation*}
$$

where

$$
\begin{aligned}
& \Lambda^{\prime}=\delta^{\prime}+a_{1}^{\prime} \cdot \mu_{1}^{\prime}+\ldots+a_{p}^{\prime} \cdot \mu_{p}^{\prime}, \quad K_{1}^{\prime}=d \cdot K_{1} \\
& K_{2}^{\prime}=K_{2} /(q-1+p) \cdot D, \quad K_{3}^{\prime}=(q-p+1) \cdot K_{3}
\end{aligned}
$$

Now, to solve (8.11) we apply the reduction process described in (i) or (ii), depending on whether $\delta^{\prime}=0$ or $\delta^{\prime} \neq 0$, and maybe more than once, if necessary, until we find a very small upper bound for A' . After having found all solutions $\left(a_{1}^{\prime}, \ldots, a_{p}^{\prime}\right)$ of (8.11), we have a lower bound $L>0$ for $\left|\Lambda^{\prime}\right|$. It is reasonable to expect that $L$ is not 'extremely small', because the integers $a_{1}^{\prime}, \ldots, a_{p}^{\prime}$ being 'small' in absolute value cannot make $|\Lambda \prime|$ 'extremely small'. Now combine $|\Lambda \prime| \geqslant L$ with the first inequality of (8.10) to get

$$
A<\frac{1}{K_{2}} \cdot \log \left(\frac{K_{1}}{L}\right)
$$

Since L is not 'very small', as argued heuristically, the above upper bound for A is 'small'.

Returning now to the general case, we point out that if the reduced upper bound for $A$ (found after some reduction steps as described above) is not small enough to admit enumeration of the remaining possibilities in a
reasonable time, then it might be necessary, or at least advisable, to use the technique of Fincke and Pohst, cf. Section 3.6. However, when solving a Thue equation, and not only an inequality for a linear form in logarithms, it may be better to avoid this method, and to use continued fractions of the roots $\xi^{(i)}$. In practice we can search for the solutions (X,Y) of (8.1) satisfying $Y_{1}<|Y| \leqslant C$ as follows, referring to Lemma 8.1. Here e.g. $C=Y_{2}$, and we can imagine $C$ here as being a large' constant compared to $Y_{1}$, but not 'very large' (cf. the introduction of $Y_{1}, Y_{2}$ in Section 8.2).

Let $\tilde{\xi}$ be a rational approximation of $\xi^{\left(i_{0}\right)}$, such that

$$
\begin{equation*}
\left|\tilde{\xi}^{-} \xi^{\left(i_{0}\right)}\right|<\frac{1}{6 \cdot c^{2}} \tag{8.12}
\end{equation*}
$$

Since $|Y|>Y_{1}$, $X / Y$ must be a convergent, $p_{k} / q_{k}$ say, from the continued fraction expansion of $\xi^{\left(i_{0}\right.}$. Denote by $a_{0}, a_{1}, a_{2}, \ldots$ the partial quotients in this expansion. First we claim that $a_{k+1} \geqslant 3$. Indeed, by (3.5)

$$
\frac{1}{\left(a_{k+1}+2\right) \cdot|Y|^{2}} \leqslant \frac{1}{\left(a_{k+1}+2\right) \cdot q_{k}^{2}}<\left|\xi^{\left(i_{0}\right)}-\frac{p_{k}}{q_{k}}\right|=\left|\xi^{\left(i_{0}\right)}-\frac{X}{Y}\right| \leqslant \frac{C_{1}}{|Y|^{n}}
$$

If $a_{k+1}=1$ or 2 , then we would have $|Y|^{n-2}<4 \cdot C_{1}$, which is absurd, since $|Y|>Y_{1}>\left(4 \cdot C_{1}\right)^{1 /(n-2)}$. Thus, $a_{k+1} \geqslant 3$, and by (3.5) we have

$$
\left|\xi^{\left(i_{0}\right)}-\frac{p_{k}}{q_{k}}\right|<\frac{1}{a_{k+1} \cdot q_{k}^{2}} \leqslant \frac{1}{3 \cdot q_{k}^{2}} .
$$

Therefore,

$$
\left|\tilde{\xi}^{-p_{k}}\right| \leqslant\left|\tilde{\xi}^{q_{k}}\right| \xi^{\left(i_{0}\right)}\left|+\left|\xi^{\left(i_{0}\right)}-\frac{p_{k}}{q_{k}}\right|<\frac{1}{6 \cdot C^{2}}+\frac{1}{3 \cdot q_{k}^{2}} \leqslant \frac{1}{2 \cdot q_{k}^{2}}\right.
$$

and this means that $p_{k} / q_{k}$ is in fact a convergent from the continued fraction expansion of $\widetilde{\xi}$ too. Moreover, in view of the inequalities

$$
\frac{1}{\left(a_{k+1}+2\right) \cdot q_{k}^{2}}<\left|\xi^{\left(i_{0}\right)}-\frac{p_{k}}{q_{k}}\right| \leqslant \frac{C_{1}}{|Y|^{n}} \leqslant \frac{C_{1}}{\left|q_{k}\right|^{n}}
$$

$a_{k+1}$ must be sufficiently large compared to $q_{k}$, namely

$$
\begin{equation*}
a_{k+1}>\frac{\left|q_{k}\right|^{n-2}}{C_{1}}-2 \tag{8.13}
\end{equation*}
$$

This inequality can be checked easily for all $k$ such that $q_{k} \leqslant C$.

To sum up, we propose the following process for every real root $\xi^{\left(i_{0}\right)}$ for $i_{0}=1, \ldots, s$ (note that $i_{0}$ is a priori not known). (1) Compute a ( $i_{0}$ ) rational approximation $\tilde{\xi}$ of $\xi\left(\begin{array}{l}0 \\ \text { satisfying (8.12) (a truncation of its }\end{array}\right.$ decimal expansion will do). (2) Expand $\widetilde{\xi}$ into its continued fraction with partial quotients $a_{0}, a_{1}, a_{2}, \ldots, a_{k+1}$ and convergents $p_{i} / q_{i}$ for all $i=1, \ldots, k$ with $q_{k} \leqslant C<q_{k+1}$. (3) Test all these convergents for the conditions (8.13) and $F\left(p_{i}, q_{i}\right)=m$. Concerning this last test, note that if $X / Y=p_{i} / q_{i}$, then $X=Z \cdot p_{i}, Y=Z \cdot q_{i}$ for some $Z \in \mathbb{Z}$ with $Z^{n} \mid m$. This simple observation excludes in general most of the reducible quotients $X / Y$, and all of them if $m$ is an $n$-th-powerfree integer.

Having tested for all solutions in the range $|Y| \leqslant C$ we may suppose that $|Y|>C . F o r$ such solutions (X,Y) we can obtain a lower bound for the corresponding $A$ as follows (the idea is due to A. Pethö, cf. also Section 1
 $\{1, \ldots, n\}$ let $\varphi_{i j}$ be the number +1 or -1 for which $\left|\varepsilon_{i}^{(j)}\right|^{\nu j} \geqslant 1$, and put $E_{j}=\prod_{i=1}^{r}\left|\varepsilon_{i}^{(j)}\right|^{\varphi j}$. Then

$$
\left|\beta^{(j)}\right|=\left|\mu^{(j)}\right| \cdot \prod_{i=1}^{r}\left|\varepsilon_{i}^{(j)}\right|^{a} \leqslant \mu_{+} \cdot E_{j}^{A}
$$

and hence for any pair $j_{1}, j_{2}$ with $j_{1} \neq j_{2}$ we have

$$
|Y|=\frac{\left|\beta^{\left(j_{1}\right)}-\beta^{\left(j_{2}\right)}\right|}{\left|\xi^{\left(j_{1}\right)}-\xi^{\left(j_{2}\right)}\right|} \leqslant \mu_{+} \cdot \frac{E_{j_{1}}^{A}+E_{j_{2}}^{A}}{\left|\xi^{\left(j_{1}\right)}-\xi^{\left(j_{2}\right)}\right|}
$$

and from this we can find a lower bound for $A$, if we know that $|Y|>C$. Of course, for an other pair $j_{1}, j_{2}$ we may find a different lower bound, and therefore we can take the larger one.

### 8.5. An application: triangular numbers that are a product of three consecutive numbers.

In this section we prove, as an application of the general theory described in the previous sections, the following result. The problem was posed by S.P.

Mohanty (cf. Mohanty [1988]; the proof in this paper is incorrect). The n-th triangular number is for $n \in \mathbb{N}$ defined as $T_{n}=\frac{1}{2} \cdot n \cdot(n+1)$.

THEOREM 8.6. The only triangular numbers that are a product of three consecutive integers, are $\mathrm{T}_{3}=1 \cdot 2 \cdot 3, \mathrm{~T}_{15}=4 \cdot 5 \cdot 6, \quad \mathrm{~T}_{20}=5 \cdot 6 \cdot 7$, $T_{44}=9 \cdot 10 \cdot 11, \quad T_{608}=56 \cdot 57 \cdot 58, \quad T_{22736}=636 \cdot 637 \cdot 638$.

Proof. We have the diophantine equation $n \cdot(n+1)=2 \cdot m \cdot(m+1) \cdot(m+2)$ in $n, m \in \mathbb{N}$. Put $x=2 \cdot m+2, y=2 \cdot n+1$. Then we are lead to the equation $y^{2}=x^{3}-4 \cdot x+1$ in $x, y \in \mathbb{N}$, with $x \geqslant 4$ even and $y \geqslant 3$ odd. Theorem 8.7 below now completes the proof.

THEOREM 8.7. The elliptic curve

$$
\begin{equation*}
y^{2}=x^{3}-4 \cdot x+1 \tag{8.14}
\end{equation*}
$$

has only the following 22 integral points:

$$
\begin{aligned}
(x, \pm y)= & (-2,1),(-1,2),(0,1),(2,1),(3,4),(4,7),(10,31), \\
& (12,41),(20,89),(114,1217),(1274,45473) .
\end{aligned}
$$

We prove this theorem in two main steps. First, we reduce the problem to the solution of two quartic Thue equations. Then we solve these equations using the general theory developed in the previous sections.

Let $L$ be the totally real field $\mathbb{Q}(\psi)$, where

$$
\psi^{3}-4 \cdot \psi+1=0
$$

Let the conjugates of $\psi$ be $\psi^{(1)}=0.254 \ldots, \psi^{(2)}=-2.114 \ldots$, $\psi^{(3)}=1.860 \ldots$. From a table of Delone and Faddeev ([1964], p. 141) we see that the class number of $L$ is 1 , its ring of integers is $\mathbb{Z}[\psi]$, its discriminant is 229 , and a pair of independent units is $\psi, 2-\psi$. From Table I of Buchmann [1986] we see that $-7+2 \cdot \psi^{2}, 2 \cdot \psi+\psi^{2}$ is a pair of fundamental units in $\mathbb{Z}[\psi]$. By $-7+2 \cdot \psi^{2}=-\psi^{-1} \cdot(2-\psi), 2 \cdot \psi+\psi^{2}=(2-\psi)^{-1}$ we see that $\psi, 2-\psi$ is also a pair of fundamental units in $\mathbb{Z}[\psi]$.

The equation (8.14) of the elliptic curve can be written as

$$
\begin{equation*}
y^{2}=(x-\psi) \cdot\left(x^{2}+x \cdot \psi+\left(\psi^{2}-4\right)\right) \tag{8.15}
\end{equation*}
$$

and the factors on the right hand side are relatively prime. Indeed, if $\pi$ were a common prime divisor of them, then $\pi$ would divide

$$
\left(x^{2}+x \cdot \psi+\left(\psi^{2}-4\right)\right)-(x+2 \cdot \psi) \cdot(x-\psi)=3 \cdot \psi^{2}-4
$$

which is prime, since its norm is -229 . Therefore we would have that $\pi$ is a unit times this prime, and then by (8.15), $x-\psi=$ unit $\times\left(3 \cdot \psi^{2}-4\right) \times$ square . Take norms, then we get $y^{2}= \pm 229 \times$ square , which is clearly impossible.

Now (8.15) implies

$$
\begin{equation*}
x-\psi= \pm \psi^{i} \cdot(2-\psi)^{j} \cdot \alpha^{2}, \quad \alpha \in \mathbb{Z}[\psi], \quad i, j \in\{0,1\} \tag{8.16}
\end{equation*}
$$

Since (8.14) is trivial to solve for $x \leqslant 0$ (the only solutions with $x \leqslant 0$ are the first three pairs stated in the theorem), we may assume that $x \geqslant 1$. Since $\psi^{(1)}=0.254 \ldots$, we see that the minus sign in (8.16) is impossible. Then, by $\psi^{(2)}=-2.114 \ldots, \quad i \neq 1$. We conclude therefore that

$$
\begin{equation*}
x-\psi=(2-\psi)^{j} \cdot\left(u+v \cdot \psi+w \cdot \psi^{2}\right)^{2}, \quad u, v, w \in \mathbb{Z}, j \in\{0,1\} \tag{8.17}
\end{equation*}
$$

First case: $j=0 . \quad$ Then (8.17) implies, on equating corresponding coefficients in both sides,

$$
\begin{equation*}
x=u^{2}-2 \cdot v \cdot w, \quad w^{2}-2 \cdot u \cdot v-8 \cdot v \cdot w=1, v^{2}+4 \cdot w^{2}+2 \cdot u \cdot w=0 \tag{8.18}
\end{equation*}
$$

Note that $w$ is odd and $v$ is even, hence $4 \mid 2 \cdot u \cdot w$, so $u$ is even. Put $u=2 \cdot u_{1}, \quad v=2 \cdot v_{1}$. The last equation of (8.18) now reads

$$
w^{2}+u_{1} \cdot w+v_{1}^{2}=0
$$

Consider this as a quadratic equation in $w$. Its discriminant must be a square, $z^{2}$ say. Then

$$
u_{1}^{2}-4 \cdot v_{1}^{2}=z^{2}, \quad w=\frac{1}{2}\left(-u_{1} \pm z\right)
$$

Note that $u_{1}$ and $z$ have the same parity. We may assume $u \geqslant 0$.

First suppose that $u_{1}$ and $z$ are even. Since $w^{2}+u_{1} \cdot w+v_{1}^{2}=0$ and $w$ is odd, we find $u_{1} \equiv 2(\bmod 4)$, and $v_{1}$ is odd. Put $u_{1}=2 \cdot u_{2}$, $z=2 \cdot z_{1}$. Then $u_{2}^{2}-v_{1}^{2}=z_{1}^{2}$, where $u_{2}$ and $v_{1}$ are odd. By $u_{2} \geqslant 0$ there exist $m, n \in \mathbb{Z}$ such that

$$
u_{2}=m^{2}+n^{2}, \quad v_{1}=m^{2}-n^{2}, \quad z_{1}=2 \cdot m \cdot n
$$

It follows that

$$
\mathrm{u}=4 \cdot\left(\mathrm{~m}^{2}+\mathrm{n}^{2}\right), \quad \mathrm{v}=2 \cdot\left(\mathrm{~m}^{2}-\mathrm{n}^{2}\right), \quad \mathrm{w}=-(m \pm \mathrm{n})^{2}
$$

Since the sign of $z$, and thus that of $n$, is of no importance, we may assume $w=-(m+n)^{2}$. After substitution in the second equation of (8.18) we obtain the Thue equation

$$
m^{4}+36 \cdot m^{3} \cdot n+6 \cdot m^{2} \cdot n^{2}-28 \cdot m \cdot n^{3}+n^{4}=1
$$

The left hand side can be factored as

$$
(m+n) \cdot\left(m^{3}+35 \cdot m^{2} \cdot n-29 \cdot m \cdot n^{2}+n^{3}\right)
$$

and therefore it can be solved very easily. Its only solutions are $\pm(m, n)=(1,0),(0,1)$. They lead to $\pm(u, v, w)=(4,2,-1),(4,-2,-1)$, and then by (8.18) we find $x=20,12$ respectively, which furnish the solutions $(x, \pm y)=(20,89),(12,41)$ for (8.14).

Secondly, we suppose that $u_{1}$ and $z$ are odd. Then $v_{1}$ is even, so by $u_{1} \geqslant 0$ there exist $m, n \in \mathbb{Z}$ with

$$
u_{1}=m^{2}+n^{2}, \quad 2 \cdot v_{1}=2 \cdot m \cdot n, \quad z=m^{2}-n^{2}
$$

It follows that

$$
\mathrm{u}=2 \cdot\left(\mathrm{~m}^{2}+\mathrm{n}^{2}\right), \quad \mathrm{v}=2 \cdot \mathrm{~m} \cdot \mathrm{n}, \quad \mathrm{w}=-\mathrm{m}^{2} \quad \text { or } \quad \mathrm{w}=-\mathrm{n}^{2}
$$

We may assume that $w=-m^{2}$. Substituting this in the second equation of (8.18) we find the Thue equation

$$
m^{4}+8 \cdot m^{3} \cdot n-8 \cdot m \cdot n^{3}=1
$$

The left hand side is again reducible. The only solutions, as is easily seen, are $\pm(m, n)=(1,0),(1,1),(1,-1)$. Since $m$ and $n$ cannot have the same parity, only the first pair is accepted. It leads to (u,v,w) = (2,0,-1) , and hence to $(x, \pm y)=(4,7)$ for (8.14).

Second case: $j=1$. Then, equating the coefficients in (8.17) we get

$$
\begin{equation*}
x=2 \cdot u^{2}+v^{2}+4 \cdot w^{2}+2 \cdot u \cdot w-4 \cdot v \cdot w, \tag{8.19}
\end{equation*}
$$

$$
\left\{\begin{array}{r}
u^{2}+4 \cdot v^{2}+18 \cdot w^{2}-4 \cdot u \cdot v+8 \cdot u \cdot w-18 \cdot v \cdot w=1  \tag{8.20}\\
2 \cdot v^{2}+9 \cdot w^{2}-2 \cdot u \cdot v+4 \cdot u \cdot w-8 \cdot v \cdot w=0
\end{array}\right.
$$

The first relation of (8.20) can be replaced by

$$
\begin{equation*}
u^{2}-2 \cdot v \cdot w=1 \tag{8.21}
\end{equation*}
$$

Note that $u$ is odd. Put $z=v-2 \cdot w$. Then the second equation of (8.20) yields

$$
w^{2}=2 \cdot z \cdot(u-z)
$$

First we suppose that $z$ is odd. Then there exist $m, n \in \mathbb{Z}$ such that

$$
z=m^{2}, \quad u-z=2 \cdot n^{2}
$$

where we use that $u \geqslant 0$ and $(u, w)=1$. Thus, choosing signs properly,

$$
u=m^{2}+2 \cdot n^{2}, \quad v=m^{2}+4 \cdot m \cdot n, \quad w=2 \cdot m \cdot n
$$

Substituting this in (8.21) we obtain the Thue equation

$$
\begin{equation*}
m^{4}-4 \cdot m^{3} \cdot n-12 \cdot m^{2} \cdot n^{2}+4 \cdot n^{4}=1 . \tag{8.22}
\end{equation*}
$$

In Theorem 8.8(i) below we prove that this equation has only the solutions $\pm(m, n)=(1,0)$, leading to $(u, v, w)=(1,1,0)$, and finally for (8.14) to $(x, \pm y)=(3,4)$.

Secondly we suppose that $z$ is even. Then there exist $m, n \in \mathbb{Z}$ with

$$
z=2 \cdot m^{2}, \quad u-z=n^{2}
$$

Thus, choosing signs properly, we find

$$
u=2 \cdot m^{2}+n^{2}, \quad v=2 \cdot m^{2}+4 \cdot m \cdot n, \quad w=2 \cdot m \cdot n
$$

Now, substituting into (8.21), we obtain the Thue equation

$$
\begin{equation*}
n^{4}-12 \cdot n^{2} \cdot m^{2}-8 \cdot n \cdot m^{3}+4 \cdot m^{4}=1 \tag{8.23}
\end{equation*}
$$

In Theorem 8.8(ii) below we prove that this equation has only the solutions $\pm(m, n)=(0,1),(1,-1),(3,1),(-1,3)$. They lead respectively to $(u, v, w)=(1,0,0),(3,-2,-2),(19,30,6),(11,-10,-6)$, which lead for (8.14)
to the solutions $(x, \pm y)=(2,1),(10,31),(1274,45473),(114,1217)$. Thus this result completes the proof of Theorem 8.7, provided the Thue equations (8.22), (8.23) have as their only solutions the pairs (m,n) mentioned above. We now proceed to prove this.

THEOREM 8.8. (i). The Thue equation

$$
\begin{equation*}
X^{4}-4 \cdot X^{3} \cdot Y-12 \cdot X^{2} \cdot Y^{2}+4 \cdot Y^{4}=1 \tag{8.24}
\end{equation*}
$$

has only the solutions $\pm(\mathrm{X}, \mathrm{Y})=(1,0)$.
(ii). The Thue equation

$$
\begin{equation*}
X^{4}-12 \cdot X^{2} \cdot Y^{2}-8 \cdot X \cdot Y^{3}+4 \cdot Y^{4}=1 \tag{8.25}
\end{equation*}
$$

has only the solutions $\pm(\mathrm{X}, \mathrm{Y})=(1,0),(1,-1),(1,3),(3,-1)$.

Proof. We use the notation and results of Sections 8.2 and 8.3. Let the algebraic numbers $\vartheta$ and $\varphi$ be defined by

$$
\vartheta^{4}-12 \cdot \vartheta^{2}-8 \cdot \vartheta+4=0, \quad \varphi^{4}-4 \cdot \varphi^{3}-12 \cdot \varphi^{2}+4=0 .
$$

Since $\varphi=2 / \vartheta$, it follows that $\vartheta$ and $\varphi$ generate the same field $K$ over Q. In the notation of Section 8.2 we have $n=4, s=4, t=0$, and $\xi=\vartheta$ or $\xi=\varphi$. Simple computations show that for $\xi=\vartheta, \varphi$ we can take

$$
\begin{gathered}
Y_{0}=1, \quad C_{1}=0.843, \quad C_{2}=0.589, \quad Y_{1}=2, \quad C_{3}=6.645, \\
Y_{2}^{*}=3, \quad \mu_{-}=\mu_{+}=1, \quad C_{4}=8.3374 .
\end{gathered}
$$

In these computations we estimated $C_{1}, C_{3}, C_{4}$ from above and $C_{2}$ from below, using the following approximations for the conjugates of $\vartheta$ and $\varphi$ :

$$
\begin{aligned}
& \vartheta^{(1)} \cong-1.080286352, \varphi^{(1)} \cong-1.851360980, \\
& \vartheta^{(2)} \cong 3.722935260, \varphi^{(2)} \cong 0.537210524, \\
& \vartheta^{(3)} \cong 0.334111716, \varphi^{(3)} \cong 5.986021747, \\
& \vartheta^{(4)} \cong-2.976760624, \varphi^{(4)} \cong-0.671871290 .
\end{aligned}
$$

Now we work in the order R of K with $\mathbb{Z}$-basis $\left\{1, \vartheta, \frac{1}{2} \cdot \vartheta^{2}, \frac{1}{2} \cdot \vartheta^{3}\right\}$ (note that $\frac{1}{2} \cdot \vartheta^{2}$ is an algebraic integer). Note that

$$
\varphi=\frac{2}{\vartheta}=4+6 \cdot \vartheta-\frac{1}{2} \cdot \vartheta^{3} \in \mathrm{R} .
$$

On the other hand, (8.24) and (8.25) are respectively equivalent to
$\operatorname{Norm}_{K / \mathbb{Q}}(X-Y \cdot \vartheta)=1$ and $\operatorname{Norm}_{K / \mathbb{Q}}(X-Y \cdot \varphi)=1$, which means that if $(X, Y)$ is a solution of (8.24) or (8.25), then $X-Y \cdot \vartheta$ or $X-Y \cdot \varphi$, respectively, is a unit of the order $R$. A system of fundamental units of $R$ is given by

$$
\varepsilon_{1}=1+\vartheta, \quad \varepsilon_{2}=3+\vartheta, \quad \varepsilon_{3}=\frac{1}{2} \cdot \vartheta^{2}
$$

We do not prove this fact here. For a proof, see Tzanakis and de Weger [1989 ${ }^{\text {a }], ~ S e c t i o n ~ I I I . ~} 2$ and Appendix I.

Thus the solution of (8.24) and (8.25) is reduced to finding all $\left(a_{1}, a_{2}, a_{3}\right) \in \mathbb{Z}^{3}$ such that the unit $\pm \varepsilon_{1}{ }^{a_{1}} \cdot \varepsilon_{2}{ }^{a_{2}} \cdot \varepsilon_{3}{ }^{a_{3}}$ has the special shape $X-Y \cdot \vartheta$ or $X-Y \cdot \varphi$, respectively. In the notation of Lemma 8.3 we have, after some numerical computations, that we leave to the reader to check, that

$$
\min _{\mathrm{I}} \mathrm{~N}\left[\mathrm{U}_{\mathrm{I}}^{-1}\right]=0.634950 \ldots, \quad \max _{\mathrm{I}} \mathrm{~N}\left[\mathrm{U}_{\mathrm{I}}^{-1}\right]=1.210070 \ldots
$$

(here, of course, $I=\{1,2,3,4\}$ ). Therefore we can take in Lemma 8.4

$$
C_{5}=1.211
$$

Also,

$$
C_{6}=6.38771 \times 10^{4}, \quad Y_{2}^{\prime}=3
$$

(The values of $C_{5}$ and $C_{6}$ are estimated from above.)

Now, relation (8.3) becomes in our case

$$
\begin{equation*}
\Lambda=\log \left|\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}-\xi^{(k)}}\right|+\sum_{i=1}^{3} a_{i} \cdot \log \left|\frac{\varepsilon_{i}^{(k)}}{\varepsilon_{i}^{(j)}}\right| \tag{8.26}
\end{equation*}
$$

where $\xi=\vartheta$ or $\varphi$. As mentioned in Section 8.2, once $i_{0}$ is fixed, we can choose $j, k$ arbitrarily. Thus we can choose

$$
\left\{\begin{array}{l}
j=3, k=4 \quad \text { if } i_{0}=1 \text { or } 2  \tag{8.27}\\
j=1, k=2 \quad \text { if } i_{0}=3 \text { or } 4
\end{array}\right.
$$

Therefore, for each $\xi \in\{\vartheta, \varphi\}$ we have four possibilities for $\Lambda$. For each of these eight cases we have, as will be shown below,

$$
C_{7}=5.71 \times 10^{38}, \quad C_{8}=6.17
$$

and therefore, by Lemma 8.5, if $|Y|>3$, then for $A=\max _{1 \leqslant i \leqslant 3}\left|a_{i}\right|$ we have
the upper bound $C_{9}=3.26 \times 10^{40}$. As is easily checked, the only solutions of either (8.24) or (8.25) with $|Y| \leqslant 3$ are those listed in the statement of the theorem. Therefore we may assume that $|Y|>3$, so that

$$
A<3.26 \times 10^{40}
$$

Before we apply the reduction method of Section 3.8 we show that the application of Lemma 2.4 yields the above constants $C_{7}, C_{8}$. We apply this result in the case of $\Lambda$ given by (8.26). In this case, we compute the $V_{i}$ 's for the various $\alpha_{i}$ 's appearing in $\Lambda$, as follows. If $\alpha_{i}=\left|\varepsilon_{i}^{(k)} / \varepsilon_{i}^{(j)}\right|$ for $i=1,2,3$, then $\alpha_{i}$ is a unit and hence $a_{0}$ (appearing in the computation of $h\left(\alpha_{i}\right)$ ) is equal to 1 . Clearly, every conjugate of $\alpha_{i}$ is in absolute value less than

$$
H_{i}=\frac{\max _{1 \leqslant h \leqslant 4}\left|\varepsilon_{i}^{(h)}\right|}{\min _{1 \leqslant h \leqslant 4}\left|\varepsilon_{i}^{(h)}\right|}
$$

and $H_{i} \geqslant 1$. Therefore, $h\left(\alpha_{i}\right) \leqslant H_{i}$, and we can take

$$
V_{i}=\max \left(\log H_{i},|\log | \varepsilon_{i}^{(k)} / \varepsilon_{i}^{(j)}| |\right)
$$

Since the latter term equals the logarithm of either $\left|\varepsilon_{i}^{(k)} / \varepsilon_{i}^{(j)}\right|$ or its inverse, it follows that

$$
\mathrm{V}_{\mathrm{i}}=\log \mathrm{H}_{\mathrm{i}}
$$

If $\alpha_{i}=\left|\xi^{\left(i_{0}\right)}-\xi^{(j)}\right| /\left|\xi^{\left(i_{0}\right)}-\xi^{(k)}\right|$, then all conjugates of $\alpha_{i}$ are in absolute value less than $C_{3}$. Therefore, $h\left(\alpha_{i}\right) \leqslant\left(\log a_{0}\right) / d+\log C_{3}$, where $a_{0}$ and $d$ are as in the definition of $h(\alpha)$ for $\alpha=\alpha_{i}$. An upper bound for $a_{\rho}$ can be computed as follows. Consider the algebraic numbers $\chi_{i h}=\frac{1}{2} \cdot\left(\xi^{(i)}-\xi^{(h)}\right)$ for $i, h \in\{1, \ldots, 4\}$ with $i \neq h$. It can be checked that the numbers $\chi_{i h}$ are algebraic integers for $\xi=\vartheta$ or $\varphi$. Now, for each permutation $\sigma=\left(\sigma_{1} \sigma_{2} \sigma_{3} \sigma_{4}\right) \in S_{4}$ we consider the number $\chi(\sigma)=\chi_{\sigma_{1} \sigma_{2}}{ }^{\prime} \chi_{\sigma_{1}} \sigma_{3}$ (independent of $\sigma_{4}$ ), and the polynomial

$$
P(X)=\prod_{\sigma \in S_{4}}(X-\chi(\sigma))
$$

Consider also the number

$$
\Delta=\prod_{1 \leqslant i<h \leqslant 4} \chi_{i h} .
$$

Note that

$$
\Delta^{2}=\frac{1}{2^{12}} \cdot \prod_{1 \leqslant i<h \leqslant 4}\left(\xi_{i}-\xi_{h}\right)^{2}=\frac{1}{2^{12}} \cdot D
$$

where $D$ is the discriminant of the defining polynomial of $\xi$, and therefore $\Delta^{2}=229$. On the other hand, the coefficients of $P(X)$ are up to the sign the elementary symmetric functions of $\chi(\sigma)$ for $\sigma \in S_{4}$, and so they are symmetrical expressions of the $\xi^{(i)}$,s with rational coefficients. This means that $P(X) \in \mathbb{Q}[X]$. On the other hand, by the definition of $\Delta$, any coefficient of $P(X)$ multiplied by $\Delta^{4}$ is a polynomial of the $\chi_{i h}$ 's with coefficients in $\mathbb{Z}$ and therefore it is an algebraic integer. Combine this with the fact that $P(X) \in \mathbb{Q}[X]$ to see that $229^{2} \cdot P(X) \in \mathbb{Z}[X]$. Hence, since $\alpha_{i}$ is a root of $P(X)$, its leading coefficient $a_{0}$ is at most $229^{2}$. To conclude, we have $h\left(\alpha_{i}\right) \leqslant 2 \cdot(\log 229) / d+\log C_{3}$ and it is clear that $\left|\log \alpha_{i}\right| / d \leqslant \log C_{3}$. Since $\alpha_{i} \notin \mathbb{Q}$ we have $d \geqslant 2$, so we can take

$$
\mathrm{V}_{\mathrm{i}}=\log 229+\log \mathrm{C}_{3}
$$

Simple computations now show that

$$
\begin{aligned}
& \log H_{1}=4.074586 \ldots, \quad \log H_{2}=5.667432 \ldots, \\
& \log H_{3}=4.821584 \ldots, \\
& \log C_{3}=1.262065 \ldots \text { if } \xi=\vartheta, \\
& \log C_{3}=1.893823 \ldots \text { if } \xi=\varphi, \\
& \log 229+\log C_{3} \leqslant 7.327545 \ldots .
\end{aligned}
$$

Therefore we apply Lemma 2.4 (Waldschmidt) with $n=4, D \leqslant 24, e(n)=73$,

$$
\alpha_{1}=\left|\frac{\varepsilon_{1}^{(k)}}{\varepsilon_{1}^{(j)}}\right|, \quad \alpha_{2}=\left|\frac{\varepsilon_{3}^{(k)}}{\varepsilon_{3}^{(j)}}\right|, \quad \alpha_{3}=\left|\frac{\varepsilon_{2}^{(k)}}{\varepsilon_{2}^{(j)}}\right|, \quad \alpha_{4}=\left|\frac{\xi^{\left(i_{0}\right)}-\xi^{(j)}}{\xi^{\left(i_{0}\right)}{ }_{-\xi^{(k)}}^{(k)}}\right|
$$

for $\xi=\vartheta$ or $\varphi$, and $b_{1}=a_{1}, b_{2}=a_{3}, b_{3}=a_{2}, b_{4}=1, B=A$, $\mathrm{V}_{1}=\log \mathrm{H}_{1}, \quad \mathrm{~V}_{2}=\log \mathrm{H}_{3}, \quad \mathrm{~V}_{3}=\mathrm{V}_{3}^{+}=\log \mathrm{H}_{2}, \mathrm{~V}_{4}=\mathrm{V}_{4}^{+}=\log 229+\log \mathrm{C}_{3}$. Thus we find that

$$
|\Lambda|>\exp \left(-C_{7} \cdot\left(\log A+C_{8}\right)\right)
$$

with $C_{7}=5.71 \times 10^{38}$ and $C_{8}=6.17$.

We have now to apply the reduction process described in Section 3.7. In our situation we have to solve (8.9) with

$$
K_{1}=C_{6}=6.38771 \times 10^{4}, \quad K_{2}=\frac{n}{C_{5}}=\frac{4}{1.211}>3.303, \quad K_{3}=3.26 \times 10^{40}
$$

( $\mathrm{K}_{2}$ is estimated from below), and

$$
\Lambda=\delta+a_{1} \cdot \mu_{1}+a_{2} \cdot \mu_{2}+a_{3} \cdot \mu_{3}
$$

where for $\delta$ and the $\mu_{i}$ 's we have, in view of (8.26) and (8.27):
or

$$
\left\{\begin{array}{l}
\delta=\delta_{3}:=\log \left|\frac{\xi^{(3)}-\xi^{(1)}}{\xi^{(3)}-\xi^{(2)}}\right| \text { or } \delta=\delta_{4}:=\log \left|\frac{\xi^{(4)}-\xi^{(1)}}{\xi^{(4)}-\xi^{(2)}}\right|,  \tag{8.29}\\
\text { where } \xi=\vartheta \text { or } \varphi,
\end{array}, \left.\frac{\varepsilon_{i}^{(2)}}{\mu_{i}^{(1)}} \right\rvert\,, \text { for } i=1,2,3 . .\right.
$$

Numerical details are given in the preprint version of Tzanakis and de Weger [1989 ${ }^{\mathrm{a}}$ ] (to be obtained from the author). We take $c_{0}=10^{140}$, and we work with the lattice with associated matrix

$$
\mathscr{A}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
{\left[c_{0} \cdot \mu_{1}\right]} & {\left[c_{0} \cdot \mu_{2}\right]} & {\left[c_{0} \cdot \mu_{3}\right]}
\end{array}\right)
$$

Note that in each of the four cases of (8.28) (resp. (8.29)) we have the same lattice, $\Gamma_{1}$ (resp. $\Gamma_{2}$ ), say. In each case $\delta \neq 0$, and we had no numerical evidence that the $\mu_{i}$ 's are $\mathbb{Q}$-dependent. Therefore we worked as in case (ii) of Section 8.4.

For each $\Gamma_{i}$ we have applied the integral version of the $L^{3}$-algorithm, and each time we have computed the integral $3 \times 3$-matrices $\mathcal{B}, \mathcal{U}, U^{-1}$, as defined in Section 3.7. In our cases, the coordinates of the vectors of the reduced bases (i.e. the elements of $\mathcal{B}$ ) turned out to have 46 to 48 digits, i.e. the lengths of the reduced basis vectors are of the size of $c_{0}^{1 / 3}$, as expected.

In each of the eight cases we computed the coordinates $s_{1}, s_{2}, s_{3}$ of

$$
\underline{x}=\left(\begin{array}{c}
0 \\
0 \\
-\left[c_{0} \cdot \delta\right]
\end{array}\right)
$$

with respect to the reduced basis $\underline{b}_{1}, \underline{b}_{2}, \underline{b}_{3}$ of the lattice. From our computations we found

$$
\begin{aligned}
& \left|\underline{\mathrm{b}}_{1}\right|>3.247 \times 10^{46} \text { in the case of lattice } \Gamma_{1}, \\
& \left|\underline{\mathrm{~b}}_{1}\right|>4.846 \times 10^{46} \text { in the case of lattice } \Gamma_{2}, \\
& I I s_{3} I>0.029 \text { in all } 8 \text { cases. }
\end{aligned}
$$

This means that in view of Lemma 3.5 , in all cases $i_{0}=3$, and

$$
\ell\left(\Gamma_{i}, \underline{x}\right)>0.029 \cdot \frac{1}{2} \cdot 3.247 \times 10^{46}>4.708 \times 10^{44}
$$

Then the assumptions of Lemma 3.10 are fulfilled with $n=3, \gamma=1, C=c_{0}$, $c=K_{1}, \delta=K_{2}, X_{0}=X_{1}=K_{3}$, since $\sqrt{ } 27 \cdot K_{3}<1.112 \times 10^{40}$, which implies

$$
A<\frac{1}{3.303} \cdot \log \left(10^{140} \cdot 6.38771 \times 10^{4} / 3.26 \times 10^{40}\right)<72.8
$$

It follows that $A \leqslant 72$. We repeat the procedure with $K_{3}=72$ and $c_{0}=10^{12}$. We found from our computations

$$
\begin{aligned}
& \left|\underline{b}_{1}\right|>1.293 \times 10^{4} \text { in the case of lattice } \Gamma_{1}, \\
& \left|\underline{b}_{1}\right|>1.092 \times 10^{4} \text { in the case of lattice } \Gamma_{2}, \\
& I s_{3} \|>0.143 \text { in all } 8 \text { cases. }
\end{aligned}
$$

This means that in view of Lemma 3.5 , in all cases $i_{0}=3$, and

$$
\ell\left(\Gamma_{i}, \underline{x}\right)>0.143 \cdot \frac{1}{2} \cdot 1.092 \times 10^{4}>7.807 \times 10^{2}
$$

Then the assumptions of Lemma 3.10 are fulfilled, since $\sqrt{ } 27 \cdot \mathrm{~K}_{3}<3.742 \times 10^{2}$, which implies

$$
A<\frac{1}{3.303} \cdot \log \left(10^{12} \cdot 6.38771 \times 10^{4} / 72\right)<10.5
$$

It follows that $A \leqslant 10$. We enumerated all remaining possibilities, and found no other solutions of (8.24) and (8.25) than those mentioned.

The computations for the proof of Theorem 8.8 took 35 sec .

### 8.6. The Thue-Mahler equation, an outline.

Let $F(X, Y)$ be as in Section 8.1. Let $p_{1}, \ldots, p_{S}$ be fixed distinct prime numbers. The diophantine equation

$$
F(X, Y)= \pm \prod_{i=1}^{s} p_{i}^{n_{i}}
$$

in the variables $X, Y \in \mathbb{Z}, n_{1}, \ldots, n_{S} \in \mathbb{N}_{0}$, with $(X, Y)=1$, is known as a Thue-Mahler equation. It was proved by Mahler [1933] that this equation has only finitely many solutions, and by Coates [1970] that they can, at least in principle, be determined effectively, since an effectively computable upper bound for the variables can be derived from the p-adic theory of linear forms in logarithms. For the history of this equation we refer to Shorey and Tijdeman [1986], Chapter 7.

We believe that it is possible to solve Thue-Mahler equations, not only in principle, but in practice. This can be done by reducing the above mentioned upper bounds, using a combination of real and p-adic computational diophantine approximation techniques, based on the $L^{3}$-algorithm for reducing bases of lattices (cf. Sections 3.7 and 3.8 for the real case, 3.11 and 3.12 for the p-adic case, Section 1.5 for a short outline of how to combine the real and p-adic techniques, and Sections 4.8 and 6.4 for some explicit examples of such combined techniques). The method can be considered as a p-adic analogue of the method for solving Thue equations, on which we reported in the preceding sections.

Such an idea (but without using the $L^{3}$-algorithm) was used by Agrawal, Coates, Hunt and van der Poorten [1980], who solved the equation

$$
X^{3}-X^{2} \cdot Y+X \cdot Y^{2}+Y^{3}= \pm 11^{n}
$$

This is to the author's knowledge the only example in the literature where a Thue-Mahler equation has been solved by the Gelfond-Baker method. Other methods may apply as well for solving Thue-Mahler equations. For example,

$$
X^{3}+3 \cdot Y^{3}=2^{n}
$$

has been solved by Tzanakis [1984] by a different method. The advantage of the Gelfond-Baker method above many other ideas is that it works in principle for any Thue-Mahler equation, because it is not very much dependent on the parameters of the particular equation that one wants to solve.

Both examples of Thue-Mahler equations mentioned above are of the simplest kind, in view of the fact that the cubic field $\mathbb{Q}(\vartheta)$, where $\vartheta$ is a root of $F(x, 1)=0$, has only one fundamental unit, and there occurs only one prime. Therefore it is sufficient to use two-dimensional real continued fractions and one-dimensional p-adic continued fractions, instead of the more complicated $L^{3}$-algorithm (which anyway was not yet available in 1980, when Agrawal, Coates, Hunt and van der Poorten did their work). With the use of the $L^{3}$-algorithm the method can in principle be extended to the general situation, where there are more than one fundamental units, and more than one primes. In a forthcoming publication, Tzanakis and the present author plan to give details and worked-out examples (Tzanakis and de Weger [1989 ${ }^{\text {b }}$ ]).

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