## NON-STANDARD ANALYSIS

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These notes are edited from those used for a course given August 10-16, 2009, at the Nesin Mathematics Village (Nesin Matematik Köyü), Kayser Dağı mevkii, Şirince, Selçuk, İzmir, Turkey. The major changes have been to $\$ \S 3$ and 4. I have made corrections throughout the notes, but this is not a thorough revision. The source files (with any further corrections and changes) will available at

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http://metu.edu.tr/~dpierce/Courses/Sirince/2009-Analysis/
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(This is a change from the address given in the original notes, for reasons beyond my control.)

I had about two dozen students in the first lecture of the course. About one dozen students came to each of the later lectures; but they were not always the same dozen. I managed to take the names of 16 students who attended at least one lecture.

In six lectures I could not expect to discuss all details in these notes. I left many theorems unproved. Another time, I might skip Archimedes's quadrature of the parabola by balancing, discussing only his quadrature by inscribed triangles. (I want to do at least this much, though, in order to connect the 'Archimedean axiom' with the person it is named for.) As it was, I did do non-standard analysis proper in the last two days. I noted and resolved the paradoxes that ${ }^{*} \mathbb{N}$ and ${ }^{*} \mathbb{R}$ should have the 'same' properties as $\mathbb{N}$ and $\mathbb{R}$, and yet cannot.

[^0]
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## Introduction

Mathematical analysis is the theoretical side of calculus. Calculus consists of methods of solving certain sorts of problems; analysis studies those methods. The standard way of doing this is founded on the 'epsilon-delta' definition of limit. The non-standard approach uses infinitesimals, with rigorous logical justification. Abraham Robinson first gave this justification; it can be found in his book, Non-standard Analysis [13].

An infinitesimal is a number whose absolute value is less than than every positive rational number. If two numbers $a$ and $b$ differ by an infinitesimal, we write

$$
a \simeq b
$$

Zero is an infinitesimal, but there are no other infinitesimals among the so-called real numbers.

In standard analysis, a function $f$ is said to be continuous at an element $a$ of its domain if

$$
\lim _{x \rightarrow a} f(x)=f(a)
$$

this means that, for all positive numbers $\varepsilon$, there is a positive number $\delta$ such that, for all $x$ in the domain of $f$, if $|x-a|<\delta$, then $|f(x)-f(a)|<\varepsilon$.

In non-standard analysis, there is an alternative formulation of continuity: $f$ is continuous at $a$ just in case, for all $x$ in the domain of $f$, if $x \simeq a$, then $f(x) \simeq f(a)$.

The alternative formulation of continuity and many other things will be worked out in the last section, $\S 6$, of these notes. The other sections are meant to provide logical justication and motivation for this work. Section 1 looks at Archimedes's solution of a calculus problem, and also mentions the Archimedean axiom, which will come up later in various contexts. Today we think of calculus as involving the complete ordered field $\mathbb{R}$ of real numbers; this field is constructed in $\S \S 2$ and 3 . Non-standard analysis requires a certain larger ordered field, ${ }^{*} \mathbb{R}$, which is an example of a non-Archimedean ordered field. Non-Archimedean ordered fields in general, and simple examples of these and related fields, are discussed in $\S 4$. The field ${ }^{*} \mathbb{R}$ can be obtained as an ultrapower of $\mathbb{R}$; this construction is treated in $\S 5$. One can jump ahead to $\S 6$ at any time, provided one understands the meaning of Theorem 45 in $\S 5.2$.

## 1. Archimedes's quadrature of the parabola

In the last chapter of Non-standard Analysis [13], Robinson treats the history of calculus in the light of non-standard analysis. Robinson begins with Leibniz; but I think it worthwhile to go back much further - about two thousand years further. In the work of Archimedes, both standard and non-standard approaches to calculus (in our terms) can be discerned. For example, Archimedes takes up the following

Problem 1. Find a square equal to a given segment of a parabola.
Parabolas will be defined below; meanwhile, a segment of a parabola can be seen in Figure 1, with an inscribed triangle. A solution to Problem 1 is called a quadrature of


Figure 1
the parabola. Archimedes's solution is given by the following
Theorem 2 (Archimedes). A parabolic segment is a third again as large ${ }^{1}$ as the inscribed triangle with the same altitude.

In Figure 1, triangle $A B C$ has the same altitude as the parabolic segment, because the tangent to the parabola at $A$ is parallel to the chord $B C$. Archimedes proves Theorem 2 in two ways in his Quadrature of the parabola: in Proposition 17 (and the propositions leading up to it), and in Proposition 24. Heath [2] provides an English version of this work, though rather than translating, he rewrites Archimedes in a way intended to be more comprehensible to his readers. Selections from the Greek text of Archimedes, with more literal English translations, are provided by Thomas [15]. The first volume of a faithful translation of all of Archimedes's works by Netz [3] has appeared; but this does not contain the works that we are particularly interested in here.

Insight into the discovery of Theorem 2 is given in Archimedes's Method. This work was lost until 1906. Then, in İstanbul, the Danish scholar J.L. Heiberg discovered the Archimedes Palimpsest: a parchment codex of the works of Archimedes that had been washed and reused for writing Christian prayers.

Archimedes does not use the word parabola [2, p. clxvii], but refers to a section of an orthogonal cone. ${ }^{2}$ Let me review what this means, sometimes following also the account of Apollonius [1]. A cone is determined by a circle, called its base, and a point, called the

[^1]apex of the cone, that is not in the plane of the base. The cone is traced out by a straight line, one endpoint of which is at the apex, the other being moved about the circumference of the base. The straight line drawn from the apex to the center of the base is the axis of the cone. A plane containing the axis intersects the cone in an axial triangle. If the axis is perpendicular to the base, then the cone is right. If an axial triangle of a right cone has a right angle at the apex, then the cone is an orthogonal ${ }^{3}$ cone. In such a cone, if a plane is perpendicular to one of the sides of the axial triangle that are about the apex, then the plane cuts the cone in a curve that -following Apollonius-we call a parabola. The intersection of the cutting plane and the axial triangle is the axis of the parabola (which is different from the axis of the cone itself); the intersection of the axis of the parabola and the parabola itself is the vertex of the parabola.

A straight line dropped from the parabola perpendicularly to the axis may be called an ordinate; the part of the axis between the foot of the ordinate and the vertex is the corresponding abscissa. The word abscissa means cut off in Latin, while the word ordinate is related to order, which is used for, among other things, any of the several classical styles of architecture. These orders feature columns standing parallel, like ordinates of a parabola. Consider for example the columns of the Ionic order erected at Priene, Söke, Aydin (which is accessible on a day trip from Sुirince): see Figure 2.


Figure 2. The ordinate $B C$ cuts off from the axis the abscissa $A C$. On the left, Priene

Apollonius's reason for using the term parabola is shown in Proposition 11 of his Conics $[1,15]$. Apollonius also shows that parabolas can be obtained from all cones, not necessarily orthogonal, not necessarily right. What is important for us are the following properties of a parabola, whose proofs can be found in Apollonius.

[^2]1. The squares on two ordinates are in the ratio of the corresponding abscissas [1, I.11]. (Below we shall talk more about what this means.)
2. Suppose a parabola has the vertex $A$, and another point $B$ is chosen on the parabola, and the ordinate $B C$ is drawn. Extend the axis $C A$ beyond $A$ to a point $D$. The straight line $B D$ is tangent to the parabola at $B$ if and only if $A D=C A[1$, I.33, 35] (see Figure 3).


Figure 3. $D B$ is tangent at $B$ if and only if $C A=A D$
3. Every straight line parallel to the axis is a diameter in the following sense. Where this diameter meets the parabola, a tangent can be drawn. If a chord of the parabola is drawn parallel to this tangent, then the diameter bisects the chord [1, I.46]. Half of such a chord is an ordinate with respect to the corresponding diameter, and the squares on two such ordinates are in the ratio of the corresponding abscissas, as in 1 [1, I.49] (see Figure 4).


Figure 4. A new diameter
In the Method, Archimedes solves Problem 1 as follows. We add some straight lines to Figure 1, getting Figure 5. Here $D$ is the midpoint of $B C$, so that (since $B C$ is parallel to the tangent at $A$ ) the straight line $A D$ must be a diameter of the parabola. The tangent to the parabola at $C$ meets $D A$ extended at $E$. Then $A$ is the midpoint of $D E$, by 2 above. A straight line from $B$ parallel to $D A$ meets $C E$ extended at $F$. Extend $C A$ to meet $B F$ at $G$, then extend further to $H$ so that $G H=C G$. The idea now is to consider $C H$ as a lever with fulcrum at $G$. If we conceive of our figures as having weights proportional to their sizes, then we shall show that, if we place the weight of the parabolic segment $A B C$ at $H$, then it will just balance triangle $B C F$ where it is.

Since $A$ is the midpoint of $D E$, also $G$ is the midpoint of $B F$. Let $D F$ be drawn, intersecting $C G$ at $K$. Since $D$ is the midpoint of $B C$, we can conclude that $K$ is the


Figure 5
center of gravity (or centroid) of triangle $B C F$. Then $G K$ is a third of $C G$, hence a third of $G H$. Therefore triangle $B C F$ is balanced by a third of its weight at $H$. If we can show that the parabolic segment balances the triangle, then the segment must be a third of the triangle. But triangle $B C F$ is four times triangle $A B C$ (why?). Then Theorem 2 will follow.

Now, pick a point $L$ at random on the parabola between $B$ and $C$. Let the straight line drawn through $L$ parallel to $B F$ meet $B C$ at $M$ and $C G$ at $N$ and $C F$ at $P$. It remains to show that

$$
\begin{equation*}
L M: M P:: G N: G H . \tag{1}
\end{equation*}
$$

This is the key point. If (1) holds, then $L M$, if its midpoint is placed at $H$, will just balance $M P$. Since $L$ was chosen arbitrarily, we conclude that, if all of the parabolic segment were placed at $H$, then it would balance $B C F$. Now, Archimedes does not find this sort of argument to be sufficiently rigorous. Indeed, in the preface to the Method, he writes

For some things first became clear to me by mechanics, though they had later to be proved geometrically owing to the fact that investigation by this method does not amount to actual proof; but it is, of course, easier to provide the proof when some knowledge of the things sought has been acquired by this method rather than to seek it with no prior knowledge. [15, p.223]

I want to look at the 'actual proof' of Archimedes presently. Meanwhile, let us establish (1). You probably think of it as an equation of fractions: $L M / M P=G N / G H$. That is fine; but (1) simply expresses a relation of proportionality among four magnitudes. Here are Definitions 3-6 from Book V of Euclid's Elements [5, 6].
3. A ratio is a sort of relation in respect of size between two magnitudes of the same kind.
4. Magnitudes are said to have a ratio to one another which are capable, when multiplied, of exceeding one another.
5. Magnitudes are said to be in the same ratio, the first to the second and the third to the fourth, when, if any equimultiples whatever be taken of the first and third, and any equimultiples whatever of the second and fourth, the former equimultiples alike exceed, are alike equal to, or alike fall short of, the latter equimultiples respectively taken in corresponding order.
6. Let magnitudes which have the same ratio be called proportional.

Briefly, 4 means you can't have a ratio between a line and a square: this may be one source of the concern expressed by Archimedes in the quote above. If $a$ and $b$ are magnitudes with a ratio, and so are $c$ and $d$, then by 5 , we may say variously
(1) $a$ is to $b$ in the same ratio that $c$ is to $d$,
(2) $a$ is to $b$ as $c$ is to $d$,
(3) $a: b:: c: d$,
provided that, whenever we take a multiple $n a$ of $a$, and the same multiple $n c$ of $c$, and a multiple $m b$ of $b$, and the same multiple $m d$ of $d$, then

$$
\begin{aligned}
& n a>m b \text { if and only if } n c>n d, \\
& n a=m b \text { if and only if } n c=n d, \\
& n a<m b \text { if and only if } n c<n d .
\end{aligned}
$$

In Books V and VI of the Elements, Euclid goes on to prove the properties of proportionality that we shall need.

To return to our problem. From $L$ draw a straight line parallel to $B C$, meeting $A D$ at $Q$ and $A C$ at $R$. Then by property 1 of the parabola given above,

$$
L Q^{2}: C D^{2}:: A Q: A D .
$$

Since $L Q=M D$, and triangle $A C D$ is similar to $N C M$, while $A R Q$ is similar to $A C D$, we can rewrite the proportion as

$$
N A^{2}: A C^{2}:: A R: A C .
$$

Therefore $N A$ is a mean proportional of $A C$ and $A R$ (see Euclid's VI.13, 20, and 23), so

$$
\begin{aligned}
N A: A C & :: A R: N A \\
& :: N A+A R: N A+A C \\
& :: N R: N C \\
& :: N L: N M .
\end{aligned}
$$

Since $A C=A G$, and $M N=N P$, we obtain

$$
\begin{gathered}
N A: A G:: N L: N M, \\
A G-N A: A G:: N M-N L: N M, \\
N G: A G:: L M: N M, \\
N G: 2 A G:: L M: 2 N M, \\
N G: G C:: L M: M P, \\
N G: G H:: L M: M P,
\end{gathered}
$$

which is (1), as desired.
Archimedes works out a rigorous formulation of this argument in the Quadrature of the parabola, but I prefer now to look at the alternative proof, given for Proposition 24 of that work.

Start over from Figure 1, getting Figure 6. Here $D$ is the midpoint of $B C$ as before,


Figure 6
and $E$ is the midpoint of $A B$. From $E$ a straight line is drawn parallel to $D A$, meeting the parabola at $F$. Draw straight lines $A F$ and $F C$. Then triangle $A C F$ has the same altitude as the parabolic segment in which it is inscribed. Similarly we can find $G$ on the parabola between $A$ and $B$ so that the inscribed triangle $A B G$ has the same height as its parabolic segment. We show that triangles $A C F$ and $A B G$ are together one fourth of triangle $A B C$.

To this end, from $E$ and $F$ we draw parallels to $B C$, meeting $A D$ at $H$ and $K$ respectively. Then

$$
F K^{2}: C D^{2}:: A K: A D .
$$

But $F K=E H$, and

$$
E H: C D:: E A: C A:: 1: 2 .
$$

Therefore $A K$ is one fourth of $A D$. Consequently $K$ is the midpoint of $A H$, and so $L$ is the midpoint of $A E$. Hence triangle $A K L$ is equal to triangle $E F L$. But $E F L$ is one fourth of $A C F$, and $A K L$ is one thirty-second of $A B C$. Therefore $A C F$ is one eighth of $A B C$. Similarly, $A B G$ is one eighth of $A B C$; so $A B G$ and $A C F$ together are one fourth of $A B C$.

We have started with the parabolic segment cut off by the chord $B C$, and we have removed from it the triangle $A B C$. Then we have removed triangles equal to a fourth
of $A B C$. We can continue, removing triangles equal to a sixteenth of $A B C$, and so on. Moreover, at each step, we remove more than one half the remainder of the original parabolic segment (why?).

Therefore, if we continue long enough, we can make the remainder of the parabolic segment less than any pre-assigned area $M$. This is the conclusion of Euclid's Proposition X.1; let us note the proof. The pre-assigned area $M$ is assumed to have a ratio with the parabolic segment, so that, by Definition V. 4 above, some multiple $n M$ of $M$ exceeds the segment. Indeed, Archimedes himself makes the assumption explicit in the preface to the Quadrature of the parabola; it is what we may refer to as the Archimedean axiom:
given [two] unequal areas, the exess by which the greater exceeds the less
can, by being added to itself, be made to exceed any given finite area.
[15, p.231]
If we take away at least half of the parabolic segment, and take $M$ from $n M$, then in the latter case we are taking not more than half; so the former remainder is still less than the latter remainder. If we repeat this process $n-1$ times, then the remainder of the parabolic segment will be less than $M$.

Suppose we have an area that is a third again as large as triangle $A B C$. If we remove triangle $A B C$, what is left is one third. If we then remove one fourth of triangle $A B C$, then what is left is one twelfth of that triangle, which is one fourth of the previous remainder. Continuing, if we remove one fourth of what we last removed, then what remains is one fourth of the previous remainder. Therefore, continuing as far as necessary, we can make the remainder as small as we like. But this is the same process as we described in the original parabolic segment.

Suppose the original parabolic segment is not a third again as large as $A B C$, but is greater. Let the difference be $M$. We can inscribe in the parabolic segment a rectilinear figure which differs from the segment by less than $M$; so it is more than a third greater than $A B C$, which is absurd. There is a similar contradiction if the parabolic segment is less than a third again as large as $A B C$. Theorem 2 now follows.

## 2. Construction of the rational numbers

You know something about the chain

$$
\begin{equation*}
\mathbb{N} \subset \mathbb{Z} \subset \mathbb{Q} \subset \mathbb{R} \tag{2}
\end{equation*}
$$

of number systems. Here $\mathbb{N}$ is the set $\{0,1,2, \ldots\}$ of natural numbers; $\mathbb{Z}$ comprises the integers; $\mathbb{Q}$, the rationals; $\mathbb{R}$, the real numbers. What comes after $\mathbb{R}$ in (2)? It depends on how we think of $\mathbb{R}$. If we think of it as a field, then we might think of $\mathbb{R}$ as included in $\mathbb{C}$, the field of complex numbers. But what if we think of $\mathbb{R}$ as an ordered field? I postpone an answer until § 4. Meanwhile I want to look at how we obtain (2) in the first place.
2.1. The natural numbers. We can understand $\mathbb{N}$ axiomatically. First of all,
(1) it has a distinguished initial element called 0 (zero);
(2) it has a distinguished singulary operation of succession, denoted by $n \mapsto n+1$ : here $n+1$ is the successor of $n$.
I propose to refer to the ordered triple $(\mathbb{N}, 0, n \mapsto n+1)$ as an iterative structure. In general, by an iterative structure, I mean any set that has a distinuished element and a distinguished singulary operation (that is, a function from the set to itself). For example, modular arithmetic involves the iterative structures ${ }^{4}(\mathbb{Z} / n, 0, k \mapsto k+1)$. The iterative structure ( $\mathbb{N}, 0, n \mapsto n+1$ ) is distinguished among iterative structures for satisfying the following axioms.
(1) 0 is not a successor: $0 \neq n+1$.
(2) Succession is injective: if $m+1=n+1$, then $m=n$.
(3) the structure admits proof by induction, in the sense that the only subset $A$ with the following two closure properties is the whole set:
(a) $0 \in A$;
(b) for all $n$, if $n \in A$, then $n+1 \in A$.

These axioms seem to have been discovered originally by Dedekind [4, II, VI (71), p. 67], although they were also written down by Peano [12] and are often known as the Peano axioms. From these axioms, Landau develops the rational, real, and complex numbers rigorously, over the course of a book [10]. I want to do the same here, though more quickly and in a different style. Landau's natural numbers start with 1 , not 0 . Also, Landau does not use the following theorem. The proof is difficult, but the result is very useful.

Theorem 3 (Recursion). For every iterative structure $(A, b, f)$, there is a unique homomorphism to this structure from $(\mathbb{N}, 0, n \mapsto n+1)$ : that is, there is a unique function $h$ from $\mathbb{N}$ to $A$ such that
(1) $h(0)=b$,
(2) $h(n+1)=f(h(n))$ for all $n$ in $\mathbb{N}$.

Proof. I use the set-theoretic conception whereby a function $g$ is just the set of ordered pairs $(x, y)$ such that $g(x)=y$; so if $(x, y)$ and $(x, z)$ belong to $g$, then $y=z$. We now seek $h$ as a particular subset of $\mathbb{N} \times A$.

[^3]Let $B$ be the set whose elements are the subsets $C$ of $\mathbb{N} \times A$ such that, if $(x, y) \in C$, then either
(1) $(x, y)=(0, b)$ or else
(2) $C$ has an element $(u, v)$ such that $(x, y)=(u+1, f(v))$.

Let $R=\bigcup B$; so $R$ is a subset of $\mathbb{N} \times A$. We may say $R$ is a relation from $\mathbb{N}$ to $A$. If $(x, y) \in R$, we may write also

$$
x R y .
$$

Since $(0, b) \in B$, we have $0 R b$. If $n R y$, then $(n, y) \in C$ for some $C$ in $B$, but then $C \cup\{(n+1, f(y))\} \in B$ by definition of $B$, so $(n+1) R f(y)$. Therefore $R$ is the desired function $h$, provided it is a function from $\mathbb{N}$ to $A$. Proving this has two stages.

1. For all $n$ in $\mathbb{N}$, there is $y$ in $A$ such that $n R y$. Indeed, let $D$ be the set of such $n$. Then we have just seen that $0 \in D$, and if $n \in D$, then $n+1 \in D$. By induction, $D=\mathbb{N}$.
2. For all $n$ in $\mathbb{N}$, if $n R y$ and $n R z$, then $y=z$. Indeed, let $E$ be the set of such $n$. Suppose $0 R y$. Then $(0, y) \in C$ for some $C$ in $B$. Since 0 is not a successor, we must have $y=b$, by definition of $B$. Therefore $0 \in E$. Suppose $n \in E$, and $(n+1) R y$. Then $(n+1, y) \in C$ for some $C$ in $B$. Again since 0 is not a successor, we must have $(n+1, y)=(m+1, f(v))$ for some $(m, v)$ in $C$. Since succession is injective, we must have $m=n$. Since $n \in E$, we know $v$ is unique such that $n R v$. Since $y=f(v)$, therefore $y$ is unique such that $(n+1) R y$. Thus $n+1 \in E$. By induction, $E=\mathbb{N}$.

So $R$ is the desired function $h$. Finally, $h$ is unique by induction.
Corollary. For every set $A$ with a distinguished element $b$, and for every function $F$ from $\mathbb{N} \times B$ to $B$, there is a unique function $H$ from $\mathbb{N}$ to $A$ such that
(1) $H(0)=b$,
(2) $H(n+1)=F(n, H(n))$ for all $n$ in $\mathbb{N}$.

Proof. Let $h$ be the unique homomorphism from $(\mathbb{N}, 0, n \mapsto n+1)$ to $(\mathbb{N} \times A,(0, b), f)$, where $f$ is the operation $(n, x) \mapsto(n+1, F(n, x)))$. In particular, $h(n)$ is always an ordered pair. By induction, the first entry of $h(n)$ is always $n$; so there is a function $H$ from $\mathbb{N}$ to $A$ such that $h(n)=(n, H(n))$. Then $H$ is as desired. By induction, $H$ is unique.

The proof of the Recursion Theorem used each of the three Peano axioms; induction alone would not enough. Indeed, if some iterative structure $\mathfrak{A}$ has the property that is guaranteed to ( $\mathbb{N}, 0, n \mapsto n+1$ ) by the Recursion Theorem, then $\mathfrak{A}$ is isomorphic to $\mathbb{N}$ (why?), and consequently $\mathfrak{A}$ satisfies the Peano axioms. ${ }^{5}$ But these axioms are independent. For example, $(\mathbb{Z} / n, 0, k \mapsto k+1)$ satisfies axioms 2 and 3 , but not 1 ; and there are examples satisfying 1 and 3 , but not 2 ; and satisfying 1 and 2 , but not 3 (can you find them?).

Moreover, it is possible to assume the Recursion Theorem and prove the Peano axioms from it.

[^4]We can now use recursion to define the binary operation $(x, y) \mapsto x+y$ of addition, along with the binary operation $(x, y) \mapsto x \cdot y$ or $(x, y) \mapsto x y$ of multiplication, on $\mathbb{N}$. The definitions are:

$$
n+0=n, \quad n+(m+1)=(n+m)+1, \quad n \cdot 0=0, \quad n \cdot(m+1)=n \cdot m+n .
$$

Lemma. For all $n$ and $m$ in $\mathbb{N}$,

$$
0+n=n, \quad(m+1)+n=(m+n)+1 .
$$

Proof. Induction.
Theorem 4. Addition on $\mathbb{N}$ is
(1) commutative: $n+m=m+n$; and
(2) associative: $n+(m+k)=(n+m)+k$.

Proof. Induction and the lemma.
Theorem 5. Addition on $\mathbb{N}$ allows cancellation: if $n+x=n+y$, then $x=y$.
Proof. Induction, and injectivity of succession.
Lemma. For all $n$ and $m$ in $\mathbb{N}$,

$$
0 \cdot n=0, \quad(m+1) \cdot n=m \cdot n+n .
$$

Proof. Induction.
Theorem 6. Multiplication on $\mathbb{N}$ is
(1) commutative: $n m=m n$;
(2) distributive over addition: $n(m+k)=n m+n k$; and
(3) associative: $n(m k)=(n m) k$.

Proof. Induction and the lemma.
Landau proves using induction alone that + and $\cdot$ exist as given by the recursive definitions above. Note however that Theorem 5 needs more than induction (why?). Also, the existence of exponentiation, as an operation $(x, y) \mapsto x^{y}$ such that

$$
n^{0}=1, \quad n^{m+1}=n^{m} \cdot n,
$$

requires more than induction.
The usual ordering $<$ of $\mathbb{N}$ is defined recursively as follows. First note that $m \leqslant n$ means simply $m<n$ or $m=n$. Then the definition of $<$ is:
(1) $m \nless 0$;
(2) $m<n+1$ if and only if $m \leqslant n$.

In particular, $n<n+1$. Really, it is the sets $\{x \in \mathbb{N}: x<n\}$ that are defined by recursion:
(1) $\{x \in \mathbb{N}: x<0\}=\varnothing$;
(2) $\{x \in \mathbb{N}: x<n+1\}=\{x \in \mathbb{N}: x<n\} \cup\{n\}$.

We now have $<$ as a binary relation on $\mathbb{N}$; we must prove that it is an ordering.
Theorem 7. The relation $<$ is transitive on $\mathbb{N}$, that is, if $k<m$ and $m<n$, then $k<n$.

Proof. Induction on $n$.
Lemma. $m \neq m+1$.
Proof. The claim is true when $m=0$, since 0 is not a successor. Suppose the claim is true when $m=k$, that is, $k \neq k+1$. Then $k+1 \neq(k+1)+1$, by injectivity of succession, so the claim is true when $m=k+1$. By induction, the claim is true for all $m$.

Theorem 8. The relation < is irreflexive on $\mathbb{N}: m \nless m$.
Proof. The claim is true when $m=0$, since $m \nless 0$ by definition. Suppose the claim fails when $m=k+1$. This means $k+1<k+1$. Therefore $k+1 \leqslant k$ by definition. By the previous lemma, $k+1<k$. But $k \leqslant k$, so $k<k+1$ by definition. So $k<k+1$ and $k+1<k$; hence $k<k$ by Theorem 7 , that is, the claim fails when $m=k$. By induction, the claim holds for all $m$.

Lemma. (1) $0 \leqslant m$.
(2) If $k<m$, then $k+1 \leqslant m$.

Proof. (1) Induction.
(2) The claim is vacuously true when $m=0$. Suppose it is true when $m=n$. Say $k<n+1$. Then $k \leqslant n$. If $k=n$, then $k+1=n+1<(n+1)+1$. If $k<n$, then $k+1<n+1$ by inductive hypothesis, so $k+1<(n+1)+1$ by transitivity. Thus the claim holds when $m=n+1$. By induction, the claim holds for all $m$.

Theorem 9. The relation $\leqslant$ is total on $\mathbb{N}$ : either $k \leqslant m$ or $m \leqslant k$.
Proof. Induction and the lemma.
Because of Theorems 7,8 , and 9 , the set $\mathbb{N}$ is (strictly) ordered by $<$.
Theorem 10. For all $m$ and $n$ in $\mathbb{N}$, we have $m \leqslant n$ if and only if the equation

$$
\begin{equation*}
m+x=n \tag{3}
\end{equation*}
$$

is soluble in $\mathbb{N}$.
Proof. By induction on $k$, if $m+k=n$, then $m \leqslant n$.
Conversely, if $m \leqslant 0$, then $m=0$ (why?), so $m+0=0$. Suppose the equation $m+x=r$ is soluble whenever $m \leqslant r$, but now $m \leqslant r+1$. If $m=r+1$, then $m+0=r+1$. If $m<r+1$, then $m \leqslant r$, so the equation $m+x=r$ has a solution $k$, and therefore $m+(k+1)=r+1$. Thus the equation $m+x=r+1$ is soluble whenever $m \leqslant r+1$. By induction, for all $n$ in $\mathbb{N}$, if $m \leqslant n$, then (3) is soluble in $\mathbb{N}$.

Theorem 11. (1) If $k<\ell$, then $k+m<\ell+m$.
(2) If $k<\ell$ and $m \neq 0$, then $k m<\ell m$.

Here part 1 is a refinement of Theorem 5 , and part 2 yields the following analogue of Theorem 5 for multiplication.

Corollary. If $k m=\ell m$ and $m \neq 0$, then $k=\ell$.
Theorem 12. $\mathbb{N}$ is well ordered by <: every nonempty set of natural numbers has a least element.

Proof. Suppose $A$ is a set of natural numbers with no least element. Let $B$ be the set of natural numbers $n$ such that, if $m \leqslant n$, then $m \notin A$. Then $0 \in B$, by the last lemma, since otherwise 0 would be the least element of $A$. Suppose $m \in B$. Then $m+1 \in B$, since otherwise $m+1$ would be the least element of $A$. By induction, $B=\mathbb{N}$, so $A=\varnothing$.
2.2. The positive rationals. The integers can be constructed from the natural numbers; and the rationals, from the integers. Since the latter construction is probably more familiar than the former, I begin with it; rather, I construct the positive rational numbers from the positive integers, which we already have: the are just the non-zero natural numbers.

Let us denote the set of positive integers by $\mathbb{N}^{+}$, so

$$
\mathbb{N}^{+}=\mathbb{N} \backslash\{0\} .
$$

We want to define a set

$$
\mathbb{Q}^{+}
$$

of positive rationals. These are certain fractions, namely numbers of the form

$$
\frac{a}{b}
$$

or $a / b$. In particular, $a / b \in \mathbb{Q}^{+}$if and only if $a$ and $b$ are in $\mathbb{N}^{+}$. But what does this mean?

One is taught in school that arithmetic of fractions obeys the following rules:

$$
\begin{equation*}
\frac{a}{b}+\frac{c}{d}=\frac{a d+b c}{b d}, \quad \quad \frac{a}{b} \cdot \frac{c}{d}=\frac{a c}{b d} \tag{4}
\end{equation*}
$$

However, in $\mathbb{Q}^{+}$, one must prove that these rules are valid, because the positive rational number $a / b$ does not uniquely determine the ordered pair $(a, b)$ of positive integers. For example, $1 / 2=2 / 4$, although $(1,2) \neq(2,4)$.

One might try defining a new operation $\oplus$ on $\mathbb{Q}^{+}$by writing down a formula like ${ }^{6}$

$$
\begin{equation*}
\frac{a}{b} \oplus \frac{c}{d}=\frac{a+c}{b+d} . \tag{5}
\end{equation*}
$$

But this implies $1 / 2 \oplus 1 / 3=2 / 5$, while $2 / 4 \oplus 1 / 3=3 / 7$. Since $1 / 2=2 / 4$, while $2 / 5 \neq 3 / 7$, we conclude that $\oplus$ is not well defined. This is a common loose way of speaking. The point is that there is no operation $\oplus$ on $\mathbb{Q}^{+}$with the property required by (5).

On the set $\mathbb{N}^{+} \times \mathbb{N}^{+}$or $\left(\mathbb{N}^{+}\right)^{2}$, let $\sim$ be the relation given by

$$
\begin{equation*}
(a, b) \sim(c, d) \Longleftrightarrow a d=b c \tag{6}
\end{equation*}
$$

One checks easily that $\sim$ is reflexive $(x \sim x)$, symmetric ( $x \sim y$ if and only if $y \sim x$ ) and transitive (if $x \sim y$ and $y \sim z$, then $x \sim z$ ). By definition therefore, $\sim$ is an equivalence relation. If $a$ and $b$ are in $\mathbb{N}^{+}$, we can now define $a / b$ precisely: it is the set of elements of $\left(\mathbb{N}^{+}\right)^{2}$ that are equivalent to $(a, b)$ with respect to $\sim$. Thus

$$
\frac{a}{b}=\left\{(x, y) \in\left(\mathbb{N}^{+}\right)^{2}: a y=b x\right\} .
$$

[^5]Traditionally, $a / b$ is therefore called the equivalence class of $(a, b)$ with respect to $\sim$. Then $\mathbb{Q}^{+}$is the set of such equivalence classes; we might write

$$
\mathbb{Q}^{+}=\left(\mathbb{N}^{+}\right)^{2} / \sim
$$

Now we can check that the rules in (4) are valid. Supposing $a / b=a^{\prime} / b^{\prime}$ and $c / d=c^{\prime} / d^{\prime}$, we have $a b^{\prime}=b a^{\prime}$ and $c d^{\prime}=d c^{\prime}$, so for example

$$
a c b^{\prime} d^{\prime}=a b^{\prime} c d^{\prime}=b a^{\prime} d c^{\prime}=b d a^{\prime} c^{\prime},
$$

and therefore $a c / b d=a^{\prime} c^{\prime} / b^{\prime} d^{\prime}$.
Considering (5) again, note that there is indeed a function $f$ from $\left(\mathbb{N}^{+}\right)^{2} \times\left(\mathbb{N}^{+}\right)^{2}$ to $\mathbb{Q}^{+}$given by

$$
f((x, y),(z, w))=\frac{x+z}{y+w} .
$$

There are just no functions $g$ from $\left(\mathbb{N}^{+}\right)^{2} \times\left(\mathbb{N}^{+}\right)^{2}$ to $\mathbb{Q}^{+} \times \mathbb{Q}^{+}$and $h$ from $\mathbb{Q}^{+} \times \mathbb{Q}^{+}$to $\mathbb{Q}^{+}$such that $f=h \circ g$.

By our construction, a positive integer is not literally a positive rational, because a positive rational is a class of pairs of positive integers. However, the positive integers embed in the positive rationals under the map

$$
x \mapsto \frac{x}{1} .
$$

This embedding respects arithmetic: $a / 1+b / 1=(a+b) / 1$ and $(a / 1)(b / 1)=(a b) / 1$. It also respects the ordering, where we define

$$
\frac{a}{b}<\frac{c}{d} \Longleftrightarrow a d<b c
$$

On $\mathbb{Q}^{+}$there is a binary operation $(x, y) \mapsto x \div y$ of division, given by

$$
\frac{a}{b} \div \frac{c}{d}=\frac{a d}{b c}
$$

one checks that division is indeed well defined. In particular, we have

$$
\frac{a}{1} \div \frac{b}{1}=\frac{a}{b} .
$$

We usually confuse a positive integer $a$ with the positive rational $a / 1$, and for $x \div y$ we write $x / y$. For $1 / a$, we write $a^{-1}$.
2.3. The integers. If $a$ and $b$ are in $\mathbb{N}^{+}$, then the equation

$$
\begin{equation*}
a=b x \tag{7}
\end{equation*}
$$

may or may not have a solution in $\mathbb{N}^{+}$. Suppose it does have a solution; this solution is unique by the corollary to Theorem 11. If $c$ and $d$ are also in $\mathbb{N}^{+}$, and the equation

$$
\begin{equation*}
c=d x \tag{8}
\end{equation*}
$$

has a solution in $\mathbb{N}^{+}$, then it is the same solution that (7) has if and only if

$$
\begin{equation*}
a d=b c \tag{9}
\end{equation*}
$$

(why?). Then $\mathbb{Q}^{+}$is defined to ensure two things:
(1) the equation $(7)$ always has a solution in $\mathbb{Q}^{+}$;
(2) equations (7) and (8) have the same solution in $\mathbb{Q}^{+}$if and only if $(9)$ holds.

The integers can be understood to arise from the natural numbers in the same way, with addition taking the place of multiplication. If $m$ and $n$ are in $\mathbb{N}$ and $m \leqslant n$, then (3) has a solution by Theorem 10; moreover, this solution is unique (why?). In this case, the equation $k+x=\ell$ has the same solution if and only if $m+\ell=n+k$. We use this idea to define an equivalence relation $\approx$ on $\mathbb{N} \times \mathbb{N}$ or $\mathbb{N}^{2}$ by

$$
\begin{equation*}
(m, n) \approx(k, \ell) \Longleftrightarrow m+\ell=n+k \tag{10}
\end{equation*}
$$

The equivalence class of $(m, n)$ with respect to $\approx$ can be denoted by

$$
m \dot{-} n
$$

We denote the set of all such classes by $\mathbb{Z}$; this is the set of integers. So we have

$$
\begin{equation*}
\mathbb{Z}=\mathbb{N}^{2} / \approx \tag{11}
\end{equation*}
$$

One checks that arithmetic can be defined on $\mathbb{Z}$ by

$$
(m \dot{-} n)+(k \dot{-} \ell)=(m+k) \dot{-}(n+\ell), \quad(m \dot{-} n) \cdot(k \dot{-} \ell)=(m k+n \ell) \dot{-}(m \ell+n k)
$$

and a strict ordering, by

$$
m \dot{-} n<k \dot{-} \ell \Longleftrightarrow m+\ell<n+k
$$

We can embed $\mathbb{N}$ in $\mathbb{Z}$ by the map

$$
x \mapsto x-0
$$

For the class $0-n$, we introduce the name

$$
-n
$$

We identify $\mathbb{N}$ with its image in $\mathbb{Z}$. The function $(x, y) \mapsto x-y$ from $\mathbb{N}^{2}$ to $\mathbb{Z}$ extends to the binary operation of subtraction on $\mathbb{Z}$, given by

$$
(m-n)-(k-\ell)=(m+\ell) \dot{-}(n+k)
$$

Lemma. Every element of $\mathbb{Z}$ is either $n$ or $-n$ for some unique $n$ in $\mathbb{N}$.
We can now define multiplication on $\mathbb{Z}$ as usual by

$$
-m \cdot-n=m \cdot n, \quad-m \cdot n=m \cdot-n=-(m \cdot n)
$$

where $m$ and $n$ are in $\mathbb{N}$.
Theorem 13. $\mathbb{Z}$ is an integral domain, that is,
(1) it contains the additive identity 0 and the multiplicative identity 1,
(2) addition is commutative and associative,
(3) equations (3) are always soluble;
(4) multiplication is commutative and associative;
(5) multiplication distributes over addition,
(6) if $x \cdot y=0$, but $x \neq 0$, then $y=0$.
2.4. The rationals. The construction of the rationals in general proceeds just as for the positive rationals in 2.2 , only now the relation $\sim$ defined in (6) must be understood as a relation on $\mathbb{Z} \times(\mathbb{Z} \backslash\{0\})$. The set of classes $m / n$, where $m \in \mathbb{Z}$ and $n \in \mathbb{Z} \backslash\{0\}$, is denoted by $\mathbb{Q}$; so

$$
\mathbb{Q}=(\mathbb{Z} \times(\mathbb{Z} \backslash\{0\})) / \sim .
$$

Lemma. $\mathbb{Q}$ is a field, that is,
(1) it is an integral domain,
(2) equations (7) are always soluble when $b \neq 0$.

Theorem 14. $\mathbb{Q}$ is an ordered field, that is,
(1) it is both a field and an ordered set,
(2) for all nonzero elements $x$, exactly one of $x$ and $-x$ is positive,
(3) the the sum and product of two positive elements is always positive.

On any ordered field, there is an operation $x \mapsto|x|$, where

$$
|a|= \begin{cases}a, & \text { if } a \geqslant 0 \\ -a, & \text { if } a<0\end{cases}
$$

Here $|a|$ is the absolute value of $a$. An absolute value is always positive or 0 .
Theorem 15. $\mathbb{Q}$ embeds uniquely in every ordered field.
Proof. Suppose $K$ is an ordered field. Then $K$ contains elements 0 and 1 , and so there is a unique homomorphism $h$ from ( $\mathbb{N}, 0, n \mapsto n+1$ ) to ( $K, 0, x \mapsto x+1$ ). By induction on $n$, if $m<n$ in $\mathbb{N}$, then $h(m)<h(n)$. Therefore $h$ is injective. Hence we can treat $\mathbb{N}$ as a subset of $K$, and then we can construct $\mathbb{Q}$ inside $K$.

We can now replace $\mathbb{Z}$ with its image in $\mathbb{Q}$.
Theorem 16. The ordering of every ordered field is dense, that is, if $x$ and $y$ are elements of the field, and $x<y$, then there is $z$ in the field such that $x<z<y$.
Proof. Let $z=(x+y) / 2$.
An ordered set is complete if every nonempty subset with an upper bound has a least upper bound. If a subset does have a least upper bound, then it is unique and is called the supremum of the subset. A greatest lower bound is called an infimum.
Theorem 17. In a complete ordered set, every nonempty subset with a lower bound has an infimum.

Proof. Suppose $A$ has an element $b$ and a lower bound. Let $C$ be the set of lower bounds of $A$. Then $C$ is nonempty and has the upper bound $b$. A supremum of $C$ is an infimum of $A$. Indeed, suppose $d$ is a supremum of $C$. If $x<d$, then there is $y$ in $C$ such that $x<y \leqslant d$, so in particular $x \notin A$. Thus $d$ is a lower bound of $A$. In particular, $d \in C$; so $d$ is the greatest of the lower bounds of $A$.

Even though $\mathbb{Q}$ is dense as an ordered set, we shall show that it is not complete.
Theorem 18. The equation

$$
\begin{equation*}
x^{2}=2 \tag{12}
\end{equation*}
$$

has no solution in $\mathbb{Q}$.

Proof. We can use the method of infinite descent. Suppose there were a solution, $n / m$. We may assume $m$ and $n$ are positive integers. Then $n^{2}=2 m^{2}$, so $n$ must be even: say $n=2 k$. So $4 k^{2}=2 m^{2}$, hence $2 k^{2}=m^{2}$. Thus $m / k$ is also a solution to (12). But $0<m<n$. Thus there is no least $n$ in $\mathbb{N}$ such that, for some $m$ in $\mathbb{N}, n / m$ solves (12). Therefore (12) has no solution, by Theorem 12.
Theorem 19. The set $\left\{x \in \mathbb{Q}: x^{2}<2\right\}$ has an upper bound in $\mathbb{Q}$, but no supremum.
Proof. Call the set $A$. It has 2 as an upper bound. Suppose $b$ is an upper bound. We show:
(1) $2<b^{2}$;
(2) $A$ has upper bounds less than $b$.

For 1 , suppose $c \in \mathbb{Q}$ and $c^{2} \leqslant 2$. We show $c$ is not an upper bound of $A$ by finding some positive $h$ in $\mathbb{Q}$ such that $(c+h)^{2}<2$. For all $h$, we have

$$
(c+h)^{2}=c^{2}+2 c h+h^{2}=c^{2}+(2 c+h) h
$$

We have $c^{2}<2$ by Theorem 18, and moreover $c<2$. If also $0<h<1$, then $2 c+h<5$, so

$$
(c+h)^{2}<c^{2}+5 h
$$

Thus, if we require also $h<\left(2-c^{2}\right) / 5$, then $(c+h)^{2}<2$. We can certainly find such $h$; just let $h$ be the lesser of $1 / 2$ and $\left(2-c^{2}\right) / 6$. Therefore $c$ is not an upper bound of $A$. This proves 1 .

For 2 , since 2 is an upper bound for $A$, we may assume $b \leqslant 2$. If $k>0$, then

$$
(b-k)^{2}=b^{2}-2 b k+k^{2}>b^{2}-2 b k \geqslant b^{2}-4 k
$$

Let also $k<\left(b^{2}-2\right) / 4$; then $(b-k)^{2}>2$, so $b-k$ is an upper bound of $A$ that is less than $b$.

## 3. Construction of the real numbers

As a consequence of Theorem 19, we can write $\mathbb{Q}$ as the union of two nonempty disjoint sets $A$ and $B$, where
(1) each element of $A$ is less than each element of $B$;
(2) $A$ has no greatest element;
(3) $B$ has no least element.

Indeed, just let $A=\left\{x \in \mathbb{Q}: x<0 \vee x^{2}<2\right\}$, and $B=\left\{x \in \mathbb{Q}: x>0 \& x^{2}>2\right\}$. See Figure 7. Here the pair $(A, B)$ is an example of a cut in the sense of Dedekind [4, I, IV.,


## Figure 7

pp. 12 f.]. Since $B$ can be obtained from $A$ as $\mathbb{Q} \backslash A$, we may just refer to $A$ as a cut. To be precise then, we define a cut of $\mathbb{Q}$ to be a nonempty proper subset $A$ of $\mathbb{Q}$ such that
(1) every element of $A$ is less than every element of $\mathbb{Q} \backslash A$,
(2) if $A$ has a supremum in $\mathbb{Q}$, then it belongs to $A$.

So a cut may be as in Figure 7 or 8. Note that condition 2 is somewhat arbitrary; one


## Figure 8

might alternatively require any supremum of $A$ not to belong to $A$. We denote the set of cuts of $\mathbb{Q}$ by

## $\mathbb{R}$.

That is, a cut of $\mathbb{Q}$ is precisely a real number.
Dedekind $[4, \mathrm{I}]$ observes that this construction of $\mathbb{R}$ results in the complete ordered field that we want. Details are worked out in Landau [10], and also in Spivak's Calculus [14, ch. 28]. Spivak writes,

The mass of drudgery which this chapter necessarily contains is relieved by one truly first-rate idea
-namely, the idea of what Dedekind calls a cut. My own view is that, in mathematics, if you think something is drudgery, then perhaps you are not looking at it the right way.
3.1. Cuts. In the interest of finding some insight in the construction of $\mathbb{R}$, I note that the notion of a cut makes sense in any ordered set. Suppose $A$ is an ordered set. If $b \in A$, let us define

$$
(b)=\{x \in A: x \leqslant b\}
$$

Note then that

$$
(x) \cup(y)=(\max (x, y))
$$

The collection of sets $(b)$ is thus closed under binary intersection; therefore these sets are basic closed sets for a topology on $A$. In particular, the intersection of any collection of basic closed subsets of $A$ is a closed set in this topology. A cut of $A$ is a non-empty
closed subset of $A$, except for $A$ itself unless it has a maximum element. In particular, every basic closed subset of $A$ is a cut; but the empty set is a closed set that is not a cut. Let us denote by

$$
\bar{A}
$$

the set of all cuts of $A$.
Theorem 20. Let $A$ be an ordered set.
(1) The set $\bar{A}$ is ordered by inclusion:

$$
X<Y \Longleftrightarrow X \subset Y
$$

(2) The map $x \mapsto(x)$ from $A$ to $\bar{A}$ is an embedding of ordered sets.
(3) The ordered set $\bar{A}$ is complete: indeed, if $B$ is a non-empty subset of $\bar{A}$ with an upper bound, then

$$
\sup (B)=\bigcap\{(x): \bigcup B \subseteq(x)\}
$$

(4) If $X \in \bar{A}$, then

$$
\begin{equation*}
X=\sup (\{(x):(x) \subseteq X\})=\bigcap\{(y): X \subseteq(y)\} \tag{13}
\end{equation*}
$$

(5) The ordered set $\bar{A}$ is a completion of $A$ in the sense that, if $f$ is an embedding of $A$ in a complete ordered set $C$, then the map

$$
X \mapsto \sup (\{f(x):(x) \subseteq X\})
$$

is an embedding $\bar{f}$ of $\bar{A}$ in $C$ such that

$$
\begin{equation*}
\bar{f}((x))=f(x) \tag{14}
\end{equation*}
$$

(6) If, further, $C$ is another completion of $A$, then $\bar{f}$ is an isomorphism.

Proof. 2. If $x<y$, then $(x) \subset(y)$.
3. Suppose $B$ is a nonempty subset of $\bar{A}$ with an upper bound. Then it has an upper bound $(c)$, and the set $\bigcap\{(x): \bigcup B \subseteq(x)\}$ is an element $B^{*}$ of $\bar{A}$ and is an upper bound of $B$, and $B^{*} \subseteq(c)$. Every upper bound of $B$ is the intersection of a nonempty collection of such upper bounds $(c)$. Therefore $B^{*}$ is the supremum of $B$.
4. Immediately $\sup (\{(x):(x) \subseteq X\}) \subseteq X$. If $Y$ is an element of $\bar{A}$ such that

$$
\sup (\{(x):(x) \subseteq X\}) \subset Y
$$

then the set $\{(x):(x) \subseteq X\}$ has an upper bound $(d)$ that is less than $Y$, so that $B \subseteq$ $(d) \subset Y$; thus $X \neq Y$.
5. Easily (14) holds. Among elements of $\bar{A}$, if $X \subset Y$, then $X \subseteq(u) \subset(v) \subseteq Y$ for some $u$ and $v$ in $A$, so that

$$
\bar{f}(X) \leqslant \bar{f}((u))<\bar{f}((v)) \leqslant \bar{f}(Y)
$$

Thus $\bar{f}$ is an embedding of ordered sets.
6. The ordered set $\bar{A}$ is the completion of $A$ in the sense that, if $C$ is also a completion, then it embeds in $\bar{A}$ under a map $g$, where

$$
g \circ f(x)=(x)
$$

By (14), we now have $g \circ \bar{f}((x))=(x)$. Say $X \in \bar{A}$. By (13), we have

$$
X \leqslant g \circ \bar{f}(X)
$$

If $X<Y$, then $X \leqslant(z)<Y$ for some $z$ in $A$, so that $g \circ \bar{f}(X) \leqslant g \circ \bar{f}((z))=(z)<Y$; this shows $g \circ \bar{f}(X) \leqslant X$. Thus $g \circ \bar{f}$ is a bijection from $\bar{A}$ to itself. Since $g$ is injective, $\bar{f}$ must be surjective.
3.2. Ordered Abelian groups. An ordered Abelian group is an Abelian group that is ordered so that

$$
\begin{equation*}
x<y \Longrightarrow x+z<y+z \tag{15}
\end{equation*}
$$

equivalently,

$$
x>0 \Longleftrightarrow-x<0, \quad x>0 \& y>0 \Longrightarrow x+y>0
$$

Here those $x$ such that $x>0$ are positive. An ordered Abelian group is complete if it is complete as an ordered set. For example, $\mathbb{Q}$ is an ordered Abelian group, but is not complete, by Theorem 19.

Theorem 21. $\mathbb{Z}$ is a complete ordered Abelian group.
Proof. Theorem 12.
However, $\mathbb{Z}$ is discrete, because it has a least positive element (namely 1 ); but $\mathbb{Q}$ is not discrete.

Theorem 22. Every ordered Abelian group that is not discrete is dense.
Proof. Suppose $G$ is an ordered Abelian group that is not discrete, and $a<b$ in $G$. Then $0<c<b-a$ for some $c$, and then $a<a+c<b$.

The positive part of an ordered Abelian group satisfies

$$
\begin{gather*}
x+y=y+x \\
x+(y+z)=(x+y)+z \\
x<x+y  \tag{16}\\
\exists z(x<y \Rightarrow x+z=y)
\end{gather*}
$$

Let us refer to any ordered set with an addition satisfying these rules as a magnitudinal structure. Indeed, such a structure behaves like a set of magnitudes that is closed under addition and subtraction; see p. 8 above. In particular, $\mathbb{N}^{+}$is magnitudinal. Note that $\mathbb{Z}$ can be obtained from $\mathbb{N}^{+}$just as from $\mathbb{N}$ : we could use

$$
\mathbb{Z}=\left(\mathbb{N}^{+}\right)^{2} / \approx
$$

in place of (11). We need not concern ourselves further with magnitudinal structures as such, because of the following.
Theorem 23. Every magnitudinal structure $B$ is the positive part of an ordered Abelian group, namely $B^{2} / \approx$, where $\approx$ is defined by (10).

Suppose $A$ is an ordered Abelian group. Then on $\bar{A}$ we can define an addition by

$$
X+Y=\sup \{(x+y): x \in X \& y \in Y\}
$$

Recall that a monoid is a group, except that it might not have inverses.

Theorem 24. If $A$ is an ordered Abelian group, then $\bar{A}$ is an Abelian monoid in which $A$ embeds under $x \mapsto(x)$; moreover, the weak form of (15), namely

$$
\begin{equation*}
x \leqslant y \Longrightarrow x+z \leqslant y+z \tag{17}
\end{equation*}
$$

holds in $\bar{A}$.
Proof. Immediately from the definition, addition on $\bar{A}$ is commutative; almost as easily, $X+(0)=X$. We also have

$$
\begin{aligned}
(X+Y)+Z & =\sup (\{(u+z): u \in X+Y \& z \in Z\}) \\
& \geqslant \sup (\{(x+y+z): x \in X \& y \in Y \& z \in Z\})
\end{aligned}
$$

Moreover, if $u \in X+Y$, then

$$
(u) \subseteq \sup (\{(x+y): x \in X \& y \in Y\})
$$

and hence

$$
(u+z) \subseteq \sup (\{(x+y+z): x \in X \& y \in Y\})
$$

this shows

$$
(X+Y)+Z \leqslant \sup (\{(x+y+z): x \in X \& y \in Y \& z \in Z\})
$$

So the two members are equal. Similarly, the right member is equal to $X+(Y+Z)$. So addition is associative. Finally, from the definitions, we have $(x)+(y)=(x+y)$, so $x \mapsto(x)$ is an embedding of monoids, and (17).

The completion of an ordered Abelian group might not be a group. For example, let the free Abelian group $\mathbb{Z} \oplus \mathbb{Z}$ be given the left lexicographic ordering, so that

$$
(a, b)<(c, d) \Longleftrightarrow(a<c \vee(a=c \& b<d))
$$

Then

$$
\cdots<(-1,0)<(-1,1)<\cdots<(0,-1)<(0,0)<(0,1)<\cdots<(1,-1)<(1,0)<\cdots
$$

The only non-principal cuts of this group (that is, cuts not of the form $(x)$ ) are the cuts $c_{n}$, where

$$
c_{n}=\{(k, m): k \leqslant n\}
$$

Identifying $\mathbb{Z} \oplus \mathbb{Z}$ with its image in the completion, we have

$$
\cdots<(-1,0)<\cdots<c_{-1}<\cdots<(0,0)<\cdots<c_{0}<\cdots<(1,0)<\cdots
$$

Also,

$$
c_{m}+c_{n}=\{(k+\ell, r): k \leqslant m \& \ell \leqslant n\}=c_{m+n}
$$

In particular, $c_{m}+c_{0}=c_{m}$, although $c_{0} \neq(0)$. Thus $\overline{\mathbb{Z} \oplus \mathbb{Z}}$ is not a group.
The problem here lies in the possibility that two positive elements of an ordered Abelian group might not have a ratio (p. 8). The problem is eliminated in an Archimedean group. Precisely, an ordered Abelian group is Archimedean if for all positive elements $a$ and $b$ there is an integer $n$ such that $b<n a$. Then $\mathbb{Z}$ and $\mathbb{Q}$ are Archimedean, but not $\mathbb{Z} \oplus \mathbb{Z}$.

Theorem 25. Every complete ordered Abelian group is Archimedean.

Proof. In a non-Archimedean ordered Abelian group, there are two positive elements $a$ and $b$ such that $b$ is an upper bound of $\{n a: n \in \mathbb{Z}\}$. Then $b-a$ is also an upper bound of this set. Therefore this set has no supremum. Hence the ordered group is not complete.

In the proof, the set $\{n a: n \in \mathbb{Z}\}$ does have a supremum $c$ in the completion; but then $a+c=\sup (\{a+n a: n \in \mathbb{Z}\})=a$. Thus (16) fails in the completion.

Theorem 26. Every discrete Archimedean ordered Abelian group is uniquely isomorphic to $\mathbb{Z}$.

Proof. Suppose $A$ is a discrete Archimedean ordered Abelian group, and let $a$ be its least positive element. If $b \in A$, then $|b| \leqslant n a$ for some minimal $n$ in $\mathbb{N}^{+}$, and then

$$
(n-1) a<|b| \leqslant n a, \quad 0<|b|-(n-1) a \leqslant a,
$$

so $|b|=n a$ by minimality of $a$. Thus $a$ generates $A$. Then there is an isomorphism from $A$ to $\mathbb{Z}$ taking $a$ to 1 , and this is the only isomorphism.
Theorem 27. The completion of an Archimedean ordered Abelian group is an ordered Abelian group.
Proof. It suffices to consider a dense Archimedean ordered Abelian group $A$. If $X \in \bar{A}$, define

$$
-X=\sup (\{(-x): X \subseteq(x)\}),
$$

and for $X+-Y$, write $X-Y$ as usual. Then

$$
X<Y \Longleftrightarrow-Y<-X
$$

and

$$
-(x)=(-x),
$$

so

$$
\begin{aligned}
X-X & =\sup (\{(x+y):(x) \leqslant X \&(y) \leqslant-X\}) \\
& =\sup (\{(x+y):(x) \leqslant X \leqslant(-y)\}) \\
& \leqslant \sup (\{(x+y): x+y \leqslant 0\}) \\
& =(0) .
\end{aligned}
$$

We have not used the Archimedean property so far; but we must use it for the reverse inequality, because of Theorem 25. Indeed, if $\bar{A}$ is indeed a group, then it is an ordered group by Theorem 24 .

So, to prove $X-X \geqslant(0)$, suppose $z<0$ in $A$. Then $z<w<0$ for some $w$ in $A$, by density. Since $\bar{A}$ is also Archimedean, for some $n$ in $\mathbb{Z}$,

$$
n(w) \leqslant X<(n-1)(w),
$$

so in particular $(w)-n(w)<-X$; but then

$$
(z)<(w) \leqslant X-X .
$$

Therefore $X-X \geqslant \sup (\{(z): z<0\})=(0)$.
We now know that $\mathbb{R}$ is a complete dense ordered Abelian group.

Theorem 28. For every complete dense ordered Abelian group with positive element a, there is a unique isomorphism $\phi$ from $\mathbb{R}$ to the group that takes 1 to a.

Proof. Let $A$ be the group, and let $\phi$ be the embedding of $\mathbb{Z}$ in $A$ that takes 1 to $a$. We extend $\phi$ to an isomorphism on $\mathbb{R}$.

Suppose $n \in \mathbb{N}^{+}$. The set $\{x \in A: n x \leqslant a\}$ contains 0 and has $a$ as an upper bound; so it has a supremum, which for the moment let us call $b$. We shall show $n b=a$.

Suppose first $n b^{\prime}<a$. By density, there are $x_{k}$ in $A$ such that

$$
n b^{\prime}=x_{0}<x_{1}<\cdots<x_{n-1}<x_{n}=a
$$

Let $c$ be the least of the $x_{k}-x_{k-1}$. Then $n c \leqslant x_{n}-x_{0}=a-n b^{\prime}$, so $n\left(b^{\prime}+c\right) \leqslant a$, which implies that $b^{\prime}$ is not an upper bound of $\{x \in A: n x \leqslant a\}$. Similarly, if $a<n b^{\prime}$, then $b^{\prime}$ is not the supremum of $\{x \in A: n x \leqslant a\}$. So $n b=a$.

For $b$ now, we may use the suggestive notation $a / n$; and we can extend the definition to negative integers by letting $a /-n=-a / n$. We can extend $\phi$ to an embedding of $\mathbb{Q}$ (as an ordered group) by defining

$$
\phi\left(\frac{m}{n}\right)=\frac{m a}{n}
$$

but we must check that this is a valid definition. If $m / n=m^{\prime} / n^{\prime}$ in $\mathbb{Q}$, so that $m n^{\prime}=m^{\prime} n$, then indeed

$$
\frac{m a}{n}=\frac{m^{\prime} a}{n^{\prime}}
$$

since if $m a=n x$, while $m^{\prime} a=n^{\prime} y$, then $m^{\prime} n x=m n^{\prime} y$, so $x=y$.
Finally, suppose $c \in A$, and let $d=\sup (\{\phi(x): \phi(x) \leqslant c\})$. Then $d \leqslant c$. If $d^{\prime}<c$, then $a<n\left(c-d^{\prime}\right)$ for some $n$ in $\mathbb{N}^{+}$, so that $a / n<c-d^{\prime}$, and therefore

$$
d^{\prime}<\frac{m a}{n} \leqslant c
$$

for some $m$ in $\mathbb{Z}$, so that $d^{\prime} \neq d$. Therefore

$$
c=\sup (\{\phi(x): \phi(x) \leqslant c\})
$$

This shows that $A$ is the completion of $\phi[\mathbb{Q}]$. Therefore $\phi$ extends to an isomorphism from $\mathbb{R}$ to $A$ by Theorem 20.
3.3. The complete ordered field of real numbers. Let

$$
\mathbb{R}^{+}=\{X \in \mathbb{R}: X \geqslant(0)\}
$$

Then there is an obvious isomorphism from $\mathbb{R}^{+}$to $\overline{\mathbb{Q}^{+}}$, namely $X \mapsto X \cap \mathbb{Q}^{+}$.
Theorem 29. $\mathbb{R}$ is a complete ordered field, and every complete ordered field is isomorphic to it.

Proof. Since $\mathbb{Q}^{+}$is a dense ordered Abelian group under multiplication, so is $\mathbb{R}^{+}$. On this group, we have

$$
\begin{aligned}
X(Y+Z) & =\sup (\{(x u): x \in X \& u \in Y+Z\}) \\
& \geqslant \sup (\{(x(y+z)): x \in X \& y \in Y \& z \in Z\})
\end{aligned}
$$

Conversely, if $u \in Y+Z$, then

$$
(u) \subseteq \sup (\{(y+z): y \in Y \& z \in Z\})
$$

and hence

$$
(x u) \subseteq \sup (\{(x(y+z)): y \in Y \& z \in Z\})
$$

this shows

$$
X(Y+Z) \leqslant \sup (\{(x(y+z)): x \in X \& y \in Y \& z \in Z\})
$$

So the two members are equal. We have also

$$
\begin{aligned}
X Y+X Z & =\sup (\{(u+v): u \in X Y \& v \in X Z\}) \\
& \geqslant \sup \left(\left\{\left(x y+x^{\prime} z\right): x \in X \& x^{\prime} \in X \& y \in Y \& z \in Z\right\}\right)
\end{aligned}
$$

this becomes equality when we note that, if $u \in X Y$ and $v \in X Z$, then

$$
(u) \subseteq \sup (\{(x y): x \in X \& y \in Y\})
$$

hence

$$
\begin{aligned}
(u+v) & \subseteq \sup (\{(x y+v): x \in X \& y \in Y\}) \\
& \subseteq \sup \left(\left\{\sup \left(\left\{\left(x y+x^{\prime} z\right): x^{\prime} \in X \& Z \in Z\right\}\right): x \in X \& y \in Y\right\}\right)
\end{aligned}
$$

which is just $\sup \left(\left\{\left(x y+x^{\prime} z\right): x \in X \& x^{\prime} \in X \& y \in Y \& z \in Z\right\}\right)$ (why?). But thie last supremum is just $\sup (\{(x y+x z)): x \in X \& y \in Y \& z \in Z\})$ (which is $X(Y+Z)$ ), since $x y+x^{\prime} z \leqslant \max \left(x, x^{\prime}\right) z$.

So multiplication distributes over addition on $\mathbb{R}^{+}$. This extends to all of $\mathbb{R}$ when we define multiplication by 0 and negative numbers in the way taught in school. Thus $\mathbb{R}$ becomes an ordered field. Its uniqueness as such follows from Theorem 28.

Henceforth we may say $\mathbb{R}$ is the complete ordered field.
3.4. Cauchy sequences. Another way to construct $\mathbb{R}$ is by means of Cauchy sequences. First of all, a sequence $\left(a_{n}: n \in \mathbb{N}\right)$ converges to the real number $b$ if, for all positive real numbers $\varepsilon$, there is a natural number $R$ such that, for all $n$ in $\mathbb{N}$, if $n \geqslant R$, then

$$
\left|a_{n}-b\right|<\varepsilon
$$

In this case, we write

$$
\lim _{n \rightarrow \infty} a_{n}=b
$$

or perhaps $\lim \left(a_{n}: n \in \mathbb{N}\right)=b$.
Lemma. A bounded monotone sequence in $\mathbb{R}$ converges.
Proof. Let $\left(a_{n}: n \in \mathbb{N}\right)$ be bounded an increasing, and let $b=\sup \left(\left\{a_{n}: n \in \mathbb{N}\right\}\right)$. Suppose $\varepsilon>0$. Then $b-\varepsilon$ is not an upper bound of $\left\{a_{n}: n \in \mathbb{N}\right\}$, so for some $R$ in $\mathbb{N}$, we have

$$
b-\varepsilon<a_{R} \leqslant b
$$

Since the sequence is increasing, if $n \geqslant R$, we have

$$
b-\varepsilon<a_{R} \leqslant a_{n} \leqslant b
$$

and therefore

$$
\left|a_{n}-b\right|=b-a_{n}<\varepsilon
$$

Thus $\left(a_{n}: n \in \mathbb{N}\right)$ converges to $b$. Similarly, bounded decreasing sequences converge.

A sequence $\left(a_{n}: n \in \mathbb{N}\right)$ of real numbers is a Cauchy sequence if, for every positive real number $\varepsilon$, there is a natural number $R$ such that, for all $m$ and $n$ in $\mathbb{N}$, if $m \geqslant R$ and $n \geqslant R$, then

$$
\left|a_{m}-a_{n}\right|<\varepsilon
$$

For example, let sequences $\left(p_{n}: n \in \mathbb{N}\right)$ and $\left(q_{n}: n \in \mathbb{N}\right)$ be defined recursively by

$$
\begin{aligned}
p_{0} & =1 \\
p_{n+1} & =p_{n}+2 q_{n}
\end{aligned}
$$

$$
q_{0}=1
$$

$$
q_{n+1}=p_{n}+q_{n}
$$

Let $a_{n}=p_{n} / q_{n}$. Then

$$
\left(a_{n}: n \in \mathbb{N}\right)=\left(1, \frac{3}{2}, \frac{7}{5}, \frac{17}{12}, \frac{41}{29}, \ldots\right)
$$

You can show that

$$
p_{n} q_{n+1}-q_{n} p_{n+1}=(-1)^{n+1}
$$

and hence

$$
a_{n+1}-a_{n}=\frac{(-1)^{n}}{q_{n+1} q_{n}}
$$

Since $\left(q_{n}: n \in \mathbb{N}\right)$ is increasing, it follows that

$$
a_{0}<a_{2}<a_{4}<\cdots<a_{5}<a_{3}<a_{1}
$$

and moreover $\left(a_{n}: n \in \mathbb{N}\right)$ is a Cauchy sequence. Finally,

$$
p_{n}{ }^{2}-2{q_{n}}^{2}=(-1)^{n+1},
$$

so $\left(a_{n}: n \in \mathbb{N}\right)$ converges to $\sqrt{ } 2$, which however is not in $\mathbb{Q}$, by Theorem 18 .
Lemma. Every Cauchy sequence in $\mathbb{R}$ is bounded.
Proof. Let $\left(a_{n}: n \in \mathbb{N}\right)$ be a Cauchy sequence. Let $R$ be such that, if $m \geqslant R$ and $n \geqslant R$, then $\left|a_{m}-a_{n}\right| \leqslant 1$. In particular, if $m \geqslant R$, then

$$
\left|a_{m}\right| \leqslant\left|a_{m}-a_{R}\right|+\left|a_{R}\right| \leqslant 1+\left|a_{R}\right| .
$$

Thus each $\left|a_{n}\right|$ is bounded by $\max \left(\left|a_{0}\right|, \ldots,\left|a_{R-1}\right|, 1+\left|a_{R}\right|\right)$.
Theorem 30. Every Cauchy sequence in $\mathbb{R}$ converges.
Proof. Let $\left(a_{n}: n \in \mathbb{N}\right)$ be a Cauchy sequence. Then the sequence is bounded, by the last lemma. In particular, we can define

$$
b_{k}=\sup \left(\left\{a_{n}: n \geqslant k\right\}\right) .
$$

Then $\left(b_{k}: k \in \mathbb{N}\right)$ is bounded (why?) and decreasing, so it converges to some $c$ by the next to last lemma. We have

$$
\left|a_{m}-c\right| \leqslant\left|a_{m}-a_{n}\right|+\left|a_{n}-b_{k}\right|+\left|b_{k}-c\right| .
$$

Let $\varepsilon>0$. There is some $R$ such that, if $k \geqslant R, m \geqslant R$, and $n \geqslant R$, then $\left|a_{m}-a_{n}\right|<\varepsilon / 3$ and $\left|b_{k}-c\right|<\varepsilon / 3$. For all $k$, there is $n$ such that $n \geqslant k$ and $\left|a_{n}-b_{k}\right|<\varepsilon / 3$. Therefore, if $m \geqslant R$, then $\left|a_{m}-c\right|<\varepsilon$. Thus ( $a_{n}: n \in \mathbb{N}$ ) converges.

For the alternative construction of $\mathbb{R}$, let us denote by

$$
\mathbb{Q}^{\mathbb{N}}
$$

the set of functions from $\mathbb{N}$ to $\mathbb{Q}$, that is, rational sequences. This becomes a commutative ring when, writing $a$ for $\left(a_{n}: n \in \mathbb{N}\right)$, we define

$$
(a+b)_{n}=a_{n}+b_{n}, \quad(-a)_{n}=-a_{n}, \quad(a b)_{n}=a_{n} b_{n}
$$

Then $\mathbb{Q}$ embeds in this ring under the map that takes $x$ to the sequence that is identically $x$. We may identify $\mathbb{Q}$ with its image in $\mathbb{Q}^{\mathbb{N}}$. Let $S$ be the set of Cauchy sequences in $\mathbb{Q}^{\mathbb{N}}$. Then $\mathbb{Q} \subseteq S$.
Lemma. $S$ is a sub-ring of $\mathbb{Q}^{\mathbb{N}}$; that is, $S$ contains 0 and 1 and is closed under,+- , and $\cdot$.

Proof. The most difficult part is closure under multiplication. Let $a$ and $b$ be in $S$. By the previous lemma, there is $R$ such that, for all $n$ in $\mathbb{N}$, we have $\left|a_{n}\right| \leqslant R$ and $\left|b_{n}\right| \leqslant R$. Hence

$$
\begin{array}{r}
\left|a_{m} b_{m}-a_{n} b_{n}\right|=\left|a_{m} b_{m}-a_{n} b_{m}+a_{n} b_{m}-a_{n} b_{n}\right| \leqslant\left|a_{m}-a_{n}\right|\left|b_{m}\right|+\left|a_{n}\right|\left|b_{m}-b_{n}\right| \\
\leqslant R\left(\left|a_{m}-a_{n}\right|+\left|b_{m}-b_{n}\right|\right)
\end{array}
$$

Then $a b$ is Cauchy.
By Theorem 30, we have a map $x \mapsto \lim x$, in fact a homomorphism, from $S$ to $\mathbb{R}$. Let $I$ be the kernel, namely the set of sequences in $\mathbb{Q}^{\mathbb{N}}$ that converge to 0 . Then $I$ is an ideal of $S$.

Theorem 31. $S / I \cong \mathbb{R}$ under $x \mapsto \lim x$; in particular, $I$ is a maximal ideal of $S$.
Proof. We just have to show $x \mapsto \lim x$ is surjective. Let $a \in \mathbb{R}$. Then $a=\sup \{x \in$ $\mathbb{Q}: x<a\}$. By the Axiom of Choice, there is a sequence $b$ in $\mathbb{Q}^{\mathbb{N}}$ such that

$$
b_{0}<a, \quad b_{n}+\frac{a-b_{n}}{2}<b_{n+1}<a
$$

Then $\lim b=a$. Thus $S$ maps onto $\mathbb{R}$ under $x \mapsto \lim x$.
Independently of the theorem-that is, without having previously defined $\mathbb{R}$-, one can show that $I$ is a maximal ideal of $S$. Then $S / I$ is a field, and one can show that it is a complete ordered field. Thus there is an alternative construction of $\mathbb{R}$. The construction is carried out more generally in Theorem 37 in the next section.

## 4. Non-Archimedean fields and valuations

In moving from $\mathbb{N}$ to $\mathbb{Z}$ to $\mathbb{Q}$ to $\mathbb{R}$, we achieve the following. The formal sentence

$$
\forall x \forall y \exists z x+z=y
$$

is false in $\mathbb{N}$, but true in $\mathbb{Z}$. The sentence

$$
\forall x \forall y \exists z(x \cdot z=y \vee x=0)
$$

is false in $\mathbb{Z}$, but true in $\mathbb{Q}$. The sentence

$$
\forall x \exists y\left(y^{2}=x \vee x<0\right)
$$

is false in $\mathbb{Q}$, but true in $\mathbb{R}$ (why?). Thus, moving to $\mathbb{R}$ allows us to solve more equations. Is there any advantage to moving beyond $\mathbb{R}$ ?
4.1. Non-Archimedean fields. In any case, we can move beyond $\mathbb{R}$. Let

$$
\mathbb{R}[x]
$$

denote the set of polynomials in $x$ over $\mathbb{R}$ : these have the form

$$
\begin{equation*}
a_{0}+a_{1} x+a_{2} x^{2}+\cdots+a_{n} x^{n} \tag{18}
\end{equation*}
$$

or

$$
\sum_{k=0}^{n} a_{k} x^{k}
$$

where the coefficients $a_{k}$ are in $\mathbb{R}$. Then $\mathbb{R}[x]$ is an integral domain in which $\mathbb{R}$ embeds. Note that, if $m \leqslant n$, then

$$
\sum_{k=0}^{m} a_{k} x^{k}=\sum_{k=0}^{n} b_{k} x^{k} \Longleftrightarrow a_{0}=b_{0} \& \ldots \& a_{m}=b_{m} \& b_{m+1}=0 \& \ldots \& b_{n}=0
$$

Thus, a polynomial is not simply an expression of the form in (18); it is an equivalenceclass of such expressions. However, every polynomial in $x$ can be written uniquely as an infinite series

$$
\sum_{k=0}^{\infty} a_{k} x^{k}
$$

but here all but finitely many of the coefficients $a_{k}$ are 0 .
Just as we construct $\mathbb{Q}$ from $\mathbb{Z}$, so from $\mathbb{R}[x]$ we construct $\mathbb{R}(x)$, the set of rational functions in $x$ over $\mathbb{R}$, consisting of the fractions

$$
\begin{equation*}
\frac{a_{0}+\cdots+a_{n} x^{n}}{b_{0}+\cdots+b_{m} x^{m}} \tag{19}
\end{equation*}
$$

Then $\mathbb{R}(x)$ is a field.
Theorem 32. $\mathbb{R}(x)$ becomes a non-Archimedean ordered field when, for every element as in (19) such that $a_{n} b_{m} \neq 0$, that element is considered positive if and only if $a_{n} b_{m}>0$.

Proof. Easily $\mathbb{R}(x)$ is an ordered field; it is non-Archimedean, since $-a+x>0$, that is, $x>a$, for all $a$ in $\mathbb{Z}$.

Suppose now $K$ is an arbitrary ordered field that includes $\mathbb{R}$. Each element of $K$ that is smaller (in absolute value) than some rational is called finite; each element that is smaller than every nonzero rational is called infinitesimal. Elements of $K$ that are not finite are infinite.

For example, in $\mathbb{R}(x)$, for all $a$ in $\mathbb{Q}$, we have

$$
\begin{equation*}
a<x<x^{2}<x^{3}<\cdots \tag{20}
\end{equation*}
$$

so the positive powers of $x$ are infinite. Hence also, if $a$ is a positive rational, we have

$$
\begin{equation*}
a>\frac{1}{x}>\frac{1}{x^{2}}>\frac{1}{x^{3}}>\cdots>0 \tag{21}
\end{equation*}
$$

so the negative powers of $x$ are infinitesimal.
With $K$ as before, let $R$ be the set of finite elements of $K$, and let $I$ be the set of infinitesimal elements. Then $R$ is a sub-ring of $K$, and $I$ is an ideal of $R$.

The units or multiplicatively invertible elements of a ring $R$ compose a multiplicative group denoted by

$$
\begin{equation*}
R^{\times} \tag{22}
\end{equation*}
$$

In our situation, an element $a$ of $K^{\times}$is infinite if and only if $a^{-1} \in I$. In particular, either $a$ or $a^{-1}$ is finite - belongs to $R$. For this reason, $R$ is called a valuation ring; the reason for the terminology will be seen below. It also follows that every element of $R \backslash I$ is a unit of $R$. Consequently, $I$ is a maximal ideal of $R$ and is moreover the unique maximal ideal of $R$. For this reason, $R$ is called a local ring. (So every valuation ring is a local ring.) Since $I$ is maximal, we know $R / I$ is a field.

Theorem 33. Let $K$ be an ordered field that includes $\mathbb{R}$, and let $R$ be the ring of finite elements of $K$, with maximal ideal I of infinitesimals. Then the quotient map $x \mapsto x+I$ determines an isomorphism from $\mathbb{R}$ onto $R / I$.

Proof. Let $h$ be $x \mapsto x+I$ on $\mathbb{R}$. Then $\operatorname{ker}(h)=I \cap \mathbb{R}$, which is $\{0\}$. Thus $h$ is injective. It remains to show $h$ is surjective onto $R / I$.

Let $a \in R$. Since $a$ is finite, the set $\{x \in \mathbb{R}: x<a\}$ has an upper bound in $\mathbb{R}$, hence a supremum, $a^{\prime}$. We shall show $h\left(a^{\prime}\right)=a+I$. To this end, suppose $b \in \mathbb{R}$, but $h(b) \neq a+I$. This means $b-a$ is not infinitesimal. In particular, for some real number $\delta$, we have

$$
0<\delta<|b-a|
$$

If $b<a$, then $b<b+\delta<a$, so $b$ is not an upper bound of $\{x \in \mathbb{R}: x<a\}$. If $a<b$, then $a<b-\delta$, so $b$ is not the supremum of $\{x \in \mathbb{R}: x<a\}$. In either case, $b \neq a^{\prime}$.

If $a$ and $b$ are arbitrary elements of $K$ such that $a-b \in I$, then $a$ and $b$ are infinitely close, and we write

$$
a \simeq b
$$

By the theorem, if $a$ is finite, then $a$ is infinitesimally close to some unique real number; this number is called the standard part of $a$. In particular, the infinisimals are the elements whose standard part is 0 .

Let us see how this all works in $\mathbb{R}(x)$. The finite elements here are those of the form

$$
\frac{a_{n} x^{n}+\cdots+a_{0}}{b_{n} x^{n}+\cdots+b_{0}}
$$

where $b_{n} \neq 0$. The standard part of this element is $a_{n} / b_{n}$, since

$$
\frac{a_{n} x^{n}+\cdots+a_{0}}{b_{n} x^{n}+\cdots+b_{0}}-\frac{a_{n}}{b_{n}}=\frac{\left(a_{n-1}-a_{n} b_{n-1} / b_{n-1}\right) x^{n-1}+\cdots}{b_{n} x^{n}+\cdots+b_{0}} .
$$

Using the division algorithm taught in school, we can formally compute the quotient of two nonzero elements of $\mathbb{R}[x]$, getting a possibly infinite series

$$
c_{0}+c_{1} x^{-1}+c_{2} x^{-2}+\cdots
$$

or simply

$$
\sum_{n=0}^{\infty} c_{k} x^{-k}
$$

this is a formal power series in $x^{-1}$ with coefficients from $\mathbb{R}$. For example, formally,

$$
\frac{x}{x-1}=1+x^{-1}+x^{-2}+\cdots .
$$

The set of all formal power series in $x^{-1}$ over $\mathbb{R}$ is denoted by $\mathbb{R}\left[\left[x^{-1}\right]\right]$ or rather

$$
\mathbb{R}[[t]],
$$

where $t=x^{-1}$. This set is an integral domain in the obvious way, and its quotient field is denoted by

$$
\mathbb{R}((t)) ;
$$

this is the field of formal Laurent series in $t$ with coefficients from $\mathbb{R}$, namely series

$$
\begin{equation*}
\sum_{n=k}^{\infty} a_{n} t^{n} \tag{23}
\end{equation*}
$$

where $k \in \mathbb{Z}$, and each $a_{n}$ is in $\mathbb{R}$. This field includes $\mathbb{R}(t)$, which is $\mathbb{R}(x)$.
The ordering of $\mathbb{R}(t)$ extends to $\mathbb{R}((t))$. Indeed, let $a$ be the element in (23), and assume $a_{k} \neq 0$. Then $a$ is
(1) positive if and only if $a_{k}>0$,
(2) finite if and only if $k \geqslant 0$,
(3) infinitesimal if and only if $k>0$.

If $a$ is finite, then its standard part is $a_{0}$ (which is 0 if $k>0$ ).
4.2. Valuations. The construction of $\mathbb{R}((t))$ as a field uses only that $\mathbb{R}$ is a field. Let $K$ be an arbitrary field, not necessarily ordered; then we can form the field

$$
K((t))
$$

of formal Laurent series in $t$ with coefficients from $K$. This has the sub-ring

$$
K[[t]]
$$

of formal power series in $t$ with coefficients from $K$. This ring is a valuation ring, with unique maximal ideal $(t)$; here $(t)$ consists of the series $\sum_{n=1}^{\infty} a_{n} t^{n}$ with no constant term.
Theorem 34. $K \cong K[[t]] /(t)$ under $\xi \mapsto \xi+(t)$.

There is another quotient we can form, namely

$$
K((t))^{\times} / K[[t]]^{\times} .
$$

Here

$$
K((t))^{\times}=K((t)) \backslash\{0\}, \quad K[[t]]^{\times}=K[[t]] \backslash(t) .
$$

The quotient map $\xi \mapsto \xi K[[t]]^{\times}$is the $t$-adic valuation. This can be understood by noting that, if $a_{k} \neq 0$, then

$$
\frac{1}{t^{k}} \sum_{n=k}^{\infty} a_{n} t^{n}=\sum_{n=0}^{\infty} a_{k+n} t^{n}
$$

which is in $K[t t]]^{\times}$. We might write

$$
\sum_{n=k}^{\infty} a_{n} t^{n} \equiv t^{k} \quad\left(\bmod K[[t]]^{\times}\right) .
$$

Thus the $t$-adic valuation maps $\langle t\rangle$ (that is, $\left\{t^{n}: n \in \mathbb{Z}\right\}$ ) bijectively onto $K((t))^{\times} / K[[t]]^{\times}$. When we assume

$$
0<\cdots<t^{2}<t<1<t^{-1}<\cdots,
$$

then an ordering is induced on $\{0\} \cup K((t))^{\times} / K[[t]]^{\times}$, and the $t$-adic valuation becomes a map $x \mapsto|x|$ such that
(1) $|x y|=|x||y|$,
(2) $|x|=0$ if and only if $x=0$,
(3) the strong triangle inequality $|x+y| \leqslant \max (|x|,|y|)$ holds.

In general, if $K$ is an arbitrary field, and $\Gamma$ is an ordered field written multiplicatively, and the ordering is extended to $\{0\} \cup \Gamma$ so that $0<v$ for all $v$ in $\Gamma$, and there is a function $x \mapsto|x|$ from $K$ to $\{0\} \cup \Gamma \cup$ so that the three properties just listed hold, then $K$, then the function $x \mapsto|x|$ may be called a strong valuation ${ }^{7}$ on $K$, and $K$ itself, considered with the valuation, may be called a strongly valued field. If, moreover, $\Gamma$ is the positive part of an ordered field, but the triangle inequality holds only in its weaker form

$$
|x+y| \leqslant|x|+|y|,
$$

then $K$ is simply a valued field with respect to the valuation $x \mapsto|x|$.
So every ordered field, such as $\mathbb{R}$, is a valued field with respect to the absolute value function; but also $\mathbb{C}$ is a valued field with respect to this function. The field $K((t))$ is strongly valued with respect to the $t$-adic valuation.

Let $\mathfrak{O}$ be an arbitrary valuation ring with unique maximal ideal $\mathfrak{p}$ and and quotient field $K$. We order the multiplicative group $K^{\times} / \mathfrak{V}^{\times}$by the rule

$$
a \mathfrak{V}^{\times} \leqslant b \mathfrak{V}^{\times} \Longleftrightarrow a / b \in \mathfrak{O},
$$

[^6]and we say 0 is less than all elements of the group. Let the quotient map from $K^{\times}$to $\mathfrak{O}^{\times}$, together with $\{(0,0)\}$ (the function taking 0 to 0 ) be denoted by
$$
\xi \mapsto|\xi|_{\mathfrak{p}}
$$
this is the $\mathfrak{p}$-adic valuation on $K$, and the terminology is justified by the following.
Theorem 35. Let $\mathfrak{O}$ be a valuation ring with maximal ideal $\mathfrak{p}$. With the $\mathfrak{p}$-adic valuation, the quotient field of $\mathfrak{O}$ is a strongly valued field.

Proof. For the strong triangle inequality, if $b \neq 0$, we have

$$
|a+b|_{\mathfrak{p}} \leqslant|b|_{\mathfrak{p}} \Longleftrightarrow \frac{a+b}{b} \in \mathfrak{O} \Longleftrightarrow \frac{a}{b}+1 \in \mathfrak{O} \Longleftrightarrow \frac{a}{b} \in \mathfrak{O} \Longleftrightarrow|a|_{\mathfrak{p}} \leqslant|b|_{\mathfrak{p}}
$$

A valuation ring can be recovered from a strong valuation, which in turn can be recovered from the valuation ring:

Theorem 36. In a strongly valued field $K$, let

$$
\mathfrak{O}=\{x \in K:|x| \leqslant 1\}, \quad \mathfrak{p}=\{x \in K:|x|<1\}
$$

Then $K$ is the quotient field of $\mathfrak{O}$, and $\mathfrak{O}=\left\{x \in K:|x|_{\mathfrak{p}} \leqslant 1\right\}$.
In the notation of the theorem, the ordered group $K^{\times} / \mathfrak{O}^{\times}$is the value group, and the field $\mathfrak{O} / \mathfrak{p}$ is the residue field. So $K((t))$, with the $t$-adic valuation, has the residue field $K$, by Theorem 34. For an arbitrary ordered field extending $\mathbb{R}$, with valuation determined by the infinitesimals, the residue field is (isomorphic to) $\mathbb{R}$, by Theorem 33 .

In an arbitrary strongly valued field $K$, the elements of the subgroup $\{n \cdot 1: n \in \mathbb{Z}\}$ of $K$ take values no greater than 1. But the elements of $K \backslash \mathfrak{O}$ take values greater than 1. For this reason, if $\mathfrak{p} \neq(0)$, the $\mathfrak{p}$-adic valuation is non-Archimedean. By contrast, the absolute value function on a subfield of $\mathbb{R}$ or $\mathbb{C}$ is Archimedean. But what about the absolute value function on a non-Archimedean ordered field? This has non-Archimedean value group, and usually the definition of valuation requires the value group to be Archimedean (or equivalently, to be a subgroup of $\mathbb{R}^{+}$). I am not making this requirement here.

In any valued field, there is the notion of Cauchy sequence and convergent sequence: the definitions are formally the same as for sequences in $\mathbb{R}$. A valued field is complete if every Cauchy sequence of its elements converges. Then $K((t))$ is complete with respect both to the $t$-adic valuation. Also, $\mathbb{R}$ and $\mathbb{C}$ are complete with respect to the absolute value function. Finally, $\mathbb{R}((t))$ is complete both with respect to the $t$-adic valuation and with respect to the absolute value function induced by the ordering in which $t$ is infinitesimal.

Lemma. In a strongly valued field, if $|a|<|b|$, then $|a \pm b|=|b|$.
Proof. Since

$$
|b|=| \pm b|=|a \pm b-a| \leqslant \max (|a \pm b|,|a|)
$$

we have $|b| \leqslant|a \pm b| \leqslant|b|$, so $|b|=|a \pm b|$.
Theorem 37. Every valued field $K$ has a completion, namely a complete valued field $\bar{K}$ in which $K$ embeds, such that any embedding of $K$ in a valued field extends to an embedding of $\bar{K}$ in that valued field.

Proof. Let $R$ consist of the Cauchy sequences of $K$, and let $I$ consist of those sequences that converge to 0 . Then $R$ is a ring with maximal ideal $I$. Indeed, suppose $\left(a_{n}: n \in\right.$ $\mathbb{N}) \in R \backslash I$. Then for some non-zero value $\varepsilon$, for all positive integers $M$, there is an integer $n$ such that $n>M$ and $\left|a_{n}\right| \geqslant \varepsilon$. We may assume $\varepsilon<1$. For some positive integer $N$, if $m>N$ and $n>N$, then $\left|a_{m}-a_{n}\right|<\varepsilon^{2}$. But we can choose $k$ so that $k>N$ and $\left|a_{k}\right| \geqslant \varepsilon$. In this case, if $m>N$, then, in case the valuation is weak, we have

$$
\begin{gathered}
\left|a_{k}\right|-\left|a_{m}\right| \leqslant\left|a_{m}-a_{k}\right|<\varepsilon^{2} \\
\varepsilon \leqslant\left|a_{k}\right|<\left|a_{m}\right|+\varepsilon^{2} \\
\varepsilon-\varepsilon^{2}<\left|a_{m}\right|
\end{gathered}
$$

while if the valuation is strong, then by the lemma, $\varepsilon^{2}<\left|a_{k}\right|=\left|a_{m}\right|=\left|a_{k}\right|$.
In particular, if $m>N$ and $n>N$, then $a_{m} a_{n} \neq 0$, and we have

$$
\left|\frac{1}{a_{n}}-\frac{1}{a_{m}}\right|=\left|\frac{a_{m}-a_{n}}{a_{n} a_{m}}\right|=\frac{\left|a_{m}-a_{n}\right|}{\left|a_{n} a_{m}\right|}
$$

Therefore, if we define

$$
b_{n}= \begin{cases}a_{n}^{-1}, & \text { if } a_{n} \neq 0 \\ 1, & \text { if } a_{n}=0\end{cases}
$$

then $\left(b_{n}: n \in \mathbb{N}\right) \in R$. Now let

$$
c_{n}= \begin{cases}0, & \text { if } a_{n} \neq 0 \\ 1, & \text { if } a_{n}=0\end{cases}
$$

Then $\left(c_{n}: n \in \mathbb{N}\right) \in I$, and $a_{n} b_{n}+c_{n}=1$. Therefore $I$ is indeed a maximal ideal of $R$. We can embed $K$ in the field $R / I$ under the quotient map. We extend the valuation to $R / I$ by letting a Cauchy sequence have the value that its terms eventually reach. (Why is $R / I$ complete, and the completion of $K$ ?)

For example, $K((t))$ is the completion of $K(t)$ with respect to the $t$-adic valuation, and $\mathbb{R}((t))$ is the completion of $\mathbb{R}(t)$ with respect to the absolute value function induced by the ordering in which $t$ is infinitesimal.

The field of rationals has a non-Archimedean completion $\mathbb{Q}_{p}$ for each prime $p$. Indeed, the $p$-adic valuation on $\mathbb{Q}$ is given by

$$
\left|p^{n} \cdot \frac{a}{b}\right|_{p}=\frac{1}{p^{n}}
$$

where $n \in \mathbb{Z}$, and $a$ and $b$ are integers indivisible by $p$. Then $\mathbb{Q}_{p}$ consists of the $p$-adic numbers, namely the formal sums

$$
\sum_{n=k}^{\infty} a_{n} p^{n}
$$

where $k \in \mathbb{Z}$, and $a_{n} \in\{0,1, \ldots, p-1\}$. For example, in $\mathbb{Q}_{p}$,

$$
-1=\sum_{n=0}^{\infty}(p-1) p^{k}
$$

Even though $\mathbb{Q}_{p}$ is of characteristic 0 , its residue field is finite, with $p$ elements.

It is possible for an ordered field to be complete with respect to the absolute value function, simply because there are no Cauchy sequences that are not eventually constant. Indeed, suppose $\kappa$ is an iterative structure that is also well-ordered so that $\alpha<\alpha+1$ for all $\alpha$ in $\kappa$. Let $X$ be a set $\left\{x_{\alpha}: \alpha \in \kappa\right\}$ of variables indexed by $\kappa$. Given a field $K$, we can define a sequence ( $K_{\alpha}: \alpha \in \kappa$ ) of fields by transfinite recursion:
(1) $K_{0}=K$,
(2) $K_{\alpha+1}=K_{\alpha}\left(x_{\alpha}\right)$,
(3) $K_{\beta}=\bigcup\left\{K_{\alpha}: \alpha<\beta\right\}$, if $\beta$ is not a successor or 0 .

Now let

$$
K(X)=\bigcup\left\{K_{\alpha}: \alpha \in \kappa\right\} .
$$

If $K$ is ordered, then $K(X)$ is ordered so that each $x_{\alpha}$ is greater than each element of $K$, and $x_{\alpha}<x_{\beta} \Longleftrightarrow \alpha<\beta$. Suppose that every countable subset of $\kappa$ is bounded in $\kappa$. (It is a set-theoretical fact that such $\kappa$ exist.) Then the only Cauchy sequences in $K(X)$ with respect to the absolute value function are eventually constant. Indeed, if $\left(a_{n}: n \in \mathbb{N}\right)$ is a sequence of elements of $K(X)$, then the set of reciprocals $1 /\left|a_{n}-a_{m}\right|$ of nonzero differences is bounded by some $x_{\alpha}$, and then

$$
\left|a_{n}-a_{m}\right|<1 / x_{\alpha} \Rightarrow a_{n}=a_{m} .
$$

Consequently, $K(X)$ is complete with respect to the absolute value function.
In sum:
(1) The field $\mathbb{R}$ is the unique complete ordered field.
(2) The field $\mathbb{R}$ is complete with respect to the absolute value function, but so too is $\mathbb{C}$.
(3) There are non-Archimedean ordered fields. The completion of one of these with respect to the ordering is never a field. But there is a completion with respect to the absolute value function determined by the ordering, and this completion is always a field.

## 5. Ultrapowers

5.1. Algebra. For notational convenience, if $n \in \mathbb{N}$, let us assume

$$
n=\{x \in \mathbb{N}: x<n\}=\{0, \ldots, n-1\} .
$$

To be precise, we can define $f$ on $\mathbb{N}$ by

$$
f(0)=\varnothing, \quad f(n+1)=f(n) \cup\{f(n)\}
$$

By induction, $f(n)=\{f(0), \ldots, f(n-1)\}$. With more work, one shows $f$ is injective. The image $f[\mathbb{N}]$ of $\mathbb{N}$ under $f$ is denoted by

$$
\omega
$$

So 0 in $\omega$ is $\varnothing$, and $n+1$ is $n \cup\{n\}$. It will be convenient to treat $\omega$ as $\mathbb{N}$.
Let $\Omega$ be a set, and $n \in \omega$. We define

$$
\Omega^{n}
$$

as the set of functions from $n$ (that is, $\{0, \ldots, n-1\}$ ) to $\Omega$. In particular,

$$
\Omega^{0}=\{\varnothing\}=\{0\}=1
$$

A subset of $\Omega^{n}$ is an $n$-ary relation on $\Omega$. A typical element of $\Omega^{n}$ might be denoted by

$$
\left(x^{0}, \ldots, x^{n-1}\right)
$$

or more simply

## $\boldsymbol{x}$.

On $\Omega$ there are just two 0 -ary relations, namely $\varnothing$ and $\{\varnothing\}$, that is, 0 and 1 .
Suppose now $m$ and $n$ are in $\omega$, and

$$
f: m \rightarrow n
$$

(If $n=0$, then $m$ must be 0 .) Then a function from $\Omega^{n}$ to $\Omega^{m}$ is induced, namely

$$
x \mapsto x \circ f
$$

Let us denote this function by

$$
f^{*}
$$

or more precisely $f_{\Omega}^{*}$. Then ${ }^{8}$

$$
f^{*}\left(x^{0}, \ldots, x^{n-1}\right)=\left(x^{f(0)}, \ldots, x^{f(m-1)}\right)
$$

For example, if $f$ is the inclusion of $n$ in $n+1$, then $f^{*}\left(x^{0}, \ldots, x^{n-1}, x^{n}\right)=\left(x^{0}, \ldots, x^{n-1}\right)$, or more simply $f^{*}(\boldsymbol{x}, y)=\boldsymbol{x}$.

In general, we get $X \mapsto f^{*}[X]$ from $\mathscr{P}\left(\Omega^{n}\right)$ to $\mathscr{P}\left(\Omega^{m}\right)$, where

$$
\begin{aligned}
f^{*}[A] & =\left\{f^{*}(\boldsymbol{x}): \boldsymbol{x} \in A\right\} \\
& =\left\{\boldsymbol{y} \in \Omega^{m}: \exists\left(x^{0}, \ldots, x^{n-1}\right)\left(\left(x^{0}, \ldots, x^{n-1}\right) \in A \& \boldsymbol{y}=\left(x^{f(0)}, \ldots, x^{f(m-1)}\right)\right)\right\}
\end{aligned}
$$

[^7]We also have a function $Y \mapsto f_{*}(Y)$ from $\mathscr{P}\left(\Omega^{m}\right)$ to $\mathscr{P}\left(\Omega^{n}\right)$ where

$$
\begin{aligned}
f_{*}(B) & =\left(f^{*}\right)^{-1}[B] \\
& =\left\{\left(x^{0}, \ldots, x^{n-1}\right) \in \Omega^{n}:\left(x^{f(0)}, \ldots, x^{f(m-1)}\right) \in B\right\}
\end{aligned}
$$

If again $f$ is the inclusion of $n$ in $n+1$, then $f_{*}(B)$ can be understood as $B \times \Omega$. If $f$ is a permutation of $n$, then

$$
f^{*}[A]=\left(f^{-1}\right)_{*}(A)
$$

Now the statement of Theorem 42 below makes some sense.
So that the proof makes sense, suppose $R$ is a commutative ring. As in $\S 3$, we obtain the commutative ring

$$
R^{\omega}
$$

If $a$ is an element $\left(a_{n}: n \in \omega\right)$ of this ring, let

$$
\operatorname{supp}(a)=\left\{n \in \omega: a_{n} \neq 0\right\}
$$

the support of $a$. In one case of interest, $R$ is $\mathbb{B}$, the two-element field $\{0,1\}$.
If $X$ and $Y$ are subsets of some set, let

$$
X \Delta Y=(X \backslash Y) \cup(Y \backslash X)
$$

the symmetric difference of $X$ and $Y$.
Theorem 38. The map $x \mapsto \operatorname{supp}(x)$ is a bijection from $\mathbb{B}^{\omega}$ onto $\mathscr{P}(\omega)$. Also

$$
\begin{gathered}
\operatorname{supp}(0)=\varnothing \\
\operatorname{supp}(1)=\omega \\
\operatorname{supp}(x y)=\operatorname{supp}(x) \cap \operatorname{supp}(y) \\
\operatorname{supp}(x+y)=\operatorname{supp}(x) \triangle \operatorname{supp}(y)
\end{gathered}
$$

Thus $\mathscr{P}(\omega)$ inherits from $\omega^{\omega}$ the structure of a ring.
A ring (not necessarily commutative) is called Boolean if in it

$$
\begin{equation*}
x^{2}=x \tag{24}
\end{equation*}
$$

So $\mathbb{B}, \mathbb{B}^{\omega}$, and $\mathscr{P}(\omega)$ are Boolean rings.
Theorem 39. Let $R$ be a Boolean ring. In $R$,

$$
\begin{equation*}
2 x=0 \tag{25}
\end{equation*}
$$

and hence

$$
-x=x
$$

Also $R$ is commutative, and $R$ can be partially ordered by the rule

$$
x \leqslant y \Longleftrightarrow x y=x
$$

Then a nonempty subset $I$ of $R$ is an ideal of $R$ if and only if

$$
\begin{gather*}
x \in I \& y \in I \Longrightarrow x+y \in I \\
x \in I \& y \leqslant x \Longrightarrow y \in I \tag{26}
\end{gather*}
$$

All prime ideals of $R$ are maximal, and and ideal $I$ is maximal if and only if

$$
x \in I \Longleftrightarrow x+1 \notin I
$$

Proof. For (25), compute

$$
2 x=(2 x)^{2}=4 x^{2}=4 x .
$$

For commutativity then, compute

$$
\begin{gathered}
x+y=(x+y)^{2}=x^{2}+x y+y x+y^{2}=x+x y+y x+y, \\
0=x y+y x .
\end{gathered}
$$

Immediately from the definitions, $x \leqslant x$. If $x \leqslant y$ and $y \leqslant x$, then $x=x y=y x=y$. If $x \leqslant y$ and $y \leqslant z$, then $x z=x y z=x y=x$, so $x \leqslant z$. Thus $\leqslant$ partially orders $R$.

For the characterization of ideals, note that (26) is equivalent to $x \in I \Longrightarrow x z \in I$.
From (24), we get

$$
x(x-1)=0,
$$

so in every Boolean integral domain, the only elements are 0 and 1 . In short, every Boolean integral domain is a field, so prime ideals of $R$ are maximal. Moreover, an ideal $I$ of $R$ is maximal if and only if $R / I$ is the disjoint union of two cosets, $I$ and $1+I$; this yields the characterization of maximal ideals.

Corollary. A subset $M$ of $\mathscr{P}(\boldsymbol{\omega})$ is a maximal ideal if and only if
(1) $x \in M \& y \in M \Longrightarrow x \cup y \in M$,
(2) $x \in M \& y \subseteq x \Longrightarrow y \in M$,
(3) $x \in M \Longleftrightarrow \omega \backslash x \notin M$.

A principal ideal $(A)$ of $\mathscr{P}(\boldsymbol{\omega})$ is maximal if and only if $A=\omega \backslash\{n\}$ for some $n$ in $\omega$. A maximal ideal of $\mathscr{P}(\boldsymbol{\omega})$ is non-principal if and only if it contains all finite subsets of $\omega$.

For example, if $n \in \omega$, then the principal ideal $(\omega \backslash\{n\})$, namely $\{x \in \mathscr{P}(\omega): n \notin x\}$, is a maximal ideal of $\mathscr{P}(\boldsymbol{\omega})$.
Theorem 40. Let $K$ be a field. The function $X \mapsto \operatorname{supp}[X]$ gives a one-to-one correspondence between the ideals of $K^{\omega}$ and the ideals of $\mathscr{P}(\omega)$.
Proof. Suppose $I$ is an ideal of $K^{\omega}$, and $a \in I$.
(1) If $\operatorname{supp}(b) \subseteq \operatorname{supp}(a)$, then $b \in I$, since $b=c a$, where

$$
c_{n}= \begin{cases}a_{n}{ }^{-1}, & \text { if } a_{n} \neq 0, \\ 0, & \text { if } a_{n}=0\end{cases}
$$

(2) If $b \subseteq \operatorname{supp}(a)$, then $b=\operatorname{supp}(c)$ for some $c$ in $I$.

The first observation shows $\operatorname{supp}^{-1}[\operatorname{supp}[I]]=I$; thus, $X \mapsto \operatorname{supp}[X]$ is injective on ideals. With the second observation, along with the identities

$$
\begin{gather*}
\operatorname{supp}(x) \cap \operatorname{supp}(y)=\operatorname{supp}(x y)  \tag{27}\\
\operatorname{supp}(x) \Delta \operatorname{supp}(y) \subseteq \operatorname{supp}(x+y)
\end{gather*}
$$

we can conclude that $\operatorname{supp}[I]$ is an ideal of $\mathscr{P}(\boldsymbol{\omega})$. Finally, if $J$ is an ideal of $\mathscr{P}(\boldsymbol{\omega})$, then (27) and

$$
\operatorname{supp}(x+y) \subseteq \operatorname{supp}(x) \cup \operatorname{supp}(y)=\operatorname{supp}(x) \Delta \operatorname{supp}(y) \Delta \operatorname{supp}(x y),
$$

along with surjectivity of $x \mapsto \operatorname{supp}(x)$, show that $\operatorname{supp}^{-1}[J]$ is an ideal.

Suppose $M$ is a maximal ideal of $\mathscr{P}(\boldsymbol{\omega})$. Then $\operatorname{supp}^{-1}[M]$ is a maximal ideal of $K^{\omega}$; let us denote this ideal also by $M$. We can form the quotient

$$
K^{\omega} / M
$$

which must be a field; it is called an ultrapower of $K$. The diagonal map

$$
x \mapsto(x: n \in \omega)+M
$$

is an embedding of $K$ in $K^{\omega} / M$; we shall identify $K$ with its image in $K^{\omega} / M$.
If $a$ and $b$ are in $K^{\omega}$, and $a+M=b+M$, let us write also

$$
\begin{equation*}
a \equiv b \quad(\bmod M), \tag{28}
\end{equation*}
$$

or simply $a \equiv b$. The elements of $M$ (as an ideal of $\mathscr{P}(\boldsymbol{\omega}))$ can be thought of as small. Then (28) holds if and only if the set $\left\{n \in \omega: a_{n} \neq b_{n}\right\}$ of indices where $a$ and $b$ differ is small. This definition makes no use of the algebraic structure of $K$. So $K$ can be just a set, although in § 6 we shall be interested only in the case where $K$ is the complete ordered field $\mathbb{R}$.

Theorem 41. Let $K$ be an infinite set, and $M$ a maximal ideal of $\mathscr{P}(\boldsymbol{\omega})$. Then the diagonal embedding of $K$ in $K^{\omega} / M$ is surjective if and only if $M$ is principal.

Proof. Suppose the element $a$ of $K^{\omega}$ is injective, so that, if $m \neq n$, then $a_{m} \neq a_{n}$. Then $a+M$ is in the image of the diagonal embedding if and only if $M$ is principal.

We may henceforth assume that $M$ is a non-principal maximal ideal of $\mathscr{P}(\omega)$, though we shall not actually use the assumption until § 6 . We have a bijection

$$
\left(\left(x_{k}^{0}: k \in \omega\right), \ldots,\left(x_{k}^{n-1}: k \in \omega\right)\right) \mapsto\left(\left(x_{k}^{0}, \ldots, x_{k}^{n-1}\right): k \in \omega\right)
$$

from $\left(K^{\omega}\right)^{n}$ onto $\left(K^{n}\right)^{\omega}$; we may write the bijection more simply as

$$
\left(x^{0}, \ldots, x^{n-1}\right) \mapsto\left(\boldsymbol{x}_{k}: k \in \omega\right)
$$

So a plainface $x$ or $x^{j}$, with a superscript at most, is an element of $K^{\omega}$ or $K$, while a boldface $\boldsymbol{x}_{k}$, with a subscript at most, is an element of $K^{n}$; but $x_{k}^{j}$, with both superscripts and subscripts, is in $K$. Instead of writing

$$
\left(x^{0}+M, \ldots, x^{n-1}+M\right)
$$

we may write simply

$$
\left(x^{0}, \ldots, x^{n-1}\right)+M
$$

and instead of

$$
x^{0} \equiv y^{0} \& \cdots \& x^{n-1} \equiv y^{n-1} \quad(\bmod M),
$$

we may write

$$
\left(x^{0}, \ldots, x^{n-1}\right) \equiv\left(y^{0}, \ldots, y^{n-1}\right) \quad(\bmod M)
$$

If $S \subseteq K^{n}$, we define

$$
\begin{equation*}
{ }^{*} S=\left\{\left(x^{0}, \ldots, x^{n-1}\right)+M:\left(\boldsymbol{x}_{k}: k \in \omega\right) \in S^{\omega}\right\} . \tag{29}
\end{equation*}
$$

Thus we have a function $X \mapsto{ }^{*} X$ from $\mathscr{P}\left(K^{n}\right)$ to $\mathscr{P}\left({ }^{*} K^{n}\right)$ for each $n$ in $\omega$. Also

$$
\begin{equation*}
{ }^{*} K=K^{\omega} / M . \tag{30}
\end{equation*}
$$

Lemma. Let $S \subseteq K^{n}$. Then

$$
\left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*} S \Longleftrightarrow\left\{k \in \omega: \boldsymbol{x}_{k} \notin S\right\} \in M
$$

Proof. Suppose $\left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*} S$. There is $\left(y^{0}, \ldots, y^{n-1}\right)$ in $\left(K^{\omega}\right)^{n}$ such that $\left(x^{0}, \ldots, x^{n-1}\right) \equiv\left(y^{0}, \ldots, y^{n-1}\right)$ and each $\boldsymbol{y}_{k}$ is in $S$. Then

$$
\begin{aligned}
\left\{k \in \omega: \boldsymbol{x}_{k} \notin S\right\} & \subseteq\left\{k \in \omega: x_{k}^{0} \neq y_{k}^{0} \vee \cdots \vee x_{k}^{n-1} \neq y_{k}^{n-1}\right\} \\
& =\left\{k \in \omega: x_{k}^{0} \neq y_{k}^{0}\right\} \cup \cdots \cup\left\{k \in \omega: x_{k}^{n-1} \neq y_{k}^{n-1}\right\} .
\end{aligned}
$$

Each of the sets $\left\{k \in \omega: x_{k}^{j} \neq y_{k}^{j}\right\}$ is in $M$, so their union is, and therefore $\left\{k \in \omega: \boldsymbol{x}_{k} \notin\right.$ $S\} \in M$, by the corollary to Theorem 39 .

Now suppose conversely $\left\{k \in \omega: x_{k} \notin S\right\} \in M$. Then in particular $S \neq \varnothing$. Pick some $\boldsymbol{y}$ in $S$, and define

$$
z_{k}= \begin{cases}\boldsymbol{x}_{k}, & \text { if } \boldsymbol{x}_{k} \in S, \\ \boldsymbol{y}, & \text { if } \boldsymbol{x}_{k} \notin S\end{cases}
$$

Then $\left(x^{0}, \ldots, x^{n-1}\right)+M=\left(z_{0}, \ldots, z^{n-1}\right)+M$, which is in ${ }^{*} S$.
Theorem 42. Let $K$ be a set, and let the functions $X \mapsto{ }^{*} X$ from $\mathscr{P}\left(K^{n}\right)$ to $\mathscr{P}\left({ }^{*} K^{n}\right)$ be as given in (29). Then

$$
\begin{equation*}
{ }^{*}\{(x, x): x \in K\}=\left\{(x, x): x \in{ }^{*} K\right\} ; \tag{31}
\end{equation*}
$$

for all $n$ in $\omega$ and all subsets $S$ and $T$ of $K^{n}$,

$$
\begin{gather*}
{ }^{*} S \cap K^{n}=S,  \tag{32}\\
{ }^{*}\left(K^{n} \backslash S\right)={ }^{*} K^{n} \backslash{ }^{*} S,  \tag{33}\\
{ }^{*}(S \cap T)={ }^{2} S \cap{ }^{*} T \tag{34}
\end{gather*}
$$

for all $m$ and $n$ in $\omega$, all $f$ from $m$ to $n$, and all subsets $S$ of $K^{n}$ and $T$ of $K^{m}$,

$$
\begin{align*}
& *  \tag{35}\\
& *\left(f^{*}[S]\right)=f^{*}\left[{ }^{*} S\right],  \tag{36}\\
& *\left(f_{*}(T)\right)\left.=f_{*}{ }^{*} T\right) .
\end{align*}
$$

(More precisely, the last equations are ${ }^{*}\left(f_{K}^{*}[S]\right)=f_{* K}^{*}\left[{ }^{*} S\right]$ and ${ }^{*}\left(f_{*}^{K}(T)\right)=f_{*}^{* K}\left({ }^{*} T\right)$.)
Proof. For (31), we have

$$
\begin{aligned}
(x, y)+M \in^{*}\{(x, x): x \in K\} & \Longleftrightarrow\left\{k \in \omega: x_{k} \neq y_{k}\right\} \in M \\
& \Longleftrightarrow x+M=y+M .
\end{aligned}
$$

For (32), we easily have $S \subseteq{ }^{*} S \cap K^{n}$. Suppose conversely $\left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*} S \cap K^{n}$. Then

$$
\left\{k \in \omega: \boldsymbol{x}_{k} \notin S\right\} \in M,
$$

and also, for some $\boldsymbol{y}$ in $K^{n}$, we have $\left\{k \in \boldsymbol{\omega}: \boldsymbol{x}_{k} \neq \boldsymbol{y}\right\} \in M$. Since

$$
\{k \in \omega: \boldsymbol{y} \notin S\} \subseteq\left\{k \in \omega: \boldsymbol{y} \neq \boldsymbol{x}_{k} \vee \boldsymbol{x}_{k} \notin S\right\}=\left\{k \in \omega: \boldsymbol{y} \neq \boldsymbol{x}_{k}\right\} \cup\left\{k \in \omega: \boldsymbol{x}_{k} \notin S\right\},
$$

we can conclude $\{k \in \omega: \boldsymbol{y} \notin S\} \in M$. (In particular, the set must be empty.) Hence $\boldsymbol{y} \in S$, so $\left(x^{0}, \ldots, x^{n-1}\right)+M \in S$.

For (33), we have

$$
\begin{aligned}
\left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*}\left(K^{n} \backslash S\right) & \Longleftrightarrow\left\{k \in \omega: x_{k} \notin K^{n} \backslash S\right\} \in M \\
& \Longleftrightarrow\left\{k \in \omega: x_{k} \notin S\right\} \notin M \\
& \Longleftrightarrow\left(x^{0}, \ldots, x^{n-1}\right)+M \notin{ }^{*} S \\
& \Longleftrightarrow\left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*} K^{n} \backslash{ }^{*} S .
\end{aligned}
$$

For (34), we have

$$
\begin{aligned}
\left(x^{0}, \ldots, x^{n-1}\right)+M \in^{*}(S \cap T) & \Longleftrightarrow\left\{k \in \omega: \boldsymbol{x}_{k} \notin S \cap T\right\} \in M \\
\Longleftrightarrow & \left\{k \in \omega: \boldsymbol{x}_{k} \notin S\right\} \cup\left\{k \in \omega: \boldsymbol{x}_{k} \notin T\right\} \in M \\
\Longleftrightarrow & \left\{k \in \omega: \boldsymbol{x}_{k} \notin S\right\} \in M \&\left\{k \in \omega: \boldsymbol{x}_{k} \notin T\right\} \in M \\
\Longleftrightarrow & \left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*} S \& \\
& \left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*} T \\
\Longleftrightarrow & \left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*} S \cap^{*} T .
\end{aligned}
$$

For (35), we may assume $S \neq \varnothing$, since ${ }^{*} \varnothing=\varnothing$. Let $\left(x^{0}, \ldots, x^{n-1}\right) \in\left(K^{\omega}\right)^{m}$. There is $\left(\boldsymbol{y}_{k}: k \in \omega\right)$ in $S^{\omega}$ such that, for all $k$ in $\omega$, if $\boldsymbol{x}_{k} \in f^{*}[S]$, then $\boldsymbol{x}_{k}=f^{*}\left(\boldsymbol{y}_{k}\right)$. Hence

$$
\begin{aligned}
\left(x^{0}, \ldots, x^{m-1}\right)+M \in^{*}\left(f^{*}[S]\right) & \Longleftrightarrow\left\{k \in \omega: \boldsymbol{x}_{k} \notin f^{*}[S]\right\} \in M \\
& \Longleftrightarrow\left\{k \in \omega: \boldsymbol{x}_{k} \neq f^{*}\left(\boldsymbol{y}_{k}\right)\right\} \in M \\
& \Longleftrightarrow\left(x^{0}, \ldots, x^{m-1}\right)+M=f^{*}\left(y^{0}+M, \ldots, y^{n-1}+M\right) \\
& \Longleftrightarrow\left(x^{0}, \ldots, x^{m-1}\right)+M \in f^{*}\left[{ }^{*} S\right] .
\end{aligned}
$$

Finally, for (36), we have

$$
\begin{aligned}
& \left(x^{0}, \ldots, x^{n-1}\right)+M \in{ }^{*}\left(f_{*}(T)\right) \\
\Longleftrightarrow & \left\{k \in \omega: \boldsymbol{x}_{k} \notin f_{*}(T)\right\} \in M \\
\Longleftrightarrow & \left\{k \in \omega: f^{*}\left(\boldsymbol{x}_{k}\right) \notin T\right\} \in M \\
\Longleftrightarrow & \left\{k \in \omega: f^{*}\left(\boldsymbol{x}_{k}\right) \neq \boldsymbol{y}_{k}\right\} \in M \text { for some }\left(\boldsymbol{y}_{k}: k \in \omega\right) \text { in } T^{\omega} \\
\Longleftrightarrow & f^{*}\left(x^{0}+M, \ldots, x^{m-1}+M\right)=\left(y^{0}, \ldots, y^{n-1}\right)+M \text { for some }\left(\boldsymbol{y}_{k}: k \in \omega\right) \text { in } T^{\omega} \\
\Longleftrightarrow & \left(x^{0}, \ldots, x^{m-1}\right)+M \in f_{*}\left({ }^{*} T\right) .
\end{aligned}
$$

5.2. Logic. It will follow from Theorem 42 that $K$ and * $K$ agree on sentences of firstorder logic. This result is stated formally as Theorem 45 , in § 6 below, for the case $K=\mathbb{R}$; but the general claim has the same proof, and the preliminary work done now will be in terms of an arbitrary set $K$.

If $S \subseteq K^{n}$, then $S$ can be understood as a name for:
(1) itself, in $K$,
(2) ${ }^{*} S$, in ${ }^{*} K$.

We can express this more symbolically by

$$
S^{K}=S, \quad S^{*} K={ }^{*} S
$$

An atomic formula is a string

$$
S t^{0} \cdots t^{n-1},
$$

where $S \subset K^{n}$, and each $t^{k}$ is either a variable or an element of ${ }^{*} K$. In case $n=2$, we customarily write

$$
t^{0} S t^{1}
$$

instead of $S t^{0} t^{1}$.
The formulas, simply, are defined recursively:
(1) atomic formulas are formulas,
(2) if $\phi$ is a formula, then so is the negation $\neg \phi$,
(3) if $\phi$ and $\psi$ are formulas, then so is the conjunction $(\phi \& \psi)$,
(4) if $\phi$ is a formula, and $x$ is a variable, then so is the instantiation ${ }^{9} \exists x \phi$ is a formula.
The following is sometimes overlooked in expositions of logic; but it is needed to allow recursive definitions on the set of formulas.

Theorem 43 (Unique Readability). Every formula is uniquely an atomic formula, a negation, a conjunction. or an instantiation. Every conjunction is ( $\phi \& \psi$ ) for some unique formulas $\phi$ and $\psi$.

We can introduce the other customary symbols as abbreviations:
(1) $(\phi \Rightarrow \psi)$ means $\neg(\phi \& \neg \psi)$,
(2) $(\phi \vee \psi)$ means $(\neg \phi \Rightarrow \psi)$,
(3) $(\phi \Leftrightarrow \psi)$ means $((\phi \Rightarrow \psi) \&(\psi \Rightarrow \phi))$,
(4) $\forall x \phi$ means $\neg \exists x \neg \phi$.

Whether a variable is free in a formula is defined recursively:
(1) all variables in an atomic formula are free,
(2) the free variables of $\neg \phi$ are those of $\phi$,
(3) the free variables of $(\phi \& \psi)$ are those of $\phi$ or $\psi$,
(4) the free variables of $\exists x \phi$ are those of $\phi$, except $x$.

If the free variables of a formula are all on the list $\left(x^{0}, \ldots, x^{n-1}\right)$, then the formula can be called $n$-ary. In this case, if $n \leqslant r$, then the formula is also $r$-ary. If we want to understand a formula $\phi$ as $n$-ary, we may write it as $\phi\left(x^{0}, \ldots, x^{n-1}\right)$.

Suppose $t^{k}$ is in ${ }^{*} K$ or is a variable for each $k$ in $\omega$. For each $n$-ary formula $\theta$, a formula $\theta\left(t^{0}, \ldots, t^{n-1}\right)$ is defined. The definition is recursive:
(1) If $\theta$ is atomic, then $\theta(\boldsymbol{t})$ is the result of replacing each $x^{k}$ with $t^{k}$.
(2) If $\theta$ is $\neg \phi$, then $\theta(\boldsymbol{t})$ is $\neg \psi$, where $\psi$ is $\phi(\boldsymbol{t})$.
(3) If $\theta$ is $(\phi \& \psi)$, then $\theta(\boldsymbol{t})$ is $(\phi(\boldsymbol{t}) \& \psi(\boldsymbol{t}))$.
(4) If $\theta$ is $\exists x^{\ell} \phi$, then we can understand $\phi$ as $r$-ary, where $r=\max (\ell+1, n)$. In this case, $\theta(\boldsymbol{t})$ is $\exists x^{\ell} \psi$, where $\psi$ is $\phi(\boldsymbol{u})$, where

$$
u^{k}= \begin{cases}x, & \text { if } k=\ell \\ t^{k}, & \text { if } k \neq \ell\end{cases}
$$

[^8]The case $n=0$ is not excluded; in this case, $\theta\left(t^{0}, \ldots, t^{n-1}\right)$ is simply $\theta$.
The parameters of a formula are the (names of) elements of * $K$ that appear in the formula. A sentence is a formula with no free variables, namely a 0 -ary or nullary formula.

A sentence $\sigma$ with parameters from $K$ may be true in $K$, in which case we write

$$
K \vDash \sigma ;
$$

otherwise $\sigma$ is false in $K$, and we write

$$
K \not \models \sigma .
$$

The definition is recursive:
(1) $K \vDash S a^{0} \cdots a^{n-1}$ if and only if $\left(a^{0}, \ldots, a^{n-1}\right) \in S$.
(2) $K \vDash \neg \sigma$ if and only if $K \nvdash \sigma$.
(3) $K \vDash(\sigma \& \tau)$ if and only if $K \vDash \sigma$ and $K \vDash \tau$.
(4) $K \vDash \exists x \phi$ if and only if, assuming $\phi$ is $n$-ary, there is $\boldsymbol{a}$ in $K^{n}$ such that $K \vDash \phi(\boldsymbol{a})$. All of the foregoing holds also with $K$ replaced by ${ }^{*} K$.

The definition of truth shows why formulas as we have defined them are more precisely called formulas of first-order logic. In our formulas, variables stand only for elements of $K$. If we allowed variables standing for relations on $K$, then our formulas would be second order. The third of the Peano axioms in $\S 2.1$ is second order; so is the definition of completeness of an ordered field. In $\S 6$ we shall note that there is no first-order definition of $\mathbb{N}$ or $\mathbb{R}$.

If $S=\{(x, x): x \in K\}$, then, instead of $t^{0} S t^{1}$, we may write

$$
t^{0}=t^{1}
$$

Then $K \vDash a^{0}=a^{1}$ if and only if $a^{0}=a^{1}$; and likewise in ${ }^{*} K$, by (32). An $n$-ary formula $\phi$ defines an $n$-ary relation on ${ }^{*} K$, namely $\left\{\boldsymbol{a} \in{ }^{*} K^{n}:{ }^{*} K \vDash \phi(\boldsymbol{a})\right\}$; this relation can be denoted by

$$
\phi^{*} K .
$$

In case $\sigma$ is nullary, we have $\sigma^{*} K=\left\{x \in\{0\}\right.$ : $\left.{ }^{*} K \vDash \sigma\right\}$, so that

$$
{ }^{*} K \vDash \sigma \Longleftrightarrow \sigma^{*} K=1
$$

If the parameters of a formula all come from $K$, then the formula similarly defines a relation on $K$, denoted by $\phi^{K}$.
Theorem 44. Let $\theta$ be a formula with parameters from $K$. Then

$$
{ }^{*}\left(\theta^{K}\right)=\theta^{* K} .
$$

Proof. Since formulas are defined recursively, we can argue inductively. The claim is true when $\theta$ is atomic, by ( $3^{2}$ ). If the claim is true when $\theta$ is $\phi$, then by (33)

$$
\begin{aligned}
*\left((\neg \phi)^{K}\right) & ={ }^{*}\left(K^{n} \backslash \phi^{K}\right) \\
& ={ }^{*} K^{n} \backslash{ }^{*}\left(\phi^{K}\right) \\
& ={ }^{*} K^{n} \backslash \phi^{*} K \\
& =(\neg \phi)^{*} K,
\end{aligned}
$$

so the claim is true when $\theta$ is $\neg \phi$. Similarly, if the claim is true when $\theta$ is $\phi$ or $\psi$, then by (34) the claim is true when $\theta$ is $(\phi \& \psi)$.

For the final case, let us first note that, if the claim is true when $\theta$ is considered as $m$-ary, and $m \leqslant n$, then the claim is still true when $\theta$ is considered as $n$-ary. Indeed, let $f$ be the inclusion of $m$ in $n$. Then

$$
\theta\left(x^{0}, \ldots, x^{n-1}\right)^{K}=\theta\left(x^{0}, \ldots, x^{m-1}\right)^{K} \times K^{n-m}=f_{*}^{K}\left(\theta\left(x^{0}, \ldots, x^{m-1}\right)^{K}\right),
$$

and likewise with ${ }^{*} K$ in place of $K$. Now use (36).
To finish then, we suppose the claim is true when $\theta$ is $\phi$, and we prove the claim when $\theta$ is $\exists x^{\ell} \phi$. We may assume $\phi$ and $\exists x^{\ell} \phi$ are both $n$-ary, where $\ell<n$. Then we can understand $\left(x^{0}, \ldots, x^{n-1}\right)$ as $(\boldsymbol{x}, y, \boldsymbol{z})$, where $\boldsymbol{x}$ is $\left(x^{0}, \ldots, x^{\ell-1}\right)$, and $y$ is $x^{\ell}$, and $\boldsymbol{x}$ is $\left(x^{\ell+1}, \ldots, x^{n-1}\right)$. Then

$$
\begin{aligned}
\left(\exists x^{\ell} \phi\right)^{K} & =\left\{(\boldsymbol{a}, b, \boldsymbol{c}) \in K^{n}: K \vDash\left(\exists x^{\ell} \phi\right)(\boldsymbol{a}, b, \boldsymbol{c})\right\} \\
& =\left\{(\boldsymbol{a}, b, \boldsymbol{c}) \in K^{n}: K \vDash \exists x^{\ell} \phi\left(\boldsymbol{a}, x^{\ell}, \boldsymbol{c}\right)\right\} \\
& =f_{*}^{K}\left(\left\{(\boldsymbol{a}, \boldsymbol{c}) \in K^{n-1}: K \vDash \exists x^{\ell} \phi\left(\boldsymbol{a}, x^{\ell}, \boldsymbol{c}\right)\right\}\right),
\end{aligned}
$$

where $f$ is the function from $n-1$ to $n$ given by

$$
f(k)= \begin{cases}k, & \text { if } k<\ell \\ k+1, & \text { if } \ell \leqslant k<n-1 .\end{cases}
$$

We have also that $K \vDash \exists x^{\ell} \phi\left(\boldsymbol{a}, x^{\ell}, \boldsymbol{c}\right)$ if and only if $K \vDash \phi(\boldsymbol{a}, b, \boldsymbol{c})$ for some $b$ in $K$. Then

$$
\left\{(\boldsymbol{a}, \boldsymbol{c}) \in K^{n-1}: K \vDash \exists x^{\ell} \phi\left(\boldsymbol{a}, x^{\ell}, \boldsymbol{c}\right)\right\}=f_{K}^{*}\left[\phi^{K}\right] .
$$

Combining these results, we have

$$
\left(\exists x^{\ell} \phi\right)^{K}=f_{*}^{K}\left(f_{K}^{*}\left[\phi^{K}\right]\right) .
$$

By (35) and (36) then, the claim holds when $\theta$ is $\exists x^{\ell} \phi$. Therefore it holds generally.
Theorem 45. Let $\sigma$ be a sentence with parameters from $K$. Then ${ }^{10}$

$$
K \vDash \sigma \Longleftrightarrow{ }^{*} K \vDash \sigma .
$$

Proof. When $n=0$, then equation (32) is simply ${ }^{*} S=S$.
$5 \cdot 3$. Mock higher-order logic. If we want our logic to be able to refer generally to subsets of a set, to functions on the set of functions on the set, and so forth, then we can proceed as follows. First, we recursively define types as certain strings:
(1) 0 is a type,
(2) if $n \in \omega \backslash\{0\}$, and $\left(t_{0}, \ldots, \tau_{n-1}\right)$ is a list of $n$ types, then the string

$$
n \tau_{0} \cdots \tau_{n-1}
$$

is a type.
Note that the type 0 is also a type of the form $n \tau_{0} \cdots \tau_{n-1}$, where $n=0$.
Theorem 46 (Unique Readability). Every type has the form $n \tau_{0} \cdots \tau_{n-1}$ for some unique $n$ in $\omega$ and some unique list $\left(\tau_{0}, \ldots, \tau_{n-1}\right)$ of types.

[^9]Given a set $K$, we can now define
(1) $K_{0}=K$,
(2) if $\tau$ is a type $n \tau_{0} \cdots \tau_{n-1}$, where $n>0$, then

$$
K_{\tau}=\mathscr{P}\left(K_{\tau(0)} \times \cdots \times K_{\tau(n-1)}\right)
$$

Here $\tau(j)$ is just $\tau_{j}$, when used as a subscript itself. The first condition is not a special case of the second: if $\tau$ is not 0 , then elements of $K_{\tau}$ are relations; but elements of $K_{0}$ are just elements of $K$. Letting $T$ be the set of types, we define

$$
\tilde{K}=\bigcup_{\tau \in T} K_{\tau}
$$

Letting $M$ be a maximal ideal of $\mathscr{P}(\omega)$ as before, we have a special case of (30):

$$
*(\tilde{K})=\tilde{K}^{\omega} / M
$$

This introduces a potential ambiguity, since $K \subseteq \tilde{K}$, but ${ }^{*} K$ is not literally a subset of ${ }^{*}(\tilde{K})$. Since a relation $S$ on $K$ is also a relation on $\tilde{K}$, its image ${ }^{*} S$ is either a relation on ${ }^{*} K$ or on ${ }^{*}(\tilde{K})$, but these two relations called ${ }^{*} S$ are not generally the same. This problem is taken care of by the following.

Theorem 47. *K embeds in ${ }^{*}(\tilde{K})$ under

$$
\left\{x \in K^{\omega}: x \equiv a\right\} \mapsto\left\{x \in \tilde{K}^{\omega}: x \equiv a\right\}
$$

Then an embedding $i$ of $\mathscr{P}\left({ }^{*} K^{n}\right)$ in $\mathscr{P}\left({ }^{*}(\tilde{K})^{n}\right)$ is induced, and the following diagram commutes.


In particular, if $S \subseteq K^{n}$, then

$$
i\left({ }^{*} S\right)={ }^{*} S
$$

where ${ }^{*} S$ is computed in ${ }^{*} K$ and ${ }^{*}(\tilde{K})$ respectively.
Another ambiguity arises when we consider that some elements of $\tilde{K}$ are also relations on $\tilde{K}$. Indeed, every element $S$ of $\tilde{K} \backslash K_{0}$ is a relation on $\tilde{K}$, so it determines both the element $(S: k \in \omega)+M$ or $S$ of ${ }^{*}(\tilde{K})$ and the relation ${ }^{*} S$ on ${ }^{*}(\tilde{K})$, but these are not literally the same. This is taken care of by the following.

Theorem 48. There is an embedding ८ from the subset $\bigcup_{\tau \in T}{ }^{*}\left(K_{\tau}\right)$ of ${ }^{*}(\tilde{K})$ into $\left(\widetilde{ }{ }^{*} K\right)$ such that
(1) $\iota(x)=x$ when $x \in{ }^{*} K$,
(2) $\iota(x) \in\left({ }^{*} K\right)_{\tau}$ when $x \in{ }^{*}\left(K_{\tau}\right)$, and
(3) if $S \in \tilde{K} \backslash K_{0}$, then

$$
\begin{equation*}
\iota(S)=\left\{\left(\iota\left(x^{0}\right), \ldots, \iota\left(x^{n-1}\right)\right): \boldsymbol{x} \in^{*} S\right\} . \tag{37}
\end{equation*}
$$

Proof. If $\tau$ is a type $n \tau_{0} \cdots \tau_{n-1}$, let

$$
E_{\tau}=\left\{(\boldsymbol{x}, y) \in \tilde{K}^{n+1}: \boldsymbol{x} \in y \& y \in K_{\tau}\right\} .
$$

Then the sentence

$$
\forall \boldsymbol{x} \forall y\left(E_{\tau} \boldsymbol{x} y \Rightarrow K_{\tau(0)} x^{0} \& \cdots \& K_{\tau(n-1)} x^{n-1} \& K_{\tau} y\right)
$$

is true in $\tilde{K}$. By Theorem 45 , it is true in ${ }^{*}(\tilde{K})$. By Theorem 44 ,

$$
\left(K_{\tau} x\right)^{*}(\tilde{K})={ }^{*}\left(K_{\tau}\right) .
$$

Hence, if $S \in{ }^{*}\left(K_{\tau}\right)$, then

$$
\left(E_{\tau} x S\right)^{*}(\tilde{K}) \subseteq{ }^{*}\left(K_{\tau(0)}\right) \times \cdots \times^{*}\left(K_{\tau(n-1)}\right),
$$

and the function that converts such $S$ to $\left(E_{\tau} x S\right)^{*(\tilde{K})}$ is an embedding. We can now define $\iota$ recursively:
(1) $\iota(x)=x$ if $x \in{ }^{*}\left(K_{0}\right)$,
(2) if $\tau=n \tau_{0} \cdots \tau_{n-1}$, and $R \in{ }^{*}\left(K_{\tau}\right)$, then

$$
\begin{equation*}
\iota(R)=\left\{\left(\iota\left(a^{0}\right), \ldots, \iota\left(a^{n-1}\right)\right):{ }^{*}(\tilde{K}) \vDash E_{\tau} \boldsymbol{a} R\right\} . \tag{38}
\end{equation*}
$$

By induction, $\iota$ maps ${ }^{*}\left(K_{\tau}\right)$ into $\left({ }^{*} K\right)_{\tau}$. Moreover, if again $\tau=n \tau_{0} \cdots \tau_{n-1}$, and $S \in K_{\tau}$, then the sentence

$$
\forall \boldsymbol{x}\left(S \boldsymbol{x} \Longleftrightarrow E_{\tau} \boldsymbol{x} S\right)
$$

is true in $\tilde{K}$, so it is true in ${ }^{*}(\tilde{K})$, which means in particular

$$
{ }^{*} S=S^{*}(\tilde{K})=(S \boldsymbol{x})^{*(\tilde{K})}=\left(E_{\tau} \boldsymbol{x} S\right)^{*(\tilde{K})} .
$$

This and (38) establish (37).
As we shall see in $\S 6, \iota$ is not generally surjective. Also, even though every element of $\tilde{K}$ is an element of some $K_{\tau}$, not every element of ${ }^{*}(\tilde{K})$ is an element of some ${ }^{*}\left(K_{\tau}\right)$.

## 6. Analysis

Let ${ }^{*} \mathbb{R}$ be the ultrapower $\mathbb{R}^{\omega} / M$ of $\mathbb{R}$, where $M$ is a non-principal maximal ideal of $\mathscr{P}(\omega)$. Then $\mathbb{R}$ embeds properly in ${ }^{*} \mathbb{R}$, by Theorem 41. Everything we do now will be based on Theorem 45 in case $K=\mathbb{R}$ or $K=\tilde{\mathbb{R}}$.

If $S$ is a relation on $\mathbb{R}$, then ${ }^{*} S$ is a standard relation and is the extension of $S$. Also, elements of $\mathbb{R}^{n}$ are standard. Note then that a standard relation might have nonstandard elements.

Suppose $S \subseteq \mathbb{R}^{n}$, and $f$ is a function from $S$ to $\mathbb{R}$. Let

$$
T=\{(\boldsymbol{x}, f(\boldsymbol{x})): \boldsymbol{x} \in S\}
$$

this is the graph of $f$. Then the sentences

$$
\begin{gathered}
\forall \boldsymbol{x} \exists y(S \boldsymbol{x} \Rightarrow T \boldsymbol{x} y), \\
\mathbb{R} \vDash \forall \boldsymbol{x} \forall y \forall z(S \boldsymbol{x} \& T \boldsymbol{x} y \& T \boldsymbol{x} z \Rightarrow y=z)),
\end{gathered}
$$

are true in $\mathbb{R}$; by Theorem 44, they are true in ${ }^{*} \mathbb{R}$. Therefore ${ }^{*} T$ is the graph of a function from ${ }^{*} S$ to ${ }^{*} \mathbb{R}$. We may denote this function by

$$
{ }^{*} f
$$

We have then

$$
\begin{equation*}
{ }^{*} f \upharpoonright S=f \tag{39}
\end{equation*}
$$

In a formula, in place of $T \boldsymbol{x} y$, we may write

$$
f(\boldsymbol{x})=y
$$

Theorem 49. ${ }^{*} \mathbb{R}$ is a non-Archimedean ordered field with respect to ${ }^{*}<,^{*}+,^{*}-$, and ${ }^{*}$, and $\mathbb{R}$ is an ordered subfield of $* \mathbb{R}$.

Proof. There is a first-order sentence $\sigma$ saying that $\mathbb{R}$ is an ordered field; but then ${ }^{*} \mathbb{R} \vDash \sigma$. By (32) and (39), $\mathbb{R}$ is an ordered subfield of ${ }^{*} \mathbb{R}$. Since $\mathbb{R}$ is a proper subset of ${ }^{*} \mathbb{R}$, the latter must be non-Archimedean.

Corollary. Being Archimedean is not a first-order property of fields.
In the notation of the proof of Theorem 48, we have the sentence

$$
\begin{aligned}
& \forall x\left(\exists y E_{10} y x \& \exists z \forall y\left(E_{10} y x \Rightarrow y \leqslant z\right) \Rightarrow\right. \\
& \left.\exists w\left(\forall y\left(E_{10} y x \Rightarrow y \leqslant w\right) \& \forall z\left(\forall y\left(E_{10} y x \Rightarrow y \leqslant z\right) \Rightarrow w \leqslant z\right)\right)\right)
\end{aligned}
$$

which says in $\tilde{\mathbb{R}}$ that every subset $x$ of $\mathbb{R}$ with an element $y$ and an upper bound $z$ has a least upper bound $w$. This is true, so the same sentence is true in ${ }^{*}(\tilde{\mathbb{R}})$. But more precisely, in $\tilde{\mathbb{R}}$, the sentence is not about subsets of $\mathbb{R}$, but about elements of $\mathbb{R} 10$, which is $\mathscr{P}(\mathbb{R})$. In $*(\tilde{\mathbb{R}})$, the sentence says that every element of $*\left(\mathbb{R}_{10}\right)$ with an upper bound has a least upper bound. We know that $\iota\left({ }^{*}\left(\mathbb{R}_{10}\right)\right) \subseteq\left({ }^{*} \mathbb{R}\right)_{10}$, which is $\mathscr{P}\left({ }^{*} \mathbb{R}\right)$. Now we can conclude that the embedding is proper, not surjective.

Theorem 5o. $\mathbb{N}$ is a proper initial segment of ${ }^{*} \mathbb{N}$. In particular, $\mathbb{N}$ consists of the finite elements of ${ }^{*} \mathbb{N}$.

Proof. For each $n$ in $\mathbb{N}$, the sentence

$$
\forall x(\mathbb{N} x \Rightarrow x=0 \vee x=1 \vee \cdots \vee x=n \vee x>n)
$$

is true in $\mathbb{R}$, hence in ${ }^{*} \mathbb{R}$, so that $\{0,1, \ldots, n\}$ is an initial segment of ${ }^{*} \mathbb{N}$. The sentence

$$
\forall x \exists y(\mathbb{N} y \& x<y)
$$

is true in $\mathbb{R}$ and hence in ${ }^{*} \mathbb{R}$. In particular, let $a$ be a positive infinite element of ${ }^{*} \mathbb{R}$. Then there is $n$ in $* \mathbb{N}$ such that $a<n$. Such $n$ must be infinite and so are not in $\mathbb{N}$.

Corollary. The Peano Axioms are not first order.
Theorem 51. A standard relation has nonstandard elements if and only if it is infinite.
Proof. Suppose $S=\left\{\boldsymbol{a}_{0}, \ldots, \boldsymbol{a}_{n-1}\right\} \subseteq \mathbb{R}^{m}$. Then the sentence

$$
\forall \boldsymbol{x}\left(S \boldsymbol{x} \Leftrightarrow \boldsymbol{x}=\boldsymbol{a}_{0} \vee \cdots \vee \boldsymbol{a}_{n-1}\right)
$$

is true in $\mathbb{R}$ and ${ }^{*} \mathbb{R}$, so ${ }^{*} S=S$.
Now suppose $f$ is an injective function from $\mathbb{N}$ into $\mathbb{R}$. Then ${ }^{*} f$ is an injective function on ${ }^{*}$ N. If ${ }^{*} f(n)$ is an element $a$ of $\mathbb{R}$ for some $n$ in ${ }^{*} \mathbb{N}$, then the sentence

$$
\exists x(\mathbb{N} x \& f(x)=a)
$$

is true in ${ }^{*} \mathbb{R}$ and $\mathbb{R}$, so $n \in \mathbb{N}$. Thus, if $n \in{ }^{*} \mathbb{N} \backslash \mathbb{N}$, then ${ }^{*} f(n) \in{ }^{*} \mathbb{R} \backslash \mathbb{R}$.

### 6.1. Sequences.

Theorem 52. Let a be a sequence $\left(a_{n}: n \in \mathbb{N}\right)$ in $\mathbb{R}$. Then ${ }^{*} a$ is a sequence $\left(a_{n}: n \in{ }^{*} \mathbb{N}\right)$ for some $a_{n}$ in ${ }^{*} \mathbb{R}$ (for $n$ in ${ }^{*} \mathbb{N} \backslash \mathbb{N}$ ). Also a converges if and only if "a converges. If ${ }^{*} a$ converges to $b$, then $b \in \mathbb{R}$.

The following theorem can be understood as an alternative definition of convergence, if one does not want to bother with the traditional definition.

Theorem 53. A standard sequence ( $a_{n}: n \in{ }^{*} \mathbb{N}$ ) converges to $L$ if and only if, for all infinite $n$,

$$
a_{n} \simeq L .
$$

Proof. Let $a$ be the sequence, and suppose it converges to $L$. For every positive standard $\varepsilon$, there is a standard $M$ such that the sentence

$$
\forall x\left(\mathbb{N} x \& x \geqslant M \Rightarrow\left|a_{n}-L\right|<\varepsilon\right)
$$

is true in $\mathbb{R}$ and ${ }^{*} \mathbb{R}$. In particular, for every infinite $n$, we have $\left|a_{n}-L\right|<\varepsilon$ for every standard positive $\varepsilon$; but this just means $a_{n} \simeq L$.

Suppose $a$ does not converge to $L$. Then there is some positive standard $\varepsilon$ such that the sentence

$$
\forall y \exists x\left(\mathbb{N} x \& x \geqslant y \&\left|a_{n}-L\right| \geqslant \varepsilon\right)
$$

is true in $\mathbb{R}$ and ${ }^{*} \mathbb{R}$. In particular, if $M$ is infinite, then there is some $n$ in ${ }^{*} \mathbb{N}$ that is greater than $M$ (and therefore infinite) such that $\left|a_{n}-L\right| \geqslant \varepsilon$.

For example, $\lim _{n \rightarrow \infty} 1 / n=0$, simply because $1 / n$ is infinitesimal when $n$ is infinite.

Theorem 54. Let $a$ and $b$ be standard convergent sequences, and $r \in \mathbb{R}$. Then

$$
\begin{gather*}
\lim (a+b)=\lim (a)+\lim (b),  \tag{40}\\
\lim (r a)=r \lim (a),  \tag{41}\\
\lim (a b)=\lim (a) \lim (b) . \tag{42}
\end{gather*}
$$

Proof. Let $R$ be the ring of finite members of ${ }^{*} \mathbb{R}$, and let $I$ be its ideal of infinitesimals. Suppose $a_{n} \simeq L$ and $b_{n} \simeq M$. Then $a_{n}-L$ and $b_{n}-M$ are in $I$, hence so are $\left(a_{n}+b_{n}\right)-$ $(L+M)$ and $r a_{n}-r L$. This shows (40) and (41). For (42), note

$$
\left|a_{n} b_{n}-L M\right|=\left|a_{n} b_{n}-a_{n} M+a_{n} M-L M\right| \leqslant\left|a_{n}\right|\left|b_{n}-M\right|+\left|a_{n}-L\right||M| .
$$

But the last is in $I$ since $\left|a_{n}\right|$ and $|M|$ are in $R$ (why?).
Theorem 55. A standard sequence is bounded if and only if every term is finite.
Proof. That $a$ is bounded means that, for some $M$, the sentence

$$
\forall x\left(x \in \mathbb{N} \Rightarrow\left|a_{x}\right|<M\right)
$$

is true in $\mathbb{R}$; then it is true in ${ }^{*} \mathbb{R}$, so every entry in ${ }^{*} a$ is bounded by $M$, hence finite.
Suppose $a$ is unbounded. Then the sentence

$$
\forall x \exists y\left(y \in \mathbb{N} \&\left|a_{y}\right|>x\right.
$$

is true in $\mathbb{R}$, hence in ${ }^{*} \mathbb{R}$. Let $L$ be positive and infinite; then there is $n$ in ${ }^{*} \mathbb{N}$ such that $\left|a_{n}\right|>L$.

Compare the following with Theorem 30.
Theorem 56. A standard sequence $\left(a_{n}\right)$ converges if and only if, for all infinite $m$ and $n$,

$$
a_{m} \simeq a_{n}
$$

Proof. If $\left(a_{n}\right)$ converges to $L$, then $a \simeq L$ for all infinite $n$, and therefore $a_{m} \simeq a_{n}$ for all infinite $m$ and $n$, since $\simeq$ is an equivalence relation.

Suppose conversely $a_{m} \simeq a_{n}$ for all infinite $m$ and $n$. If each $a_{n}$ is finite, and $n$ is in particular infinite, then $\left(a_{n}\right)$ converges to the standard part of $a_{n}$. Suppose some $a_{n}$ is infinite. Then by Theorem 55 , the sequence ( $a_{n}: n \in \mathbb{N}$ ) is unbounded. Hence the sentence

$$
\forall x \exists y\left(x \in \mathbb{N} \Rightarrow y \in N \& x \leqslant y \&\left|a_{x}\right|+1 \leqslant\left|a_{y}\right|\right)
$$

is true in $\mathbb{R}$ and ${ }^{*} \mathbb{R}$, so $a_{m}$ and $a_{n}$ fail to be infinitely close for some infinite $m$ and $n$.
Traditionally, $L$ is a limit point of $\left(a_{n}\right)$ if for all positive $\varepsilon$ and for all $m$ in $\mathbb{N}$, there is $n$ such that $m<n$ and $\left|L-a_{n}\right|<\varepsilon$. The following can be used as an alternative definition.

Theorem 57. A finite number $L$ is a limit point of the standard sequence $\left(a_{n}\right)$ if and only if, for some infinite $n$,

$$
a_{n} \simeq L
$$

Proof. If $L$ is a limit point of $\left(a_{n}\right)$, then the sentence

$$
\forall x \forall y \exists z\left(x>0 \& y \in \mathbb{N} \Rightarrow z \in N \& y<z \&\left|L-a_{z}\right|<x\right)
$$

is true in $\mathbb{R}$ and $* \mathbb{R}$, so for an infinitesimal $\varepsilon$ there is an infinite $n$ such that $\left|L-a_{n}\right|<\varepsilon$ and hence $a_{n} \simeq L$.

Suppose $L$ is not a limit point of $\left(a_{n}\right)$. Then there is some positive $\varepsilon$ and some $n$ in $\mathbb{N}$ such that the sentence

$$
\forall x\left(x \in \mathbb{N} \Rightarrow\left|L-a_{x}\right| \geqslant \varepsilon\right)
$$

is true in $\mathbb{R}$ and ${ }^{*} \mathbb{R}$. This means $\left|L-a_{n}\right| \geqslant \varepsilon$ whenever $n$ is infinite.
The non-standard proof of the following should be compared with the traditional divide-and-conquer proof.
Theorem 58 (Bolzano-Weierstraß). Every bounded standard sequence has a limit point.
Proof. Indeed, by Theorem 55 , if $\left(a_{n}\right)$ is bounded, then each $a_{n}$ has a standard part when $n$ is infinite, and that standard part is a limit point of the sequence by Theorem 57 .
6.2. Continuity. Suppose now $f$ is a standard function defined on an interval $[a, b]$. If $c \in[a, b]$, we say, classically, that

$$
\lim _{x \rightarrow c} f(x)=\lim _{c}(f)=L
$$

if for all positive $\varepsilon$ there is a positive $\delta$ such that, for all $x$ in $[a, b]$,

$$
0<|x-c|<\delta \Longrightarrow|L-f(x)|<\varepsilon .
$$

Theorem 59. If $f$ is a standard function defined on an interval $I$ that contains $c$, then $\lim _{c} f=L$ if and only if, for all $x$ in $I$ that are distinct from $c$,

$$
x \simeq c \Longrightarrow f(x) \simeq L .
$$

Theorem 60. If $\lim _{c} f$ and $\lim _{c} g$ exist, then

$$
\begin{aligned}
\lim _{c}(f+g) & =\lim _{c}(f)+\lim _{c}(g), \\
\lim _{c}(f g) & =\lim _{c}(f) \lim _{c}(g) ;
\end{aligned}
$$

if also $\lim _{c}(f) \neq 0$, then

$$
\begin{equation*}
\lim _{c}\left(\frac{1}{f}\right)=\frac{1}{\lim _{c}(f)} . \tag{43}
\end{equation*}
$$

Proof. For (43), if $f(x) \simeq L$, and $L \neq 0$, then $|f(x)|>|L| / 2$, so that

$$
\left|\frac{1}{f(x)}-\frac{1}{L}\right|=\frac{|L-f(x)|}{|f(x) L|}<\frac{2}{L^{2}}|L-f(x)| \simeq 0 .
$$

The function $f$ is continuous at $c$, if $\lim _{c}(f)=f(c)$; continuous on $[a, b]$, if continuous at every point of $[a, b]$.
Theorem 61 (Intermediate Value). If $f$ is continuous on $[a, b]$, and $d$ lies between $f(a)$ and $f(b)$, then for some $c$ in $(a, b)$,

$$
f(c)=d .
$$

Proof. Suppose $f(a)<d<f(b)$. In $\mathbb{R}$, for all $n$ in $\mathbb{N}$, there is some least $j$ in $\mathbb{N}$ such that

$$
\begin{equation*}
f\left(a+\frac{j}{n}(b-a)\right)<d \leqslant f\left(a+\frac{j+1}{n}(b-a)\right) . \tag{44}
\end{equation*}
$$

Then the same is true in ${ }^{*} \mathbb{R}$, with $* \mathbb{N}$ replacing $\mathbb{N}$. In particular, we have (44) for some $j$ in $* \mathbb{N}$, where $n$ is in ${ }^{*} \mathbb{N} \backslash \mathbb{N}$. Let $c$ be the standard part of $a+(j / n)(b-a)$. Then

$$
f\left(a+\frac{j}{n}(b-a)\right) \simeq f(c) \simeq f\left(a+\frac{j+1}{n}(b-a)\right)
$$

Therefore $f(c)=d$.
Theorem 62 (Extreme Value). If $f$ is continuous on $[a, b]$, then it attains a maximum and minimum value on the interval.

Proof. For all positive natural numbers $n$, for some natural number $j$ such that $j \leqslant n$, the value of

$$
f\left(a+\frac{j}{n}(b-a)\right)
$$

is maximized. In particular, this is so when $n$ is infinite. If $i \leqslant n$, we now have

$$
f\left(a+\frac{i}{n}(b-a)\right) \leqslant f\left(a+\frac{j}{n}(b-a)\right)
$$

Let $d$ be the standard part of $a+(j / n)(b-a)$. For every $c$ in $[a, b]$, there is a natural number $i$ such that

$$
a+\frac{i}{n}(b-a) \leqslant c<a+\frac{i+1}{n}(b-a) .
$$

Then these three numbers are infinitely close, so

$$
f(c) \simeq f\left(a+\frac{i}{n}(b-a)\right)
$$

Therefore $f(c) \leqslant f(d)$.
6.3. Derivatives. If

$$
\lim _{x \rightarrow c} \frac{f(x)-f(c)}{x-c}=d
$$

then we write

$$
f^{\prime}(c)=d
$$

saying $f$ is differentiable at $c$, with derivative $d$ at $c$. So $f^{\prime}(c)=d$ if and only if, whenever $c \simeq c$ but $x \neq c$, we have

$$
\frac{f(x)-f(c)}{x-c} \simeq d
$$

Theorem 63. If $f$ is differentiable at $c$, then it is continuous at $c$.
Proof. If $f$ is differentiable at $c$ and $x \simeq c$, then

$$
f(x)-f(c) \simeq(x-c) f^{\prime}(c) \simeq 0
$$

so $f$ is continuous at $c$.
Theorem 64. If $f$ and $g$ are differentiable at $c$, then so are $f+g$ and $f g$, and

$$
(f+g)^{\prime}(c)=f^{\prime}(c)+g^{\prime}(c), \quad(f g)^{\prime}(c)=f^{\prime}(c) g(c)+f(c) g^{\prime}(c)
$$

Proof. If $x \simeq c$, then

$$
\frac{(f g)(x)-(f g)(c)}{x-c}=\frac{f(x)-f(c)}{x-c} g(c)+f(x) \frac{g(x)-g(c)}{x-c} \simeq f^{\prime}(c) g(c)+f(c) g^{\prime}(c) .
$$

The standard function $f$ has a local maximum at $c$ if, for some positive $\delta$, the function $f$ is defined on $(c-\delta, c+\delta)$ and on this interval is maximized at $c$.

Theorem 65. A standard function $f$ has a local maximum at $c$ if and only if $f$ is defined on $\{x: x \simeq c\}$ and on this interval is maximized at $c$.

Proof. Necessity of the condition is immediate. To prove sufficiency, suppose $f$ does not have a local maximum at $c$. Then for every positive $\delta$, and in particular for $\delta$ that are infinitely close to $c$, either $f$ is not defined on $(c-\delta, c+\delta)$, or else $f$ is not maximized there at $c$. But that interval is a subset of $\{x: x \simeq c\}$.

Theorem 66. If $f$ has a local maximum and is differentiable at $c$, then $f^{\prime}(c)=0$.
Proof. We assume that, if $x \simeq c$, but $x \neq c$, then $f(x) \leqslant c$, so

$$
\frac{f(x)-f(c)}{x-c} \begin{cases}\geqslant 0, & \text { if } x<c, \\ \leqslant 0, & \text { if } x>c .\end{cases}
$$

Since $(f(x)-f(c)) /(x-c) \simeq f^{\prime}(c)$, we can conclude that $f^{\prime}(c)=0$.
Theorem 67 (Rolle). If $f$ is continuous on $[a, b]$ and differentiable on $(a, b)$, and $f(a)=$ $f(b)$, then, for some $c$ in $(a, b)$,

$$
f^{\prime}(c)=0 .
$$

Proof. Theorems 62 and 66.
6.4. Integrals. Classically, the integral of a bounded function $f$ on an interval $[a, b]$ can be defined as follows. Suppose $I$ is a number such that, for all positive $\varepsilon$, there is a positive $\delta$ such that, for all positive integers $n$, for all lists $\left(a_{0}, \ldots, a_{n}\right)$ and $\left(\xi_{1}, \ldots, \xi_{n}\right)$ of numbers such that

$$
\begin{equation*}
a=a_{0} \leqslant \xi_{1} \leqslant a_{1} \leqslant \cdots \leqslant a_{n-1} \leqslant \xi_{n} \leqslant a_{n}=b \tag{45}
\end{equation*}
$$

and also

$$
\min \left(a_{1}-a_{0}, \ldots, a_{n}-a_{n-1}\right) \leqslant \delta
$$

we have

$$
\left|I-\sum_{i=1}^{n} f\left(\xi_{i}\right)\left(a_{i}-a_{i-1}\right)\right|<\varepsilon .
$$

Then $f$ is integrable on $[a, b]$, and

$$
\int_{a}^{b} f=I .
$$

This is not a first-order statement in $\mathbb{R}$, so we move to $\tilde{\mathbb{R}}$. Let $A_{[a, b]}$ be the set of finite sequences ( $a_{0}, \xi_{1}, a_{1}, \ldots, x_{n}, a_{n}$ ) with entries from $\mathbb{R}$, where $n \in \mathbb{N}$ and (45) holds. Such
sequences can be understood as binary relations on $\mathbb{R}$, so that $A_{[a, b]} \in \mathbb{R}_{200}$. If $f$ is a bounded function on $[a, b]$, let $S_{f, a, b}$ be the function

$$
\left(a_{0}, \xi_{1}, a_{1}, \ldots, \xi_{n}, a_{n}\right) \mapsto \sum_{i=1}^{n} f\left(\xi_{i}\right)\left(a_{i}-a_{i-1}\right)
$$

on $A$. So $S_{f, a, b} \in \mathbb{R}_{22000}$. An element of ${ }^{*} A_{[a, b]}$ also takes the form $\left(a_{0}, \xi_{1}, a_{1}, \ldots, \xi_{n}, a_{n}\right)$, where again (45) holds; but now $n \in * \mathbb{N}$. Such an element can be called fine if $a_{i-1} \simeq a_{i}$ for each $i$ in $\{1, \ldots, n\}$. It must be noted that fine elements of ${ }^{*} A$ do exist: for example,

$$
\left(a, a+\frac{1}{n}(b-a), a+\frac{1}{n}(b-a), \ldots, a+\frac{n-1}{n}(b-a), a+\frac{n-1}{n}(b-a)\right),
$$

where $n$ is infinite.
Theorem 68. Given a bounded function $f$ on $[a, b]$, Then $f$ is integrable on $[a, b]$ if and only if, for any two fine elements $a$ and $a^{\prime}$ of ${ }^{*} A_{[a, b]}$,

$$
{ }^{*} S_{f, a, b}(a) \simeq{ }^{*} S_{f, a, b}\left(a^{\prime}\right)
$$

In this case, $\int_{a}^{b} f$ is the standard part of either of these sums.
Theorem 69 (Fundamental, of Calculus). Suppose $f$ is continuous on $[a, b]$. If $a \leqslant c \leqslant b$, then $f$ is integrable on $[a, c]$. The function

$$
x \mapsto \int_{a}^{x} f
$$

is differentiable on $[a, b]$, and its derivative is $f$.

## Appendix A. The Greek alphabet

| capital | minuscule | transliteration | name |
| :---: | :---: | :---: | :---: |
| $A$ | $a$ | a | alpha |
| $B$ | $\beta$ | b | beta |
| $\Gamma$ | $\gamma$ | g | gamma |
| $\Delta$ | $\delta$ | d | delta |
| $E$ | $\epsilon$ | e | epsilon |
| $Z$ | $\zeta$ | z | zeta |
| $H$ | $\eta$ | è | eta |
| $\Theta$ | $\theta$ | th | theta |
| $I$ | $\iota$ | i | iota |
| $K$ | $\kappa$ | k | kappa |
| $\Lambda$ | $\lambda$ | l | lambda |
| $M$ | $\mu$ | m | mu |
| $N$ | $\nu$ | n | nu |
| $\Xi$ | $\xi$ | x | xi |
| $O$ | $o$ | o | omicron |
| $\Pi$ | $\pi$ | p | pi |
| $P$ | $\rho$ | r | rho |
| $\Sigma$ | $\sigma, s$ | s | sigma |
| $T$ | $\tau$ | t | tau |
| $Y$ | $v$ | $\mathrm{y}, \mathrm{u}$ | upsilon |
| $\Phi$ | $\phi$ | ph | phi |
| $X$ | $\chi$ | ch | chi |
| $\Psi$ | $\psi$ | ps | psi |
| $\Omega$ | $\omega$ | o | omega |
|  |  |  |  |
|  |  |  |  |

The following remarks pertain to ancient Greek. The vowels are $a, \epsilon, \eta, \iota, o, v, \omega$, where $\eta$ is a long $\epsilon$, and $\omega$ is a long $o$; the other vowels ( $\alpha, \iota, v$ ) can be long or short. Some vowels may be given tonal accents ( $\dot{\alpha}, \hat{a}, \grave{\alpha})$. An initial vowel takes either a rough-breathing mark (as in $\dot{\alpha}$ ) or a smooth-breathing mark ( $\dot{\alpha}$ ): the former mark is transliterated by a preceding h , and the latter can be ignored, as in $\dot{v} \pi \epsilon \rho \beta \circ \lambda \dot{\eta}$ hyperbolê hyperbola, ò $\rho \theta o \gamma \dot{\omega} \nu \iota o \nu$ orthogônion rectangle. Likewise, $\dot{\rho}$ is transliterated as rh, as in $\dot{\rho}$ ó $\mu$ ßos rhombos rhombus. A long vowel may have an iota subscript ( $\alpha, \eta, \omega$ ), especially in case-endings of nouns. Of the two forms of minuscule sigma, the $s$ appears at the ends of words; elsewhere, $\sigma$ appears, as in $\beta$ ávıs basis base.

## Appendix B. The German script

Writing in 1993, Wilfrid Hodges [8, Ch. 1, p. 21] observes
Until about a dozen years ago, most model theorists named structures in horrible Fraktur lettering. Recent writers sometimes adopt a notation according to which all structures are named $M, M^{\prime}, M^{*}, \bar{M}, M_{0}, M_{i}$ or occasionally $N$.
For Hodges, structures are $A, B, C$, and so forth; he refers to their universes as domains and denotes these by $\operatorname{dom}(A)$ and so forth. This practice is convenient if one is using a typewriter (as in the preparation of another of Hodges's books [9], from 1985). In 2002, David Marker [11] uses 'calligraphic' letters for structures, so that $M$ is the universe of $\mathcal{M}$. I still prefer the Fraktur letters:

| $\mathfrak{A}$ | $\mathfrak{B}$ | $\mathfrak{C}$ | $\mathfrak{D}$ | $\mathfrak{E}$ | $\mathfrak{F}$ | $\mathfrak{G}$ | $\mathfrak{H}$ | $\mathfrak{I}$ | $\mathfrak{a}$ | $\mathfrak{b}$ | $\mathfrak{c}$ | $\mathfrak{d}$ | $\mathfrak{e}$ | $\mathfrak{f}$ | $\mathfrak{g}$ | $\mathfrak{h}$ | $\mathfrak{i}$ |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| $\mathfrak{J}$ | $\mathfrak{K}$ | $\mathfrak{L}$ | $\mathfrak{M}$ | $\mathfrak{N}$ | $\mathfrak{O}$ | $\mathfrak{P}$ | $\mathfrak{Q}$ | $\mathfrak{R}$ | $\mathfrak{j}$ | $\mathfrak{k}$ | $\mathfrak{l}$ | $\mathfrak{m}$ | $\mathfrak{n}$ | $\mathfrak{o}$ | $\mathfrak{p}$ | $\mathfrak{q}$ | $\mathfrak{r}$ |
| $\mathfrak{S}$ | $\mathfrak{T}$ | $\mathfrak{U}$ | $\mathfrak{V}$ | $\mathfrak{W}$ | $\mathfrak{X}$ | $\mathfrak{Y}$ | $\mathfrak{Z}$ |  | $\mathfrak{s}$ | $\mathfrak{t}$ | $\mathfrak{u}$ | $\mathfrak{v}$ | $\mathfrak{w}$ | $\mathfrak{x}$ | $\mathfrak{y}$ | $\mathfrak{z}$ |  |

A way to write these by hand is seen in a textbook of German from 1931 [7]:
A

Cc Dd
$\stackrel{E}{E}$






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[^0]:    Date: January 8, 2010.

[^1]:    ${ }^{1} \underset{\epsilon}{\boldsymbol{\epsilon}} \boldsymbol{i} \boldsymbol{i} \tau \rho \iota \tau o s$, one and a third times as much. See Appendix A for the Greek letters.
    

[^2]:    ${ }^{3}$ Or 'right-angled'; but I want to avoid confusion with right cones in general.

[^3]:    4Some people might write these structures as $(\mathbb{Z} / n, \overline{0}, \bar{k} \mapsto \overline{k+1})$, to make sure that the reader distinguishes a natural number $k$ from its congruence class modulo $n$.

[^4]:    ${ }^{5}$ Here $\mathfrak{A}$ is the Fraktur version of $A$. The idea is that $A$ is just a set, but $\mathfrak{A}$ is that set together with some other things - here a distinguished element and singulary operation. Then $A$ is the universe of $\mathfrak{A}$. I shall not later distinguish notationally between structures and their universes. If I did want to make a distinction, I would use the letters depicted in Appendix B.

[^5]:    ${ }^{6}$ This example, and the difficulty it illustrates, are discussed on [Timothy] Gowers's Weblog at http: //gowers.wordpress.com/2009/06/08/why-arent-all-functions-well-defined/.

[^6]:    ${ }^{7}$ This is not standard terminology. Usually what I am calling a strong valuation is just called a valuation, and what I shall call a valuation is called an 'absolute value,' at least if its range is included in $\mathbb{R}$. I object to this usage on linguistic grounds: An absolute value should be the value of a function at a particular argument, not the whole function itself. (Similarly a product is the value or result of a multiplication; the two terms are not properly interchangeable. In linear algebra then, if one is being linguistically strict, an 'inner product' should be called an 'inner multiplication,' if the operation itself is meant.)

[^7]:    ${ }^{8}$ In the language of category theory, the pair ( $m \mapsto \Omega^{m}, f \mapsto f^{*}$ ) is a contravariant functor from the category ( $\omega$, , functions $\}$ ) to the category (\{sets\}, \{functions\}).

[^8]:    9I don't know of a common term for formulas $\exists x \phi$; instantiation seems to work, though, since the formula will be interpreted as saying that $\phi$ is true for some instance of $x$.

[^9]:    ${ }^{10}$ In model-theoretic terms, the full structure on $K$ is an elementary substructure of the structure induced on ${ }^{*} K$ by $X \mapsto{ }^{*} X$.

