PRIMES
What is Riemann’s Hypothesis?

(August 2013 Draft)

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Foreword

The Riemann Hypothesis is one of the great unsolved problems of mathematics and the reward of $1,000,000 of Clay Mathematics Institute prize money awaits the person who solves it. But—with or without money—its resolution is crucial for our understanding of the nature of numbers.

There are at least four full-length books recently published, written for a general audience, that have the Riemann Hypothesis as their main topic. A reader of these books will get a fairly rich picture of the personalities engaged in the pursuit, and of related mathematical and historical issues. [1]

This is not the mission of the book that you now hold in your hands. We aim—instead—to explain, in as direct a manner as possible and with the least mathematical background required, what this problem is all about and why it is so important. For even before anyone proves this hypothesis to be true (or false!), just getting familiar with it and with some of the the ideas behind it, is exciting. Moreover, this hypothesis is of crucial importance in a wide range of mathematical fields; for example, it is a confidence-booster for computational mathematics: even if the Riemann Hypothesis is never proved, its truth gives us an excellent sense of how long certain computer programs will take to run, which, in some cases, gives us the assurance we need to initiate a computation that might take weeks or even months to complete.

Figure 1: Peter Sarnak

Here is how the Princeton mathematician Peter Sarnak described the broad impact the Riemann Hypothesis has had:

“The Riemann hypothesis is the central problem and it implies many, many things. One thing that makes it rather unusual in mathematics
today is that there must be over five hundred papers—somebody should go and count—which start ‘Assume the Riemann hypothesis,’ and the conclusion is fantastic. And those [conclusions] would then become theorems ... With this one solution you would have proven five hundred theorems or more at once.” [2]

Our book is profusely illustrated, containing over 125 figures, diagrams, and pictures that accompany the text [3]. There are comparatively fewer mathematical equations in Part [1].

So, what is the Riemann Hypothesis? Below is a first description of what it is about. The task of our book is to develop the following boxed paragraph into a fuller explanation and to convince you of the importance and beauty of the mathematics it represents. We will be offering, throughout our book, a number of different—but equivalent—ways of precisely formulating this hypothesis (we display these in boxes). When we say that two mathematical statements are “equivalent” we mean that, given the present state of mathematical knowledge, we can prove that if either one of those statements is true, then the other is true. The endnotes will guide the reader to the mathematical literature that establishes these equivalences.

**What sort of Hypothesis is the Riemann Hypothesis?**

Consider the seemingly innocuous series of questions:

- How many prime numbers (2, 3, 5, 7, 11, 13, 17, ...) are there less than 100?
- How many less than 10,000?
- How many less than 1,000,000?

More generally, how many primes are there less than any given number $X$?

Riemann proposed, a century and a half ago, a strikingly simple-to-describe “very good approximation” to the number of primes less than a given number $X$. We now see that if we could prove this Hypothesis of Riemann we would have the key to a wealth of powerful mathematics. Mathematicians are eager to find that key.
A famous mathematician, Raoul Bott, once said—giving advice to some young mathematicians—that whenever one reads a mathematics book or article, or goes to a math lecture, one should aim to come home with something very specific (it can be small, but should be *specific*) that has application to a wider class of mathematical problem than was the focus of the text or lecture. If we were to suggest some possible *specific* items to come home with, after read our book, three key phrases—prime numbers, square-root accurate, and spectrum—would head the list. As for words of encouragement to think hard about the first of these, i.e., prime numbers, we can do no better than to quote a paragraph of Don Zagier’s classic 12-page exposition, *The First 50 Million Prime Numbers*:

“There are two facts about the distribution of prime numbers of which I hope to convince you so overwhelmingly that they will be permanently engraved in your hearts. The first is that, [they are] the most arbitrary and ornery objects studied by mathematicians: they grow like weeds among the natural numbers, seeming to obey no other law than that of chance, and nobody can predict where the next one will sprout. The second fact is even more astonishing, for it states just the opposite: that the prime numbers exhibit stunning regularity, that there are laws governing their behavior, and that they obey these laws with almost military precision.”
Part I of our book is devoted to conveying the essence of the Riemann Hypothesis and explaining why it is so intensely pursued. It requires a minimum of mathematical background, and does not, for example, use calculus, although it would be helpful to know—or to learn on the run—the meaning of the concept of function. Given its mission, Part I is meant to be complete, in that it has a beginning, middle, and end.

Part II does require calculus, and is meant as a general preparation for the type of Fourier analysis that will occur in the later chapters. The notion of spectrum is key.

Part III is for readers who wish to see, more vividly, the phenomenology that links the placement of prime numbers and (what we call there) the Riemann Spectrum.

Part IV requires some complex analysis, and returns to Riemann’s original viewpoint. In particular it relates the “Riemann Spectrum” that we discuss in Part III to the nontrivial zeroes of the Riemann’s zeta-function.

The end-notes are meant to provide more technical commentary, and link the text to references.
Part I

The Riemann Hypothesis
Chapter 1

Thoughts about numbers: ancient, medieval, and modern

If we are to believe the ancient Greek philosopher Aristotle the early Pythagoreans thought that the principles governing Number are “the principles of all things,” the concept of Number being more basic than earth, air, fire, or water, which were according to ancient tradition the four building blocks of matter. To think about number is to get close to the architecture of “what is.”

So, how far along are we in our thoughts about numbers?

Figure 1.1: René Descartes

The French philosopher and mathematician René Descartes, almost four centuries ago, expressed the hope that there soon would be “almost nothing more to discover in geometry.” Contemporary physicists dream of a final theory \[4\]. But despite its venerability and its great power and beauty, the pure mathematics of numbers may still be in the infancy of its development, with depths to be explored as endless as the human soul, and never a final theory.
Numbers are obstreperous things. Don Quixote encountered this when he requested that the “bachelor” compose a poem to his lady Dulcinea del Toboso, the first letters of each line spelling out her name. The “bachelor” found

“It must fit in, however, you do it,” pleaded Quixote, not willing to grant the imperviousness of the number 17 to division.

Seventeen is indeed a prime number: there is no way of factoring it as the product of smaller numbers, and this accounts—people tell us—for its occurrence in some phenomena of nature, as when the 17-year cicadas all emerged to celebrate a “reunion” of some sort in our fields and valleys.

Prime numbers, despite their primary position in our modern understanding of number, were not specifically doted over in the ancient literature before Euclid, at least not in the literature that has been preserved. Primes are mentioned as
a class of numbers in the writings of Philolaus (a predecessor of Plato); they are not mentioned specifically in the Platonic dialogues, which is surprising given the intense interest Plato had in mathematical developments; and they make an occasional appearance in the writings of Aristotle, which is not surprising, given Aristotle’s emphasis on the distinction between the *composite* and the *incomposite*. “The incomposite is prior to the composite,” writes Aristotle in Book 13 of the Metaphysics.

The notion of prime number occurs in Euclid’s *Elements* and play a role there as the extraordinary mathematical concept, central to any deep understanding of arithmetic phenomena, that it is now understood to be.

There is an extraordinary wealth of established truths about whole numbers; these truths provoke sheer awe for the beautiful complexity of prime numbers. But each of the important new discoveries we make give rise to a further richness of questions, educated guesses, heuristics, expectations, and unsolved problems.
Chapter 2

What are prime numbers?

*Primes as atoms.* To begin from the beginning, think of the operation of multiplication as a bond that ties numbers together: the equation $2 \times 3 = 6$ invites us to imagine the number 6 as (a molecule, if you wish) built out of its smaller constituents 2 and 3. Reversing the procedure, if we start with a whole number, say 6 again, we may try to factor it (that is, express it as a product of smaller whole numbers) and, of course, we would eventually, if not immediately, come up with $6 = 2 \times 3$ and discover that 2 and 3 factor no further; the numbers 2 and 3, then, are the indecomposable entities (atoms, if you wish) that comprise our number.

![Diagram of 6 = 2 × 3]

Figure 2.1: The number $6 = 2 \times 3$

By definition, a **prime number** (colloquially, *a prime*) is a whole number, bigger than 1, that cannot be factored into a product of two smaller whole numbers. So, 2 and 3 are the first two prime numbers. The next number along the line, 4, is not prime, for $4 = 2 \times 2$; the number after that, 5, is. Primes are, multiplicatively speaking, the building blocks from which all numbers can be made. A fundamental theorem of arithmetic tells us that any number (bigger than 1) can be factored as a product of primes, and the factorization is unique except for rearranging the order of the primes.

For example, if you try to factor the number 12 as a product of smaller numbers—ignoring the order of the factors—there are two ways to begin to do this:

$$12 = 2 \times 6 \quad \text{and} \quad 12 = 3 \times 4$$

But neither of these ways is a full factorization of 12, for both 6 and 4 are
not prime, so can be themselves factored, and in each case after changing the ordering of the factors we arrive at:

$$12 = 2 \times 2 \times 3.$$ 

![Figure 2.2: Factorizations of 12](image)

If you try to factor the number 300, there are many ways to begin:

$$300 = 30 \times 10 \quad \text{or} \quad 300 = 6 \times 50$$

and there are various other starting possibilities. But if you continue the factorization (“climbing down” any one of the possible “factoring trees”) to the bottom, where every factor is a prime number as in Figure 2.3, you always end up with the same collection of prime numbers:

$$300 = 2^2 \times 3 \times 5^2.$$ 

![Figure 2.3: Factor trees that illustrates the factorization of 300 as a product of primes](image)
The Riemann Hypothesis probes the question: how intimately can we know prime numbers, those atoms of multiplication? Prime numbers are an important part of our daily lives. For example, anytime we visit a website and purchase something online, prime numbers having hundreds of decimal digits are used to keep our bank transactions private. This ubiquitous use to which these giant primes are put depends upon a very simple principle: it is much easier to multiply numbers together than to factor them. If you had to factor, say, the number 391 you might scratch your head for a few minutes before discovering that 391 is $17 \times 23$. But if you had to multiply 17 by 23 you would do it straightaway. Offer two primes, say, $P$ and $Q$ each with a few hundred digits, to your computing machine and ask it to multiply them together: you will get their product $N = P \times Q$ with its hundreds of digits in about a microsecond. But present that number $N$ to any current desktop computer, and ask it to factor $N$, and the computer will (almost certainly) fail to do the task. [7] [8]

The safety of much encryption depends upon this guaranteed failure! [9]

If we were latter-day number-phenomenologists we might revel in the discovery and proof that

$$p = 2^{43,112,609} - 1 = 3164702693 \ldots \ldots \text{(millions of digits)} \ldots \ldots 6697152511$$

is a prime number, this number having 12,978,189 digits! This prime, which was discovered on August 23, 2008 by the GIMPS project (see [10]), is the first prime ever found with more than ten million digits.

Now $2^{43,112,609} - 1$ is quite a hefty number! Suppose someone came up to you saying “surely $p = 2^{43,112,609} - 1$ is the largest prime number!” (which it is not) how might you convince that person that he or she is wrong? [11]

Here is a neat—and, we hope, convincing—strategy to show there are prime numbers even larger than $p = 2^{43,112,609} - 1$. Imagine forming the following
humungous number: let $M$ be the product of all prime numbers up to and including $p = 2^{43,112,609} - 1$. Now go one further than $M$ by taking the next number $N = M + 1$.

OK, even though this number $N$ is wildly large, it is either a prime number itself—which would mean that there would indeed be a prime number larger than $p = 2^{43,112,609} - 1$, namely $N$; or in any event it is surely divisible by some prime number, call it $P$.

Here, now, is a way of seeing that this $P$ is bigger than $p$: Since every prime number smaller than or equal to $p$ divides $M$, these prime numbers cannot divide $N = M + 1$ (since they divide $M$ evenly, if you tried to divide $N = M + 1$ by any of them you would get a remainder of 1). So, since $P$ does divide $N$ it must not be any of the smaller prime numbers: $P$ is therefore a prime number bigger than $p = 2^{43,112,609} - 1$.

This strategy, by the way, is not very new: it is, in fact, well over two thousand years old, since it already occurred in Euclid’s Elements. The Greeks did know that there are infinitely many prime numbers and they showed it via the same method as we showed that our $p = 2^{43,112,609} - 1$ is not the largest prime number.

Here is the argument again, given very succinctly: Given primes $p_1, \ldots, p_m$, let $n = p_1 p_2 \cdots p_m + 1$. Then $n$ is divisible by some prime not equal to any $p_i$, so there are more than $m$ primes.

You can think of this strategy as a simple game that you can play. Start with any finite bag of prime numbers (say the bag that only contains one prime, the prime 2). Now each “move” of the game consists of multiplying together all the primes you have in your bag to get a number $M$, then adding 1 to $M$ to get the even larger number $N = M + 1$, then factoring $N$ into prime number factors, and then including all those new prime numbers in your bag. Euclid’s proof gives us that we will—with each move of this game—be finding more prime numbers: the bag will increase. After, say, a million moves our bag will be guaranteed to contain more than a million prime numbers.

For example, starting the game with your bag containing only one prime number 2, here is how your bag grows with after successive moves of the game:

\[
\{2\}
\{2, 3\}
\{2, 3, 7\}
\{2, 3, 7, 43\}
\{2, 3, 7, 43, 13, 139\}
\{2, 3, 7, 43, 13, 139, 3263443\}
\{2, 3, 7, 43, 13, 139, 3263443, 547, 607, 1033, 31051\}
\{2, 3, 7, 43, 13, 139, 3263443, 547, 607, 1033, 31051, 29881, 67003, 9119521, 6212157481\}
\]

etc. [12]

Though there are infinitely many primes, actually finding them is a major chal-
In the 1990s, the Electronic Frontier Foundation [http://www.eff.org/awards/coop](http://www.eff.org/awards/coop) offered a $100,000 cash reward to the first group to find a prime with at least 10,000,000 decimal digits (the record prime $p$ above won this prize [13]), and offers another $150,000 cash prize to the first group to find a prime with at least 100,000,000 decimal digits.

The number $p = 2^{43,112,609} - 1$ was for a time the largest prime known, where by “know” we mean that we know it so explicitly that we can compute things about it. For example, the last two digits of $p$ are both 1 and the sum of the digits of $p$ is 58,416,637. Of course $p$ is not the largest prime number since there are infinitely many primes, e.g., the next prime $q$ after $p$ is a prime. But there is no known way to efficiently compute anything interesting about $q$. For example, what is the last digit of $q$ in its decimal expansion?
Chapter 3

“Named” Prime Numbers

Prime numbers come in all sorts of shapes, some more convenient to deal with than others. For example, the number we have been talking about,

\[ p = 2^{43,112,609} - 1, \]

is given to us, by its very notation, in a striking form; i.e., one less than a power of 2. It is no accident that the largest “currently known” prime number has such a form. This is because there are special techniques we can draw on to show primality of a number, if it is one less than a power of 2 and—of course—if it also happens to be prime. The primes of that form have a name, Mersenne Primes, as do the primes that are one more than a power of 2, those being called Fermat Primes. [14]

Here are two exercises that you might try to do, if this is your first encounter with primes that differ from a power of 2 by 1:

1. Show that if a number of the form \( M = 2^n - 1 \) is prime, then the exponent \( n \) is also prime. [Hint: This is equivalent to proving that if \( n \) is composite, then \( 2^n - 1 \) is also composite.] For example: \( 2^2 - 1 = 3 \), \( 2^3 - 1 = 7 \) are primes, but \( 2^4 - 1 = 15 \) is not. So Mersenne primes are numbers that are

- of the form

\[ 2^{\text{prime number}} - 1, \]

and

- are themselves prime numbers.

2. Show that if a number of the form \( F = 2^n + 1 \) is prime, then the exponent \( n \) is a power of two. For example: \( 2^2 + 1 = 5 \) is prime, but \( 2^3 + 1 = 9 \) is not. So Fermat primes are numbers that are
• of the form $2^\text{power of two} + 1$, 

and

• are themselves prime numbers.

Not all numbers of the form $2^\text{prime number} - 1$ or of the form $2^\text{power of two} + 1$ are prime. We currently know only finite many primes of either of these forms. How we have come to know what we know is an interesting tale. See, for example, [http://www.mersenne.org/](http://www.mersenne.org/)
Chapter 4

Sieves

Eratosthenes, the mathematician from Cyrene (and later, librarian at Alexandria) explained how to sift the prime numbers from the series of all numbers: in the sequence of numbers,

\[
\begin{align*}
2 & \quad 3 & \quad 4 & \quad 5 & \quad 6 & \quad 7 & \quad 8 & \quad 9 & \quad 10 & \quad 11 & \quad 12 & \quad 13 & \quad 14 & \quad 15 & \quad 16 & \quad 17 & \quad 18 & \quad 19 & \quad 20 & \quad 21 & \quad 22 & \quad 23 & \quad 24 & \quad 25 & \quad 26,
\end{align*}
\]

for example, start by circling the 2 and crossing out all the other multiples of 2. Next, go back to the beginning of our sequence of numbers and circle the first number that is neither circled nor crossed out (that would be, of course, the 3), then cross out all the other multiples of 3. This gives the pattern: go back again to the beginning of our sequence of numbers and circle the first number that is neither circled nor crossed out; then cross out all of its other multiples. Repeat this pattern until all the numbers in our sequence are either circled, or crossed out, the circled ones being the primes.

In Figures 4.1–4.4 we use the primes 2, 3, 5, and finally 7 to sieve out the primes up to 100, where instead of crossing out multiples we grey them out, and instead of circling primes we color their box red.
Since all the even numbers greater than two are eliminated as being composite numbers and not primes they appear as gray in Figure 4.1, but none of the odd numbers are eliminated so they still appear in white boxes.

Looking at Figure 4.3, we see that for all but six numbers up to 100 we have (after sieving by 2, 3, and 5) determined which are primes and which composite.
Figure 4.4: Using the primes 2,3,5,7 to sieve for primes up to 100

Finally, as in Figure 4.4 sieving by 2,3,5, and 7 we have determined which are primes for all but two numbers. [15]
Chapter 5

Questions about primes
that any person might ask

We become quickly stymied when we ask quite elementary questions about the spacing of the infinite series of prime numbers.

For example, are there infinitely many pairs of primes whose difference is 2? The sequence of primes seems to be rich in such pairs

\[ 5 - 3 = 2, \quad 7 - 5 = 2, \quad 13 - 11 = 2, \quad 19 - 17 = 2, \]

and we know that there are loads more such pairs (see [16]) but the answer to our question, are there infinitely many?, is not known. Nevertheless there is very exciting recent work in this direction, specifically, Yitang Zhang proved that there are infinitely many pairs of primes that differ by no more than \( 7 \times 10^7 \). See the wikipedia entry [http://en.wikipedia.org/wiki/Yitang_Zhang]. Are there infinitely many pairs of primes whose difference is 4? Answer: equally unknown.

Is every even number greater than 2 a sum of two primes? Answer: unknown.

Are there infinitely many primes which are 1 more than a perfect square? Answer: unknown.

Remember the Mersenne prime \( p = 2^{43,112,609} - 1 \)? and how we proved—by pure thoughts—that there must be a prime \( P \) larger than that? Suppose, though, someone asked us whether there was a Mersenne Prime larger than this \( p \); that is, is there a prime number of the form

\[ 2^{\text{some prime number}} - 1 \]

bigger than \( p = 2^{43,112,609} - 1 \)? Answer: For many years we did not know; however, in 2013 Curtis Cooper discovered the even bigger Mersenne prime \( 2^{57,885,161} - 1 \), with a whopping 17,425,170 digits! Again we can ask if there is a Mersenne prime larger than Cooper’s. Answer: we do not know. It is possible
that there are infinitely many Mersenne primes but we’re far from being able
to answer such questions.

Figure 5.1: Mersenne

Is there some neat formula giving the next prime? More specifically, If I give you
a number $N$, say $N = $ one million, and ask you for the first number after $N$ that
is prime, is there a method that answers that question without, in some form or
other, running through each of the successive odd numbers after $N$ rejecting the
nonprimes until the first prime is encountered? Answer: unknown.

One can think of many ways of “getting at” some understanding of the placement
of prime numbers among all numbers. Up to this point we have been mainly
just counting them, trying to answer the question “how many primes are there
up to $X$?” and we have begun to get some feel for the numbers behind this
question, and especially for the current “best guesses” about estimates.

What is wonderful about this subject is that people attracted to it cannot resist
asking questions that lead to interesting, and sometimes surprising numerical
experiments. Moreover, given our current state of knowledge, many of the
questions that come to mind are still unapproachable: we don’t yet know enough
about numbers to answer them. But asking interesting questions about the
mathematics that you are studying is a high art, and is probably a necessary
skill to acquire, in order to get the most enjoyment—and understanding—from
mathematics. So, we offer this challenge to you:

Come up with with your own question about primes that

- is interesting to you,
- is not a question whose answer is known to you,
- is not a question that you’ve seen before; or at least not exactly,
• is a question about which you can begin to make numerical investigations.

If you are having trouble coming up with a question, read on for more examples that provide further motivation.
Chapter 6

Further questions about primes

Let us, for variety, dice the question differently by concentrating on the gaps between one prime and the next, rather than the tally of all primes. Of course, it is no fun at all to try to guess how many pairs of primes $p, q$ there are with gap $q - p$ equal to a fixed odd number, since the difference of two odd numbers is even. The fun, though, begins in earnest if you ask for pairs of primes with difference equal to 2 (these being called twin primes) for it has long been guessed that there are infinitely many such pairs of primes, but no one has been able to prove this yet.

As of 2013, the largest known twin primes are

$$3756801695685 \cdot 2^{666669} \pm 1$$

These enormous primes have 200700 digits each (see [17]).

Similarly, it interesting to consider primes $p$ and $q$ with difference 4, or 8, or— in fact— any even number $2k$. That is, people have guessed that there are infinitely many pairs of primes with difference 4, with difference 6, etc. but none of these guesses have yet been proved.

So, define

$$\text{Gap}_k(X)$$

to be the number of pairs of successive primes $(p, q)$ with $q < X$ that have “gap $k$” (i.e., such that their difference $q - p$ is $k$). Here $p$ is a prime, $q > p$ is a prime, and there are no primes between $p$ and $q$. For example, $\text{Gap}_2(10) = 2$, since the pairs $(3, 5)$ and $(5, 7)$ are the pairs less than 10 with gap 2. See Table 6.1 for various values of $\text{Gap}_k(X)$ and Figure 6.1 for the distribution of prime gaps for $X = 10^7$. 

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CHAPTER 6. FURTHER QUESTIONS ABOUT PRIMES

Table 6.1: Values of \( \text{Gap}_k(X) \)

<table>
<thead>
<tr>
<th>( X )</th>
<th>( \text{Gap}_2(X) )</th>
<th>( \text{Gap}_4(X) )</th>
<th>( \text{Gap}_6(X) )</th>
<th>( \text{Gap}_8(X) )</th>
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<td>10</td>
<td>2</td>
<td>1</td>
<td>0</td>
<td>0</td>
</tr>
<tr>
<td>( 10^2 )</td>
<td>8</td>
<td>8</td>
<td>7</td>
<td>1</td>
</tr>
<tr>
<td>( 10^3 )</td>
<td>35</td>
<td>40</td>
<td>44</td>
<td>15</td>
</tr>
<tr>
<td>( 10^4 )</td>
<td>205</td>
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<td>299</td>
<td>101</td>
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<td>( 10^6 )</td>
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<td>8143</td>
<td>13549</td>
<td>5569</td>
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<td>( 10^7 )</td>
<td>58980</td>
<td>58621</td>
<td>99987</td>
<td>42352</td>
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</tbody>
</table>

Figure 6.1: Frequency histogram showing the distribution of prime gaps of size \( \leq 50 \) for all primes up to \( 10^7 \). Six is the most popular gap in this data. The vertical axis labels such as “6e4” mean \( 6 \cdot 10^4 = 60,000 \).

Figure 6.2: Plots of \( \text{Gap}_k(X) \) for \( k = 2, 4, 6, 8 \). Which wins?

Here is yet another question that deals with the spacing of prime numbers that we do not know the answer to:

**Racing Gap 2, Gap 4, Gap 6, and Gap 8 against each other:**

Challenge: As \( X \) tends to infinity which of \( \text{Gap}_2(X) \), \( \text{Gap}_4(X) \), \( \text{Gap}_6(X) \), or \( \text{Gap}_8(X) \) do you think will grow faster? How much would you bet on the truth of your guess? [18]
Here is a curious question that you can easily begin to check out for small numbers. We know, of course, that the even numbers and the odd numbers are nicely and simply distributed: after every odd number comes an even number, after every even, an odd, there are an equal number of odd number as even numbers less than any given odd number, and there may be nothing else of interest to say about the matter. Things change considerably, though, if we focus our concentration on multiplicatively even numbers and multiplicatively odd numbers.

A **multiplicatively even** number is one that can be expressed as a product of *an even number* of primes; and a **multiplicatively odd** number is one that can be expressed as a product of *an odd number* of primes. So, any prime is multiplicatively odd, the number $4 = 2 \cdot 2$ is multiplicatively even, and so is $6 = 2 \cdot 3$, $9 = 3 \cdot 3$, and $10 = 2 \cdot 5$; but $12 = 2 \cdot 2 \cdot 3$ is multiplicatively odd. Below we list the numbers up to 25, and underline and bold the multiplicatively odd numbers.

1 2 3 4 5 6 7 8 9 10 11 12 13 14 15 16 17 18 19 20 21 22 23 24 25

Table 6.2 gives some data:

Table 6.2: Count of multiplicatively odd and multiplicatively even positive numbers ≤ X

<table>
<thead>
<tr>
<th>X</th>
<th>1</th>
<th>2</th>
<th>3</th>
<th>4</th>
<th>5</th>
<th>6</th>
<th>7</th>
<th>8</th>
<th>9</th>
<th>10</th>
<th>11</th>
<th>12</th>
<th>13</th>
<th>14</th>
<th>15</th>
<th>16</th>
</tr>
</thead>
<tbody>
<tr>
<td>m. odd</td>
<td>0</td>
<td>1</td>
<td>2</td>
<td>2</td>
<td>3</td>
<td>3</td>
<td>4</td>
<td>5</td>
<td>5</td>
<td>5</td>
<td>6</td>
<td>7</td>
<td>8</td>
<td>8</td>
<td>8</td>
<td>8</td>
</tr>
<tr>
<td>m. even</td>
<td>1</td>
<td>1</td>
<td>1</td>
<td>2</td>
<td>2</td>
<td>3</td>
<td>3</td>
<td>3</td>
<td>4</td>
<td>5</td>
<td>5</td>
<td>5</td>
<td>5</td>
<td>6</td>
<td>7</td>
<td>8</td>
</tr>
</tbody>
</table>

Now looking at this data, a natural, and simple, question to ask about the concept of multiplicative oddness and evenness is:

*Is there some $X \geq 2$ for which there are more multiplicatively even numbers less than or equal to $X$ than multiplicatively odd ones?*

Each plot in Figure 6.3 gives the number of multiplicatively odd numbers between 2 and $X$ minus the number of multiplicatively even numbers between 2 and $X$, for $X$ equal to 10, 100, 1000, 10000, 100000, and 1000000. The above question asks whether these graphs would, for sufficiently large $X$, ever cross the $X$-axis.
A negative response to this question—i.e., a proof that any plot as drawn in Figure 6.3 never crosses the X-axis—would imply the Riemann Hypothesis! In contrast to the list of previous questions, the answer to this question is known (see [19]): alas, there is such an $X$. In 1960, Lehman showed that for $X = 906,400,000$ there are 708 more multiplicatively even numbers up to $X$ than multiplicatively odd numbers (Tanaka found in 1980 that the smallest $X$ such that there are more multiplicative even than odd numbers is $X = 906,150,257$).

These are questions that have been asked about primes (and we could give bushels more as is done in [20]), questions expressible in simple vocabulary, that we can’t answer today. We have been studying numbers for over two millenia and yet we are indeed in the infancy of our understanding.

So we’ll continue our discussion by returning to the simplest counting question about prime numbers.
How many primes are there?

Slow as we are to understand primes, at the very least we can try to count them. You can see that there are 10 primes less than 30, so you might encapsulate this by saying that the chances that a number less than 30 is prime is 1 in 3. This frequency does not persist, though; here is some more data: There are 25 primes less than 100 (so 1 in 4 numbers up to 100 are prime), there are 168 primes less than a thousand (so we might say that among the numbers less than a thousand the chances that one of them is prime is roughly 1 in 6).
There are 78,498 primes less than a million (so we might say that the chances that a random choice among the first million numbers is prime have dropped to roughly 1 in 13).

There are 455,052,512 primes less than ten billion; i.e., 10,000,000,000 (so we might say that the chances are down to roughly 1 in 22).

Primes, then, seem to be thinning out. We return to the sifting process we carried out earlier, and take a look at a few graphs, to get a sense of why that might be so. There are a 100 numbers less than or equal to 100, a thousand numbers less than or equal to 1000, etc.: the shaded graph in Figure 7.4 that looks like a regular staircase, each step the same length as each riser, climbing up at, so to speak, a 45 degree angle, counts all numbers up to and including X.

Following Eratosthenes, we have sifted those numbers, to pan for primes. Our first move was to throw out roughly half the numbers (the even ones!) after the number 2. The cross-hatched bar graph in this figure which is, with one hiccup, a regular staircase climbing at a smaller angle, each step twice the lengths of each riser, illustrates the numbers that are left after one pass through Eratosthenes’ sieve, which includes, of course, all the primes. So, the chances that a number
bigger than 2 is prime is at most 1 in 2. Our second move was to throw out a good bunch of numbers bigger than 3. So, the chances that a number bigger than 3 is prime is going to be even less. And so it goes: with each move in our sieving process we are winnowing the field more extensively, reducing the chances that the later numbers are prime.

The red curve in these figures actually counts the primes: it is the beguilingly irregular staircase of primes. Its height above any number \( X \) on the horizontal line records the number of primes less than or equal to \( X \), the accumulation of primes up to \( X \). Refer to this number as \( \pi(X) \). So \( \pi(2) = 1 \), \( \pi(3) = 2 \), \( \pi(30) = 10 \); of course, could plot a few more values of \( \pi(X) \), like \( \pi(\text{ten billion}) = 455,052,512 \).

Let us accompany Eratosthenes for a few further steps in his sieving process. Figure 7.6 contains a graph of all whole numbers up to 100 after we have removed the even numbers greater than 2, and the multiples of 3 greater than 3 itself.

From this graph you can see that if you go “out a way” the likelihood that
a number is a prime is less than 1 in 3. Figure 7.7 contains a graph of what Eratosthenes sieve looks like up to 100 after sifting 2, 3, 5, and 7.

This data may begin to suggest to you that as you go further and further out on the number line the percentage of prime numbers among all whole numbers tends towards 0% (it does).

To get a sense of how the primes accumulate, we will take a look at the staircase of primes for \( X = 25 \) and \( X = 100 \) in Figures 7.8 and 7.9.
Chapter 8

Prime numbers viewed from a distance

The striking thing about these figures is that as the numbers get large enough, the jagged accumulation of primes, those quintessentially discrete entities, becomes smoother and smoother, to the eye. How strange and wonderful to watch, as our viewpoint zooms out to larger ranges of numbers, the accumulation of primes taking on such a smooth and elegant shape.

Figure 8.1: Staircases of primes up to 1,000 and 10,000

Figure 8.2: Primes up to 100,000. Note that the axis label “6e4” means $6 \times 10^4$. 

33
But don’t be fooled by the seemingly smooth shape of the curve in the last figure above: it is just as faithful a reproduction of the staircase of primes as the typographer’s art can render, for there are thousands of tiny steps and risers in this curve, all hidden by the thickness of the print of the drawn curve in the figure. It is already something of a miracle that we can approximately describe the build-up of primes, somehow, using a smooth curve. But what smooth curve?

That last question is not rhetorical. If I draw a curve with chalk on the blackboard, this can signify a myriad of smooth (mathematical) curves all encompassed within the thickness of the chalk-line, all—if you wish—reasonable approximations of one another. So, there are many smooth curves that fit the chalk-curve. With this warning, but very much fortified by the data of Figure 8.2, let us ask: what is a smooth curve that is a reasonable approximation to the staircase of primes?
Chapter 9

Pure and applied mathematics

Mathematicians seems to agree that, loosely speaking, there are two types of mathematics: pure and applied. Usually—when we judge whether a piece of mathematics is pure or applied—this distinction turns on whether or not the math has application to the “outside world,” i.e., that world where bridges are built, where economic models are fashioned, where computers churn away on the Internet (for only then do we unabashedly call it applied math), or whether the piece of mathematics will find an important place within the context of mathematical theory (and then we label it pure). Of course, there is a great overlap (as we will see later, Fourier analysis plays a major role both in data compression and in pure mathematics).

Moreover, many questions in mathematics are “hustlers” in the sense that, at first view, what is being requested is that some simple task be done (e.g., the question raised in this book, to find a smooth curve that is a reasonable approximation to the staircase of primes). And only as things develop is it discovered that there are payoffs in many unexpected directions, some of these payoffs being genuinely applied (i.e., to the practical world), some of these payoffs being pure (allowing us to strike behind the mask of the mere appearance of the mathematical situation, and get at the hidden fundamentals that actually govern the phenomena), and some of these payoffs defying such simple classification, insofar as they provide powerful techniques in other branches of mathematics. The Riemann Hypothesis—even in its current unsolved state—has already shown itself to have all three types of payoff.

The particular issue before us is, in our opinion, twofold, both applied, and pure: can we curve-fit the “staircase of primes” by a well approximating smooth curve given by a simple analytic formula? The story behind this alone is marvelous, has a cornucopia of applications, and we will be telling it below. But our
curiosity here is driven by a question that is pure, and less amenable to precise formulation: are there mathematical concepts at the root of, and more basic than (and “prior to,” to borrow Aristotle’s use of the phrase,) prime numbers—concepts that account for the apparent complexity of the nature of primes?
Chapter 10

A probabilistic “first” guess

The search for such approximating curves began, in fact, two centuries ago when Carl Friedrich Gauss defined a certain beautiful curve that, experimentally, seemed to be an exceptionally good fit for the staircase of primes.

Figure 10.2: Plot of $\pi(X)$ and Gauss’s smooth curve $G(X)$
Let us denote Gauss’s curve \( G(X) \); it has an elegant simple formula comprehensible to anyone who has had a tiny bit of calculus. If you make believe that the chances that a number \( X \) is a prime is inversely proportional to the number of digits of \( X \) you might well hit upon Gauss’s curve. That is,

\[
G(X) \quad \text{is roughly proportional to} \quad \frac{X}{\text{the number of digits of } X}.
\]

But to describe Gauss’s guess precisely we need to discuss the natural logarithm “\( \log(X) \)” which is an elegant smooth function of real numbers \( X \) that is roughly proportional to the number of digits of the whole number part of \( X \).

Euler’s famous constant \( e = 2.71828182\ldots \), which is the limit of the sequence

\[
\left(1 + \frac{1}{2}\right)^2, \left(1 + \frac{1}{3}\right)^3, \left(1 + \frac{1}{4}\right)^4, \ldots,
\]

is used in the definition of \( \log \):

\[
A = \log(X) \quad \text{is the number } A \quad \text{for which } e^A = X.
\]

Before electronic calculators, logarithms were frequently used to speed up calculations, since logarithms translate difficult multiplication problems into easier addition problems which can be done mechanically. Such calculations use that the logarithm of a product is the sum of the logarithms of the factors; that is,

\[
\log(X \cdot Y) = \log(X) + \log(Y).
\]

Figure 10.3: Plot of the natural logarithm \( \log(X) \)

Figure 10.4: A slide rule computes \( 2X \) by using that \( \log(2X) = \log(2) + \log(X) \)
In Figure 10.4 the numbers printed (on each of the slidable pieces of the rule) are spaced according to their logarithms, so that when one slides the rule arranging it so that the printed number $X$ on one piece lines up with the printed number 1 on the other, we get that for every number $Y$ printed on the first piece, the printed number on the other piece that is aligned with it is the product $XY$; in effect the “slide” adds $\log(X)$ to $\log(Y)$ giving $\log(XY)$.

In 1791, when Gauss was 14 years old, he received a book that contained logarithms of numbers up to 7 digits and a table of primes up to 10,009. Years later, in a letter written in 1849 (see Figure 10.5), Gauss claimed that as early as 1792 or 1793 he had already observed that the density of prime numbers over intervals of numbers of a given rough magnitude $X$ seemed to average $1/\log(X)$.

Very very roughly speaking, this means that the number of primes up to $X$ is approximately $X$ divided by twice the number of digits of $X$. For example, the number of primes less than 99 should be roughly

$$\frac{99}{2 \times 2} = 24.75 \approx 25,$$

which is pretty amazing, since the correct number of primes up to 99 is 25. The number of primes up to 999 should be roughly

$$\frac{999}{2 \times 3} = 166.5 \approx 167,$$
which is again close, since there are 168 primes up to 1000. The number of primes up to 999,999 should be roughly

\[ \frac{999999}{2 \times 6} = 83333.25 \approx 83,333, \]

which is close to the correct count of 78,498.

Gauss guessed that the expected number of primes up to \( X \) is approximated by the area under the graph of \( \frac{1}{\log(X)} \) from 2 to \( X \) (see Figure 10.6). The area under \( \frac{1}{\log(X)} \) up to \( X = 999,999 \) is 78,626.43..., which is remarkably close to the correct count 78,498 of the primes up to 999,999.

Figure 10.6: The expected tally of the number of primes \( \leq X \) is approximated by the area underneath the graph of \( \frac{1}{\log(X)} \) from 1 to \( X \).

Gauss was an inveterate computer: he wrote in his 1849 letter that there are 216,745 prime numbers less than three million (This is wrong: the actual number of these primes is 216,816). Gauss’s curve \( G(X) \) predicted that there would be 216,970 primes—a miss, Gauss thought, by

\[ 225 = 216,970 - 216,745 \]

(but actually he was closer than he thought: the prediction of the curve \( G(X) \) missed by a mere 154 = 216,970 − 216,816) Gauss’s computation brings up two queries: will this spectacular “good fit” continue for arbitrarily large numbers? and, the (evidently prior) question: what counts as a good fit?
Chapter 11

What is a “good approximation”? 

If you are trying to estimate a number, say, around ten thousand, and you get it right to within a hundred, let us celebrate this kind of accuracy by saying that you have made an approximation with square-root error (\(\sqrt{10,000} = 100\)). Of course, we should really use the more clumsy phrase “an approximation with at worst square-root error.” Sometimes we’ll simply refer to such approximations as good approximations. If you are trying to estimate a number in the millions, and you get it right to within a thousand, let’s agree that—again—you have made an approximation with square-root error (\(\sqrt{1,000,000} = 1,000\)). Again, for short, call this a good approximation. So, when Gauss thought his curve missed by 226 in estimating the number of primes less than three million, it was well within the margin we have given for a “good approximation.” [21]

More generally, if you are trying to estimate a number that has \(D\) digits and you get it almost right, but with an error that has no more than, roughly, half that many digits, let us say, again, that you have made an approximation with square-root error or synonymously, a good approximation.

This rough account almost suffices for what we will be discussing below, but to be more precise, the specific gauge of accuracy that will be important to us is not for a mere single estimate of a single error term,

\[
\text{Error term} = \text{Exact Value} - \text{Our “good approximation”}
\]

but rather for infinite sequences of estimates of error terms. Generally, if you are interested in a numerical quantity \(q(X)\) that depends on the real number parameter \(X\) (e.g., \(q(X)\) could be \(\pi(X)\), “the number of primes < \(X\)” and if you have an explicit candidate “approximation,” \(q_{\text{approx}}(X)\), to this quantity, let us say that \(q_{\text{approx}}(X)\) is essentially a square-root accurate approximation to \(q(X)\) if for any given exponent greater than 0.5 (you choose it: 0.501, 0.5001, 0.50001, ... for example) and for large enough \(X\)—where the phrase
“large enough” depends on your choice of exponent—the **error term**—i.e., the difference between \( q_{\text{approx}}(X) \) and the **true** quantity, \( q(X) \), is, in absolute value, less than \( q_{\text{approx}}(X) \) raised to that exponent (e.g. \(< X^{0.501}, < X^{0.5001}, \text{etc.} \)).

Readers who know calculus and wish to have a technical formulation of this definition of **good approximation** might turn to the endnote 2 for a precise statement. If you found the above confusing, don’t worry: again, a square-root accurate approximation is one in which at least roughly half the digits are correct.

**Remark 11.1.** To get a feel for how basic the notion of **approximation to data being square root close to the true values of the data** is—and how it represents the “gold standard” of accuracy for approximations, consider this fable.

Imagine that the devil had the idea of saddling a large committee of people with the task of finding values of \( \pi(X) \) for various large numbers \( X \). This he did in the following manner, having already worked out which numbers are prime himself. Since the devil is, as everyone knows, **into the details**, he has made no mistakes: his work is entirely correct. He gives each committee member a copy of the list of all **prime numbers** between 1 and one of the large numbers \( X \) in which he was interested. Now each committee member would count the number of primes by doing nothing more than considering each number, in turn, on their list and tallying them up, much like a canvasser counting votes. But since they are human, they will indeed be making mistakes, say 0.001% of the time. Assume further that it is just as likely for them to make the mistake of undercounting or overcounting. If many people are engaged in such pursuit, some of them might over-count \( \pi(X) \); some of them might under-count it. The average error (over-counted or undercounted) would be proportional to \( \sqrt{X} \).
Chapter 12

What is Riemann’s Hypothesis? (first formulation)

Recall from Chapter [10] that a rough approximation to $\pi(X)$, the number of primes $< X$, is given by the function $X/\log(X)$; and Gauss’s guess for an approximation to $\pi(X)$ was in terms of the area of the region from 2 to $X$ under the graph of $1/\log(X)$, a quantity sometimes referred to as $\text{Li}(X)$. “Li” (pronounced Li) is short for logarithmic integral, because the area of the region from 2 to $X$ under $1/\log(X)$ is the integral $\int_2^X 1/\log(X)\,dt$.

Figure 12.1 contains a graph of the three functions $\text{Li}(X)$, $\pi(X)$, and $X/\log X$ for $X \leq 200$. But data, no matter how impressive, may be deceiving (as we learned in Chapter [6]). If you think that the three graphs never cross for all large values of $X$, and that we have the simple relationship $X/\log(X) < \pi(X) < \text{Li}(X)$ for large $X$, turn to this endnote. [22]

Figure 12.1: Plots of $\text{Li}(X)$ (top), $\pi(X)$ (in the middle), and $X/\log(X)$ (bottom).
Let $X = 10^{24}$. Then (see [23])

\[
\begin{align*}
\pi(X) &= 18,435,599,767,349,200,867,866 \\
\text{Li}(X) &= 18,435,599,767,366,347,775,143,10580 \ldots \\
X / (\log(X) - 1) &= 18,429,088,896,563,917,716,962,93869 \ldots \\
\text{Li}(X) - \pi(X) &= 17,146,907,277,105803 \ldots \\
\sqrt{X} \cdot \log(X) &= 55,262,042,231,857,096416 \ldots
\end{align*}
\]

Since Li$(X)$ seems to start out impressively close in value to our $\pi(X)$ (at least in this range) it is natural to hope that in full generality it is essentially a square root approximation to $\pi(X)$. This gives us our first formulation of Riemann’s Hypothesis:

**The Riemann Hypothesis (first formulation)**

For any real number $X$ the number of prime numbers less than $X$ is approximately Li$(X)$ and this approximation is essentially square root accurate. See [24].
Take a look at Figure 12.1 again. All three functions, $X/\log(X)$, $\text{Li}(X)$ and $\pi(X)$ are “going to infinity with $X$” (this means that for any real number $R$, if $X$ is taken to be sufficiently large, the values of these functions at $X$ will exceed $R$).

Are these functions “going to infinity” at the same rate?

To answer such a question, we have to know what we mean by going to infinity at the same rate. So, here’s a definition. Two functions, $A(X)$ and $B(X)$, that each go to infinity will be said to go to infinity at the same rate if their ratio

$$A(X)/B(X)$$

tends to 1 as $X$ goes to infinity.

If for example two functions, $A(X)$ and $B(X)$ that take positive whole number values, have the same number of digits for large $X$ and if, for any number you give us, say: a million (or a billion, or a trillion) and if $X$ is large enough, then the “leftmost” million (or billion, or trillion) digits of $A(X)$ and $B(X)$ are the same, then $A(X)$ and $B(X)$ go to infinity at the same rate.

While we’re defining things, let us say that two functions, $A(X)$ and $B(X)$, that each go to infinity go to infinity at similar rates if there are two positive constants $c$ and $C$ such that for $X$ sufficiently large the ratio

$$A(X)/B(X)$$

is greater than $c$ and smaller than $C$. 
For example, two polynomials in $X$ with positive leading coefficient go to infinity at the same rate if and only if they have the same degrees and the same leading coefficient; they go to infinity at similar rates if they have the same degree. See Figure 13.1.

Now a theorem from elementary calculus tells us that the ratio of $\text{Li}(X)$ to $X/\log(X)$ tends to 1 as $X$ gets larger and larger. That is—using the definition we’ve just introduced— $\text{Li}(X)$ and $X/\log(X)$ go to infinity at the same rate (see [25]).

Recall (on page 10 above) that if $X = 10^{23}$, the top eleven digits of $\pi(X)$ and $\text{Li}(X)$ are the same: 19253203916. Well, that’s a good start. Can we guarantee that for $X$ large enough, the “top” million (or billion, or trillion) digits of $\pi(X)$ and $\text{Li}(X)$ are the same? I.e., that these two functions go to infinity at the same rate?

The Riemann Hypothesis, as we have just formulated it, would tell us that the difference between $\text{Li}(X)$ and $\pi(X)$ is pretty small in comparison with the size of $X$. This information would imply (but would be much more precise information
than) the statement that the ratio $\text{Li}(X)/\pi(X)$ tends to 1, i.e., that $\text{Li}(X)$ and $\pi(X)$ go to infinity at the same rate.

This last statement gives, of course, a far less precise relationship between $\text{Li}(X)$ and $\pi(X)$ than the Riemann Hypothesis (once it is proved!) would give us. The advantage, though, of the less precise statement is that it is currently known to be true, and—in fact—has been known for over a century. It goes under the name of

**The Prime Number Theorem:** $\text{Li}(X)$ and $\pi(X)$ go to infinity at the same rate.

Since $\text{Li}(X)$ and $X/\log(X)$ go to infinity at the same rate, we could equally well have expressed the “same” theorem by saying:

**The Prime Number Theorem:** $X/\log(X)$ and $\pi(X)$ go to infinity at the same rate.

This fact is a very hard-won piece of mathematics! It was proved in 1896 independently by Hadamard and de la Vallée Poussin.

A milestone in the history leading up to the proof of Prime Number Theorem is the earlier work of Chebyshev (see [http://en.wikipedia.org/wiki/Chebyshev_function](http://en.wikipedia.org/wiki/Chebyshev_function)) showing that (to use the terminology we introduced) $X/\log(X)$ and $\pi(X)$ go to infinity at similar rates.

The elusive Riemann Hypothesis, however, is much deeper than the Prime Number Theorem, and takes its origin from some awe-inspiring, difficult to interpret, lines in Bernhard Riemann’s magnificent 8-page paper, “On the number of primes less than a given magnitude,” published in 1859 (see [26]).

Figure 13.3: From Riemann’s 1859 Manuscript

Riemann’s hypothesis, as it is currently interpreted, turns up as relevant, as a key, again and again in different parts of the subject: if you accept it as hypothesis you have an immensely powerful tool at your disposal: a mathematical magnifying glass that sharpens our focus on number theory. But it also has a wonderful protean quality—there are many ways of formulating it, any of these formulations being provably equivalent to any of the others.
This Riemann Hypothesis remains unproved to this day, and therefore is “only a hypothesis,” as Osiander said of Copernicus’s theory, but one for which we have overwhelming theoretical and numerical evidence in its support. It is the kind of conjecture that contemporary Dutch mathematician Frans Oort might label a *suffusing conjecture* in that it has unusually broad implications: many many results are now known to follow, if the conjecture, familiarly known as RH, is true. A proof of RH would, therefore, fall into the *applied* category, given our discussion above in Chapter 9. But however you classify RH, it is a central concern in mathematics to find its proof (or, a counter-example!). RH is one of the weightiest statements in all of mathematics.
We have borrowed the phrase “staircase of primes” from the popular book *The Music of Primes* by Marcus du Sautoi, for we feel that it captures the sense that there is a deeply hidden architecture to the graphs that compile the number of primes (up to $N$) and also because—in a bit—we will be tinkering with this carpentry. Before we do so, though, let us review in Figure 14.1 what this staircase looks like, for different ranges.
The mystery of this staircase is that the information contained within it is—in effect—the full story of where the primes are placed. This story seems to elude any simple description. Can we “tinker with” this staircase without destroying this valuable information?
Chapter 15

Tinkering with the carpentry of the staircase of primes

For starters, notice that all the risers of this staircase (Figure 14.1 above) have unit length. That is, they contain no numerical information except for their placement on the $x$-axis. So, we could distort our staircase by changing (in any way we please) the height of each riser; and as long as we haven’t brought new risers into—or old risers out of—existence, and have not modified their position over the $x$-axis, we have retained all the information of our original staircase.

A more drastic-sounding thing we could do is to judiciously add new steps to our staircase. At present, we have a step at each prime number $p$, and no step anywhere else. Suppose we built a staircase with a new step not only at $x = p$ for each prime number but also at $x = 1$ and $x = p^n$ where $p^n$ runs through all powers of prime numbers as well. Such a staircase would have, indeed, many more steps than our original staircase had, but, nevertheless, would retain much of the quality of the old staircase: namely it contains within it the full story of the placement of primes and their powers.

A final thing we can do is to perform a distortion of the $x$-axis (elongating or shortening it, as we wish) in any specific way, as long as we can perform the inverse process, and “undistort” it if we wish. Clearly such an operation may have mangled the staircase, but hasn’t destroyed information irretrievably.

We shall perform all three of these kinds of operations eventually, and will see some great surprises as a result. But for now, we will perform distortions only of the first two types. We are about to build a new staircase that retains the precious information we need, but is constructed according to the following architectural plan.
• We first build a staircase that has a new step precisely at \( x = 1 \), and \( x = p^n \) for every prime power \( p^n \) with \( n \geq 1 \). That is, there will be a new step at \( x = 1, 2, 3, 4, 5, 7, 8, 9, 11, \ldots \)

• Our staircase starts on the ground at \( x = 0 \) and the height of the riser of the step at \( x = 1 \) will be \( \log(2\pi) \). The length of the riser of the step at \( x = p^n \) will not be 1 (as was the length of all risers in the old staircase of primes) but rather: the step at \( x = p^n \) will have the height of its riser equal to \( \log p \). So for the first few steps listed in the previous item, the risers will be of length \( \log(2\pi) \), \( \log 2 \), \( \log 3 \), \( \log 2 \), \( \log 5 \), \( \log 7 \), \( \log 2 \), \( \log 3 \), \( \log 11 \), \ldots 
Since \( \log(p) > 1 \), these vertical dimensions lead to a steeper ascent but no great loss of information.

Although we are not quite done with our architectural work, Figure 15.1 shows what our staircase looks like, so far.

Figure 15.1: The newly constructed staircase that counts prime powers

Notice that this new staircase looks, from afar, as if it were nicely approximated by the 45 degree straight line, i.e., by the simple function \( X \). In fact, we have—by this new architecture—a second equivalent way of formulating Riemann’s hypothesis. For this, let \( \psi(X) \) denote the function of \( X \) whose graph is depicted in Figure 15.1 (see [27]).

The Riemann Hypothesis (second formulation)

This new staircase is essentially square root close to the 45 degree straight line; i.e., the function \( \psi(X) \) is essentially square root close to the function \( f(X) = X \).

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52 \hspace{1cm} \textit{CHAPTER 15. TINKERING WITH THE STAIRCASE OF PRIMES}
Do not worry if you do not understand why our first and second formulations of Riemann’s Hypothesis are equivalent. Our aim, in offering the second formulation—a way of phrasing Riemann’s guess that mathematicians know to be equivalent to the first one—is to celebrate the variety of equivalent ways we have to express Riemann’s proposed answers to the question “How many primes are there?” and to point out that some formulations would reveal a startling simplicity—not immediately apparent—to the behavior of prime numbers, no matter how erratic primes initially appear to us to be. After all, what could be simpler than a 45 degree straight line?
Chapter 16

What do computer music files, data compression, and prime numbers have to do with each other?

Sounds of all sorts—and in particular the sounds of music—travel as vibrations of air molecules at roughly 768 miles an hour. These vibrations—fluctuations of pressure—are often represented, or “pictured,” by a graph where the horizontal axis corresponds to time, and the vertical axis corresponds to pressure at that time. The very purest of sounds—a single sustained note—would look something like this (called a “sine wave”) when pictured (see Figure 16.1), so that if you fixed your position somewhere and measured air pressure due to this sound at that position, the peaks correspond to the times when the varying air pressure is maximal or minimal and the zeroes to the times when it is normal pressure.

You’ll notice that there are two features to the graph in Figure 16.1.
1. *The height of the peaks of this sine wave:* This height is referred to as the **amplitude** and corresponds to the **loudness** of the sound.

2. *The number of peaks per second:* This number is referred to as the **frequency** and corresponds to the **pitch** of the sound.

Of course, music is rarely—perhaps never—just given by a single pure sustained note and nothing else. A next most simple example of a sound would be a simple chord (say a C and an E played together on some electronic instrument that could approximate pure notes). Its graph would be just the **sum** of the graphs of each of the pure notes (see Figures 16.2 and 16.3).

![Figure 16.2: Graph of Two Sine Waves with Different Frequencies](image)

![Figure 16.3: Graph of Sum of the Two Sine Waves with Different Frequencies](image)

So the picture of the changing frequencies of this chord would be already a pretty complicated configuration. What we have described in these graphs are two sine waves (our C and our E) when they are played *in phase* (meaning they start at the same time) but we could also “delay” the onset of the E note and play them with some different phase relationship, for example, as illustrated in Figures 16.4 and 16.5.
So, all you need to reconstruct the chord graphed above is to know five numbers:

- the two frequencies—the collection of frequencies that make up the sound is called the spectrum of the sound,
- the two amplitudes of each of these two frequencies,
- the phase between them.

Now suppose you came across such a sound as pictured in Figure 16.5 and wanted to “record it.” Well, one way would be to sample the amplitude of the sound at many different times, as for example in Figure 16.6.

Then, fill in the rest of the points to obtain Figure 16.7.
But this sampling would take an enormous amount of storage space, at least compared to storing five numbers, as explained above! Current audio compact discs do their sampling 44,100 times a second to get a reasonable quality of sound.

Another way is to simply record the five numbers: the spectrum, amplitudes, and phase. Surprisingly, this seems to be roughly the way our ear processes such a sound when we hear it [28].

Even in this simplest of examples (our pure chord: the pure note C played simultaneously with pure note E) the efficiency of data compression that is the immediate bonus of analyzing the picture of the chords as built just with the five numbers giving spectrum, amplitudes, and phase is staggering.

This type of analysis, in general, is called Fourier Analysis and is one of the glorious chapters of mathematics. One way of picturing spectrum and amplitudes of a sound is by a bar graph which might be called the spectral picture of the sound, the horizontal axis depicting frequency and the vertical one depicting amplitude: the height of a bar at any frequency is proportional to the amplitude of that frequency “in” the sound.

So our CE chord would have the spectral picture in Figure [16.9]
This spectral picture ignores the phase but is nevertheless a very good portrait of the sound. The spectral picture of a graph gets us to think of that graph as “built up by the superposition of a bunch of pure waves,” and if the graph is complicated enough we may very well need infinitely many pure waves to build it up! Fourier analysis is a mathematical theory that allows us to start with any graph—we are thinking here of graphs that picture sounds, but any graph will do—and actually compute its spectral picture (and even keep track of phases).

The operation that starts with a graph and goes to its spectral picture that records the frequencies, amplitudes, and phases of the pure sine waves that, together, compose the graph is called the Fourier transform and nowadays there are very fast procedures for getting accurate Fourier transforms (meaning accurate spectral pictures including information about phases) by computer [29].

The theory behind this operation (Fourier transform giving us a spectral analysis of a graph) is quite beautiful, but equally impressive is how—given the power of modern computation—you can immediately perform this operation for yourself to get a sense of how different wave-sounds can be constructed from the superposition of pure tones.

The sawtooth wave in Figure 16.10 has a spectral picture, its Fourier transform, given in Figure 16.11.

![Figure 16.9: Spectral Picture of a CE chord](image-url)
Figure 16.12: A Complicated Sound Wave

Suppose you have a complicated sound wave, say as in Figure 16.12 and you want to record it. Standard audio CD’s record their data by intensive sampling as we mentioned. In contrast, current mp3 audio compression technology uses Fourier transforms plus sophisticated algorithms based on knowledge of which frequencies the human ear can hear. With this, mp3 technology manages to get a compression factor of 8–12 with little perceived loss in quality, so that you can fit your entire music collection on your iPod, instead of just your favorite 10 CD’s.
Chapter 17

Spectra and Trigonometric Sums

As we saw in Chapter 16, a pure tone can be represented by a periodic \textit{sine wave}—a function of time \( f(t) \)—the equation of which might be:

\[
f(t) = a \cdot \cos(b + \theta t).
\]

![Figure 17.1: Plot of the periodic sine wave \( f(t) = 2 \cdot \cos(1 + t/2) \)](image)

The \( \theta \) determines the \textit{frequency} of the periodic wave, the larger \( \theta \) is the higher the “pitch.” The coefficient \( a \) determines the envelope of size of the periodic wave, and we call it the \textit{amplitude} of the periodic wave.

Sometimes we encounter functions \( F(t) \) that are not pure tones, but that can be expressed as (or we might say “decomposed into”) a finite sum of pure tones, for example three of them:

\[
F(t) = a_1 \cdot \cos(b_1 + \theta_1 t) + a_2 \cdot \cos(b_2 + \theta_2 t) + a_3 \cdot \cos(b_3 + \theta_3 t)
\]
Figure 17.2: Plot of the sum $5 \cos (-t - 2) + 2 \cos (t/2 + 1) + 3 \cos (2t + 4)$

We’ll refer to such functions $F(t)$ as *finite trigonometric sums*, because—well—they are. In this example, there are three frequencies involved—i.e., $\theta_1, \theta_2, \theta_3$—and we’ll say that the spectrum of $F(t)$ is the set of these frequencies, i.e.,

$$\text{The spectrum of } F(t) = \{\theta_1, \theta_2, \theta_3\}.$$  

More generally we might consider a sum of any finite number of pure cosine waves—or in a moment we’ll also see some infinite ones as well. Again, for these more general trigonometric sums, their *spectrum* will denote the set of frequencies that compose them.
Chapter 18

The spectrum and the staircase of primes

In view of the amazing data-compression virtues of Fourier analysis, it isn’t unnatural to ask these questions:

- Is there a way of using Fourier analysis to better understand the complicated picture of the staircase of primes?

- Does this staircase of primes (or, perhaps, some tinkered version of the staircase that contains the same basic information) have a spectrum?

- If such a spectrum exists, can we compute it conveniently, just as we have done for the saw-tooth wave above, or for the major third CE chord?

- Assuming the spectrum exists, and is computable, will our understanding of this spectrum allow us to reproduce all the pertinent information about the placement of primes among all whole numbers, elegantly and faithfully?
• And here is a most important question: will that spectrum show us order and organization lurking within the staircase that we would otherwise be blind to?

Strangely enough, it is towards questions like these that Riemann’s Hypothesis takes us. We began with the simple question about primes: how to count them, and are led to ask for profound, and hidden, regularities in structure.
Chapter 19

To our readers of Part I

The statement of the Riemann Hypothesis—admittedly as elusive as before—has, at least, been expressed elegantly and more simply, given our new staircase that approximates (conjecturally with essential square root accuracy) a 45 degree straight line.

We have offered two equivalent formulations of the Riemann Hypothesis, both having to do with the manner in which the prime numbers are situated among all whole numbers.

In doing this, we hope that we have convinced you that—in the words of Don Zagier—primes seem to obey no other law than that of chance and yet exhibit stunning regularity. This is the end of Part I of our book, and is largely the end of our main mission, to explain—in elementary terms—what is Riemann’s Hypothesis?

For readers who have at some point studied Differential Calculus, in Part II we shall discuss Fourier analysis, a fundamental tool that will be used in Part III where we show how Riemann’s hypothesis provides a key to some deeper structure of the prime numbers, and to the nature of the laws that they obey. We will—if not explain—at least hint at how the above series of questions have been answered so far, and how the Riemann Hypothesis offers a surprise for the last question in this series.
Part II

Distributions
Chapter 20

How Calculus manages to find the slopes of graphs that have no slopes

Differential Calculus, initially the creation of Newton and/or Leibniz in the 1680s, acquaints us with slopes of graphs of functions of a real variable. So, to discuss this we should say a word about what a function is, and what its graph is.

Figure 20.1: Isaac Newton and Gottfried Leibniz created Calculus

A function (let us refer to it in this discussion as \( f \)) is often described as a kind of machine that for any specific input numerical value \( a \) will give, as output, a well-defined numerical value.

This “output number” is denoted \( f(a) \) and is called the “value” of the function \( f \) at \( a \). For example, the machine that adds 1 to any number can be thought of as the function \( f \) whose value at any \( a \) is given by the equation \( f(a) = a + 1 \). Often we choose a letter—say, \( X \)—to stand for a “general number” and we
denote the function \( f \) by the symbol \( f(X) \) so that this symbolization allows to “substitute for \( X \) any specific number \( a \)” to get its value \( f(a) \).

The graph of a function provides a vivid visual representation of the function in the Euclidean plane where over every point \( a \) on the \( x \)-axis you plot a point above it of “height” equal to the value of the function at \( a \), i.e., \( f(a) \). In Cartesian coordinates, then, you are plotting points \((a, f(a))\) in the plane where \( a \) runs through all real numbers.

![Graph of the function \( f(a) = a + 1 \)](image)

In this book we will very often be talking about “graphs” when we are also specifically interested in the functions—of which they are the graphs. We will use these words almost synonymously since we like to adopt a very visual attitude towards the behavior of the functions that interest us.

![Calculus illustration](image)

Figure 20.3 illustrates a function (blue), the slope at a point (black straight line), and the derivative (red) of the function; the red derivative is the function whose value at a point is the slope of the blue function at that point. Differential Calculus explains to us how to calculate slopes of graphs, and finally, shows us the power that we then have to answer problems we could not answer if we couldn’t compute those slopes.

Usually, in elementary Calculus classes we are called upon to compute slopes only of smooth graphs, i.e., graphs that actually have slopes at each of their points, such as in the illustration just above. What could Calculus possibly do
if confronted with a graph that has jumps, such as in Figure 20.4:

\[ f(x) = \begin{cases} 
1 & x \leq 3 \\
2 & x > 3. 
\end{cases} \]

(Note that for purely aesthetic reasons, we draw a vertical line at the point where the jump occurs, though technically that vertical line is not part of the graph of the function.)

![Graph](image-url)

Figure 20.4: The graph of the function \( f(x) \) above that jumps—it is 1 up to 3 and then 2 after that point.

The most comfortable way to deal with the graph of such a function is to just approximate it by a differentiable function as in Figure 20.5. Note that this function is not smooth, since the derivative is continuous but not differentiable; in our discussion all we will need is that our approximating function is once differentiable.

![Graph](image-url)

Figure 20.5: A picture of a smooth graph approximating the graph that is 1 up to some point \( x \) and then 2 after that point, the smooth graph being flat mostly.

Then take the derivative of that smooth function. Of course, this is just an approximation, so we might try to make a better approximation, which we do in each successive graph starting with Figure 20.6 below.
Figure 20.6: A picture of the derivative of a smooth approximation to a function that jumps.

Note that—as you would expect—in the range where the initial function is constant, its derivative is zero. In the subsequent figures, our initial function will be nonconstant for smaller and smaller intervals about the origin. Note also that, in our series of pictures below, we will be successively rescaling the $y$-axis; all our initial functions have the value 1 for “large” negative numbers and the value 2 for large positive numbers.

Figure 20.7: Second picture of the derivative of a smooth approximation to a function that jumps.

Figure 20.8: Third picture of the derivative of a smooth approximation to a function that jumps.
Figure 20.9: Fourth picture of the derivative of a smooth approximation to a function that jumps.

Notice what is happening: as the approximation gets better and better, the derivative will be zero mostly, with a blip at the point of discontinuity, and the blip will get higher and higher. In each of these pictures, for any interval of real numbers \([a, b]\) the total area under the red graph over that interval is equal to

\[
\text{the height of the blue graph at } x = b \quad \text{minus} \quad \text{the height of the blue graph at } x = a.
\]

This is a manifestation of one of the fundamental facts of life of Calculus relating a function to its derivative:

Given any real-valued function \(F(x)\)—that has a derivative—for any interval of real numbers \([a, b]\) the total area under the graph of the derivative of \(F(x)\) over that interval is equal to \(F(b) - F(a)\).

What happens if we take the series of figures 20.6–20.9, etc. to the limit? This is quite curious:

- **the series of red graphs:** these are getting thinner and thinner and higher and higher: can we make any sense of what the red graph might mean in the limit (even though the only picture of it that we have at present makes it infinitely thin and infinitely high)?

- **the series of blue graphs:** these are happily looking more and more like the tame Figure 20.4

Each of our red graphs is the derivative of the corresponding blue graph. Wouldn’t it be tempting to think of the limit of the red graphs—whatever we might construe this to be—as standing for the derivative of the limit of the blue graphs, i.e., of the graph in Figure 20.4?

Well, the temptation is so great that, in fact, mathematicians and physicists of the early twentieth century struggled to give a meaning to things like the limit of the red graphs—such things were initially called **generalized functions** which
might be considered the derivative of the limit of the blue graphs, i.e., of the graph of Figure 20.4.

Of course, to achieve progress in mathematics, all the concepts that play a role in the theory have to be unambiguously defined, and it took a while before generalized functions such as the limit of our series of red graphs had been rigorously introduced.

But many of the great moments in the development of mathematics occur when mathematicians—requiring some concept not yet formalized—work with the concept tentatively, dismissing—if need be—mental tortures, in hopes that the experience they acquire by working with the concept will eventually help to put that concept on sure footing. For example, early mathematicians (Newton, Leibniz)—in replacing approximate speeds by instantaneous velocities by passing to limits—had to wait a while before later mathematicians (e.g., Weierstrass) gave a rigorous foundation for what they were doing.

Figure 20.10: Karl Weierstrass (1815–1897) and Laurent Schwartz (1915–2002)

Karl Weierstrass, who worked during the latter part of the nineteenth century, was known as the “father of modern analysis.” He oversaw one of the glorious moments of rigorization of concepts that were long in use, but never before systematically organized. He, and other analysts of the time, were interested in providing a rigorous language to talk about functions and more specifically continuous functions and smooth (i.e., differentiable) functions. They wished to have a firm understanding of limits (i.e., of sequences of numbers, or of functions).

For Weierstrass and his companions, even though the functions they worked with needn’t be smooth, or continuous, at the very least, the functions they studied had well-defined—and usually finite—values. But our “limit of red graphs” is not so easily formalized as the concepts that occupied the efforts of Weierstrass. Happily however, this general process of approximating discontinuous functions
more and more exactly by smooth functions, and taking their derivatives to get the blip-functions as we have just seen in the red graphs above was eventually given a mathematically rigorous foundation; notably, by the French mathematician, Laurent Schwartz who provide a beautiful theory that we will not go into here, that made perfect sense of “generalized functions” such as our limit of the series of red graphs, and that allows mathematicians to work with these concepts with ease. These “generalized functions” are called distributions in Schwartz’s theory [30].
Chapter 21

Distributions: sharpening our approximating functions even if we have to let them shoot out to infinity

The curious limit of the red graphs of the previous section, which you might be tempted to think of as a “blip-function” that vanishes for $t$ nonzero and is somehow “infinite” (whatever that means) at 0 is an example of a generalized function (in the sense of the earlier mathematicians) or a distribution in the sense of Laurent Schwartz.

This particular limit of the red graphs also goes by another name (it is officially called a Dirac $\delta$-function, the adjective “Dirac” being in honor of the physicist who first worked with this concept, the “$\delta$” being the symbol he assigned to these objects). The noun “function” should be in quotation marks for, properly speaking, the Dirac $\delta$-function is not—as we have explained above—a bona fide function but rather a distribution.
Now may be a good time to summarize what the major difference is between honest functions and generalized functions or distributions.

An honest (by which we mean integrable) function of a real variable $f(t)$ possesses two “features.”

- **It has values.** That is, at any real number $t$, e.g., $t = 2$ or $t = 0$ or $t = \pi$ etc., our function has a definite real number value ($f(2)$ or $f(0)$ or $f(\pi)$ etc.) and if we know all those values we know the function.

- **It has areas under its graph.** If we are given any interval of real numbers, say the interval between $a$ and $b$, we can talk unambiguously about the area “under” the graph of the function $f(t)$ over the interval between $a$ and $b$. That is, in the terminology of Integral Calculus, we can talk of the integral of $f(t)$ from $a$ to $b$. And in the notation of Calculus, this—thanks to Leibniz—is elegantly denoted

$$\int_{a}^{b} f(t)dt.$$ 

Figure 21.2: This figure illustrates $\int_{-\infty}^{\infty} f(x)dx$, which is the signed area between the graph of $f(x)$ and the $x$-axis, where area below the $x$-axis (yellow) counts negative, and area above (grey) is positive.

In contrast, a generalized function or distribution
• **may not have** “definite values” at all real numbers if it is not an honest function, nevertheless:

• **It has well-defined areas under portions of its “graph.”** If we are given any interval of real numbers, say the (open) interval between \(a\) and \(b\), we can still talk unambiguously about the area “under” the graph of the generalized function \(D(t)\) over the interval between \(a\) and \(b\) and we will denote this—extending what we do in ordinary calculus—by the symbol

\[
\int_a^b D(t)\,dt.
\]

This description is important to bear in mind and it gives us a handy way of thinking about “generalized functions” (i.e., distributions) as opposed to functions: when we consider an (integrable) function of a real variable, \(f(t)\), we may invoke its value at every real number and for every interval \([a, b]\) we may consider the quantity \(\int_a^b f(t)\,dt\). BUT when we are given a generalized function \(D(t)\) we only have at our disposal the latter quantities. In fact, a generalized function of a real variable \(D(t)\) is (formally) nothing more than a rule that assigns to any finite interval \([a, b]\) \((a \leq b)\) a quantity that we might denote \(\int_a^b D(t)\,dt\) and that behaves as if it were the integral of a function and in particular—for three real numbers \(a \leq b \leq c\) we have the additivity relation

\[
\int_a^c D(t)\,dt = \int_a^b D(t)\,dt + \int_b^c D(t)\,dt.
\]

SO, any honest function integrable over finite intervals clearly *is* a distribution (forget about its values!) but ... there are many more generalized functions, and including them in our sights gives us a very important tool.

It is natural to talk, as well, of Cauchy sequences, and limits, of distributions. We’ll say that such a sequence \(D_1(t), D_2(t), D_3(t), \ldots\) is a **Cauchy sequence** if for every interval \([a, b]\) the quantities

\[
\int_a^b D_1(t)\,dt, \quad \int_a^b D_2(t)\,dt, \quad \int_a^b D_3(t)\,dt, \ldots
\]

form a Cauchy sequence of real numbers. Now, any Cauchy sequence of distributions **converges to a limiting distribution** \(D(t)\) which is defined by the rule that for every interval \([a, b]\),

\[
\int_a^b D(t)\,dt = \lim_{i \to \infty} \int_a^b D_i(t)\,dt.
\]

If, by the way, you have an infinite sequence—say—of honest, continuous, functions that converges uniformly to a limit (which will again be a continuous
function) then that sequence certainly converges—in the above sense—to the same limit when these functions are viewed as generalized functions. BUT, there are many important occasions where your sequence of honest continuous functions doesn’t have that convergence property and yet when they are viewed as generalized functions they do converge to some generalized function as a limit. We will see this soon when we get back to the “sequence of the red graphs.” This sequence does converge (in the above sense) to the Dirac δ-function when these red graphs are thought of as a sequence of generalized functions.

The integral notation for distribution is very useful, and allows us the flexibility to define, for nice enough—and honest—functions \( c(t) \) useful expressions such as

\[
\int_a^b c(t)D(t).
\]

Figure 21.3: The Dirac δ—“function” (actually distribution), where we draw a vertical arrow to illustrate the delta function with support at a given point.

For example, for the Dirac δ-function we have been discussing (i.e., the limit of the red graphs of the previous section) is an honest function away from \( t = 0 \) and—in fact—is the “trivial function” zero away from the origin. And at 0, we may say that it has the “value” infinity, in honor of it being the limit of blip functions getting taller and taller at 0 but the feature that pins it down as a distribution is given by its behavior relative to the second feature above, the area of its graph over the open interval between \( a \) and \( b \):

- If both \( a \) and \( b \) have the same sign (i.e., if the origin is not in the open interval spanned by \( a \) and \( b \)) then the “area under the graph of our Dirac δ-function” is 0.

- If \( a \) is negative and \( b \) is positive then the “area under the graph of our Dirac δ-function” is 1—in notation

\[
\int_a^b \delta = 1.
\]

We sometimes summarize the fact that these areas vanish so long as the origin is not included in the interval we are considering by saying that the support of this δ-function is “at the origin.”
Once you’re happy with this Dirac δ-function, you’ll also be happy with a Dirac δ-function—call it $\delta_x$—with support concentrated at any specific real number $x$ gotten by “translating” the one we’ve been talking about appropriately; $\delta_x$ vanishes for $t \neq x$ and intuitively speaking, has an infinite blip at $t = x$.

So, the original delta-function we were discussing, i.e., $\delta(t)$ would be denoted $\delta_0(t)$.

**A question:** If you’ve never seen distributions before, but know the Riemann integral, can you guess at what the definition of $\int_a^b c(t)D(t)$ is, and can you formulate hypotheses on $c(t)$ that would allow you to endow this expression with a definite meaning?

**A second question:** If you have not seen distributions before, and answered the first question above, let $c(t)$ be an honest function for which your definition of

$$\int_a^b c(t)D(t)$$

applies. Now let $x$ be a real number. Can you use your definition to compute

$$\int_{-\infty}^{+\infty} c(t)\delta_x(t)?$$

The answer, by the way, is: $\int_{-\infty}^{+\infty} c(t)\delta_x(t) = c(x)$. This will be useful in the later sections!

The theory of distributions gives a partial answer to the following funny question:

**How in the world can you “take the derivative” of a function $F(t)$ that doesn’t have a derivative?**

The short answer to this question is that this derivative $F'(t)$ which doesn’t exist as a function may exist as a distribution. What then is the integral of that distribution? Well, it is given by the original function!

$$\int_a^bF'(t)dt = F(b) - F(a).$$

Let us practice this with simple staircase functions. For example, what is the derivative—in the sense of the theory of distributions—of the function in Figure 21.4? Answer $\delta_0 + 2\delta_1$. 
Figure 21.4: The staircase function that is 0 for $t \leq 0$, 1 for $0 < t \leq 1$ and 3 for $1 < t \leq 2$ has derivative $\delta_0 + 2\delta_1$.

We’ll be dealing with much more complicated staircase functions in the next section, but the general principles discussed here will nicely apply there.\[31\]
Chapter 22

Fourier transforms: second visit

In Chapter 16 above we wrote:

The operation that starts with a graph and goes to its spectral picture that records the frequencies, amplitudes, and phases of the pure sine waves that, together, compose the graph is called the Fourier transform.

Now let’s take a closer look at this operation Fourier transform.

We will focus our discussion on an even function $f(t)$ of a real variable $t$. “Even” means that its graph is symmetric about the $y$-axis; that is, $f(-t) = f(t)$. See Figure 22.1.

Figure 22.1: The graph of an even function is symmetrical about the $y$-axis.

When we get to apply this discussion to the staircase of primes $\pi(t)$ or the tinkered staircase of primes $\psi(t)$ both of which being defined only for positive values of $t$, then we would “lose little information” in our quest to understand them if we simply “symmetrized their graphs” by defining their values on negative numbers $-t$ via the formulas $\pi(-t) = \pi(t)$ and $\psi(-t) = \psi(t)$ thereby turning each of them into even functions.
The idea behind the Fourier transform is to express \( f(t) \) as made up out of sine and cosine wave functions. Since we have agreed to consider only even functions, we can dispense with the sine waves—they won’t appear in our Fourier analysis—and ask how to reconstruct \( f(t) \) as a sum (with coefficients) of cosine functions (if only finitely many frequencies occur in the spectrum of our function) or more generally, as an integral if the spectrum is more elaborate. For this work, we need a little machine that tells us, for each real number \( \theta \) whether or not \( \theta \) is in the spectrum of \( f(t) \), and if so, what the amplitude is of the cosine function \( \cos(\theta t) \) that occurs in the Fourier expansion of \( f(t) \)—this amplitude answers the awkwardly phrased question: how much \( \cos(\theta t) \) “occurs in” \( f(t) \)? We will denote this amplitude by \( \hat{f}(\theta) \), and refer to it as the Fourier transform of \( f(t) \). The spectrum, then, of \( f(t) \) is the set of all frequencies \( \theta \) where the amplitude is nonzero.

Now in certain easy circumstances—specifically, if \( \int_{-\infty}^{+\infty} |f(t)| \, dt \) (exists, and) is finite—the Integral Calculus provides us with an easy construction of that machine (see Figure 22.3); namely:

\[
\hat{f}(\theta) = \int_{-\infty}^{+\infty} f(t) \cos(-\theta t) \, dt.
\]

This concise machine manages to “pick out” just the part of \( f(t) \) that has frequency \( \theta \)!! It provides for us the analysis part of the Fourier analysis of our function \( f(t) \).

But there is a synthesis part to our work as well, for we can reconstruct \( f(t) \) from its Fourier transform, by a process intriguingly similar to the analysis part;
namely: if \( \int_{-\infty}^{+\infty} |\hat{f}(\theta)|d\theta \) (exists, and) is finite, we retrieve \( f(t) \) by the integral

\[
f(t) = \frac{1}{2\pi} \int_{-\infty}^{+\infty} \hat{f}(\theta) \cos(\theta t) d\theta.
\]

We are not so lucky to have \( \int_{-\infty}^{+\infty} |f(t)|dt \) finite when we try our hand at a Fourier analysis of the staircase of primes, but we'll work around this!
Chapter 23

What is the Fourier transform of a delta function?

Consider the δ-function that we denoted δ(t) (or δ₀(t)). This is also the “generalized function” that we thought of as the “limit of the red graphs” in Chapter 21 above. Even though δ(t) is a distribution and not a bona fide function, it is symmetric about the origin, and also

$$\int_{-\infty}^{+\infty} |δ(t)| dt$$

exists, and is finite (its value is, in fact, 1). All this means that, appropriately understood, the discussion of the previous section applies, and we can feed this delta-function into our Fourier Transform Machine (Figure 22.3) to see what frequencies and amplitudes arise in our attempt to express—whatever this means!—the delta-function as a sum, or an integral, of cosine functions.

So what is the Fourier transform, ˆδ₀(θ), of the delta-function?

Well, the general formula would give us:

$$\hat{δ}_0(θ) = \int_{-\infty}^{+\infty} \cos(-θt)δ₀(t)dt$$

and as we mentioned in section 18, for any nice function c(t) we have that the integral of the product of c(t) by the distribution δₓ(t) is given by the value of the function c(t) at t = x. SO:

$$\hat{δ}_0(θ) = \int_{-\infty}^{+\infty} \cos(-θt)δ₀(t)dt = \cos(0) = 1.$$
In other words, the Fourier transform of $\delta_0(t)$ is the constant function

$$\hat{\delta}_0(\theta) = 1.$$ 

One can think of this colloquially as saying that the delta-function is a perfect example of white noise in that every frequency occurs in its Fourier analysis and they all occur in equal amounts.

To generalize this computation let us consider for any real number $x$ the symmetrized delta-function with support at $x$ and $-x$, given by

$$d_x(t) = (\delta_x(t) + \delta_{-x}(t))/2$$

![Figure 23.1: The sum $(\delta_x(t) + \delta_{-x}(t))/2$, where we draw vertical arrows to illustrate the Dirac delta functions.](image)

What is the Fourier transform of this $d_x(t)$? The answer is given by making the same computation as we’ve just made:

$$\hat{d}_x(\theta) = \frac{1}{2} \left( \int_{-\infty}^{+\infty} \cos(-\theta t) \delta_x(t) dt + \int_{-\infty}^{+\infty} \cos(-\theta t) \delta_{-x}(t) dt \right)$$

$$= \frac{1}{2} \left( \cos(-\theta x) + \cos(+\theta x) \right)$$

$$= \cos(x\theta)$$

To summarize this in ridiculous (!) colloquial terms: for any frequency $\theta$ the amount of $\cos(\theta t)$ you need to build up the generalized function $(\delta_x(t) + \delta_{-x}(t))/2$ is $\cos(x\theta)$.

So far, so good, but remember that the theory of the Fourier transform has—like much of mathematics—two parts: an analysis part and a synthesis part. We’ve just performed the analysis part of the theory for these symmetrized delta functions $(\delta_x(t) + \delta_{-x}(t))/2$.

Can we synthesize them—i.e., build them up again—from their Fourier transforms?

We’ll leave this, at least for now, as a question for you.
Chapter 24

Trigonometric series

Given our interest in the ideas of Fourier, it is not surprising that we’ll want to deal with things like

\[ F(\theta) = \sum_{k=1}^{\infty} a_k \cos(s_k \cdot \theta) \]

where the \( s_k \) are real numbers tending (strictly monotonically) to infinity. These we’ll just call trigonometric series without asking whether they converge in any sense for all values of \( \theta \), or even for any value of \( \theta \). The \( s_k \)’s that occur in such a trigonometric series we will call the spectral values or for short, the spectrum of the series, and the \( a_k \)’s the (corresponding) amplitudes. We repeat that we impose no convergence requirements at all. But we also think of these things as providing “cutoff” finite trigonometric sums, which we think of as functions of two variables, \( \theta \) and \( C \) (the “cutoff”) where

\[ F(\theta, C) := \sum_{s_k \leq C} a_k \cos(s_k \cdot \theta). \]

These functions \( F(\theta, C) \) are finite trigonometric series and therefore “honest functions” having finite values everywhere.

24.1 Spike values

**Definition 24.1.** Say that a trigonometric series \( F(\theta) \) has a spike at the real number \( \theta = \tau \in \mathbb{R} \) if the set of absolute values \( |F(\tau, C)| \) as \( C \) ranges through positive number cutoffs is unbounded. A real number \( \tau \in \mathbb{R} \) is, in contrast, a non-spike if those values admit a finite upper bound.

One main mission in this book will be furthered just by paying attention to the spike values of certain (infinite) trigonometric series. The trigonometric sums we
study are going to be particularly nice. Under the assumption of the Riemann Hypothesis, they will have the property that the set of their spike values are particularly interesting discrete subsets of the real line.

24.2 Trigonometric Series as Fourier Transforms

Recall, as in Chapter 23, that for any real number $x$, we considered the symmetrized delta-function with support at $x$ and $-x$, given by

$$d_x(t) = \frac{(\delta_x(t) + \delta_{-x}(t))}{2}$$

![Figure 24.1](image-url)

Figure 24.1: The sum $(\delta_x(t) + \delta_{-x}(t))/2$, where we draw vertical arrows to illustrate the Dirac delta functions.

and noted that the Fourier transform of this $d_x(t)$ is

$$\hat{d}_x(\theta) = \cos(x\theta).$$

It follows, of course, that a cutoff finite trigonometric series, $F(\theta, C)$ associated to an infinite trigonometric series

$$F(\theta) = \sum_{k=1}^{\infty} a_k \cos(s_k \cdot \theta)$$

is the Fourier transform of the distribution

$$D(t, C) := \sum_{s_k \leq C} a_k d_{s_k}(t).$$

Given the discreteness of the set of spectral values $s_k (k = 1, 2, \ldots)$ and given the efficacy of the theory of distributions, we can perfectly well consider the infinite sum

$$D(t) := \sum_{k=1}^{\infty} a_k d_{s_k}(t),$$

viewed as distribution, and playing the role of the ‘inverse Fourier transform’ of our trigonometric series $F(t)$. 
Part III

The Riemann Spectrum of the Prime Numbers
Chapter 25

A sneak preview

To get a sense of what we’re in for, let us consider two infinite trigonometric sums—that seem to be related one to the other in that the frequencies of the terms in the one trigonometric sum give the spike values of the other, and vice versa: the frequencies of the other give the spike values of the one: a kind of duality as in the theory of Fourier transforms. We show this duality by exhibiting the graphs of more and more accurate finite approximations (cutoffs) of these infinite sums. More specifically,

1. The first infinite trigonometric sum is a sum of pure cosine waves with frequencies given by logarithms of powers of primes and with amplitudes that will be described below. The graphs of longer and longer truncations of these trigonometric sums, as you will see, have “higher and higher peaks” concentrated more and more accurately at a certain infinite discrete set of real numbers that what we will be referring to as the Riemann spectrum indicated in our pictures below (Figures 25.2–25.5) by the series of vertical red lines.

2. In contrast, the second infinite trigonometric sum is a sum of pure cosine waves with frequencies given by what we have dubbed above the Riemann spectrum and with amplitudes all equal to 1. These graphs will have “higher and higher peaks” concentrated more and more accurately at the logarithms of powers of primes indicated in our pictures below (see Figure 25.6) by the series of vertical blue lines.

That the series of blue lines (i.e., the logarithms of powers of primes) in our pictures below determines—via the the trigonometric sums we describe—the series of red lines (i.e., what we are calling the spectrum) and conversely is a consequence of the Riemann Hypothesis.
1. Getting the Riemann Spectrum as the spike values of a trigonometric series with frequencies equal to (logs of) powers of the primes:

To get warmed up, let’s plot the positive values of the following sum of (co-)sine waves:

$$f(t) = -\frac{\log(2)}{2^{1/2}} \cos(t \log(2)) - \frac{\log(3)}{3^{1/2}} \cos(t \log(3))$$

$$- \frac{\log(2)}{4^{1/2}} \cos(t \log(4)) - \frac{\log(5)}{5^{1/2}} \cos(t \log(5))$$

Figure 25.1: Plot of $f(t)$

Look at the peaks of this graph. There is nothing very impressive about them, you might think; but wait, for this is just a very “early” piece of an expression that consists of a sum of infinitely many (co-)sine waves:

$$- \sum_{p^n} \frac{\log(p)}{p^{n/2}} \cos(t \log(p^n))$$

the summation being over all powers $p^n$ of all prime numbers $p$.

Let us cut this infinite sum off taking only finitely many terms, by choosing various “cutoff values” $C$ and forming the finite sums

$$- \sum_{p^n \leq C} \frac{\log(p)}{p^{n/2}} \cos(t \log(p^n))$$

and plotting their positive values. Figures 25.2-25.5 show what we get for a few values of $C$.

In each of the graphs, we have indicated by red vertical arrows the real numbers that give the values of the Riemann spectrum that we will be discussing. These numbers at the red vertical arrows in the graphs above,

$$\theta_1, \theta_2, \theta_3, \ldots$$

Here we make use of the Greek symbol $\sum$ as a shorthand way of expressing a sum of many terms. We are not requesting this sum to converge.
are *spike values*—as described in Chapter 24—of the infinite trigonometric series

\[- \sum_{p^n < C} \frac{\log(p)}{p^n/2} \cos(t \log(p^n)).\]

They constitute what we are calling the Riemann spectrum and are key to the staircase of primes [32].

- **The sum with** $p^n \leq C = 5$
  
  Here is the function $f(t)$ we displayed above; it consists in the sum of the first four terms of our infinite sum, and doesn’t yet show very much “structure”:

![Figure 25.2](image)

Figure 25.2: Plot of $- \sum_{p^n \leq 5} \frac{\log(p)}{p^n/2} \cos(t \log(p^n))$ with arrows pointing to the spectrum of the primes

- **The sum with** $p^n \leq C = 20$
  
  Something, (don’t you agree?) is already beginning to happen here:

![Figure 25.3](image)

Figure 25.3: Plot of $- \sum_{p^n \leq 20} \frac{\log(p)}{p^n/2} \cos(t \log(p^n))$ with arrows pointing to the spectrum of the primes

- **The sum with** $p^n \leq C = 50$
  
  Note that the high peaks seem to be lining up more accurately with the vertical red lines. Note also that the $y$-axis has been rescaled.
Figure 25.4: Plot of $- \sum_{p^m \leq 50} \frac{\log(p)}{p^n/2} \cos(t \log(p^n))$ with arrows pointing to the spectrum of the primes

- **The sum with** $p^n \leq C = 500$

Here, the peaks are even sharper, and note that again they are higher; that is, we have rescaled the $y$-axis.

Figure 25.5: Plot of $- \sum_{p^m \leq 500} \frac{\log(p)}{p^n/2} \cos(t \log(p^n))$ with arrows pointing to the spectrum of the primes

We will pay attention to:

- how this “plays out” as we take the sums of longer and longer pieces of the infinite sum of cosine waves above, given by larger and larger cutoffs $C$,
- how this spectrum of red lines more closely matches the high peaks of the graphs of the positive values of these finite sums,
- how these peaks are climbing higher and higher,
- what relationship this has to the Fourier analysis of the staircase of primes,
- and, equally importantly, what these mysterious red lines signify.

2. **Towards (logs of) powers of the primes, starting from the Riemann Spectrum:**
Here we will be making use of the series of numbers
\[ \theta_1, \theta_2, \theta_3, \ldots \]
comprising what we called the \textit{spectrum}. We consider the infinite trigonometric series
\[ 1 + \cos(\theta_1 t) + \cos(\theta_2 t) + \cos(\theta_3 t) + \ldots \]
or, using the \( \sum \) notation,
\[ 1 + \sum_{\theta} \cos(\theta t) \]
where the summation is over the spectrum, \( \theta = \theta_1, \theta_2, \theta_3, \ldots \). Again we will consider finite cutoffs \( C \) of this infinite trigonometric sum,
\[ \hat{\Phi}_{\leq C}(\theta) = 1 + \sum_{\theta \leq C} \cos(\theta t) \]
and plot their graphs, over various ranges.

![Plot of \( \hat{\Phi}_{\leq C}(\theta) \) for different \( C \).](image)

Figure 25.6: Plots of \( \hat{\Phi}_{\leq C}(\theta) \) for \( C = 5 \) (top), 20, 50, and 500 (bottom).

This passage—thanks to the Riemann Hypothesis—from spectrum to prime powers and back again via consideration of the “high peaks” in the graphs of the appropriate trigonometric sums provides a kind of visual duality emphasizing,
for us, that the information inherent in the wild spacing of prime powers, is somehow “packaged” in the series of mysterious numbers we have called the Riemann Spectrum, and reciprocally, the information given in that series of mysterious numbers is obtainable from the sequence of prime powers.
Chapter 26

On losing no information

To manage to repackage the “same” data in various ways—where each way brings out some features that would be kept in the shadows if the data were packaged in some different way—is a high art, in mathematics. In a sense every mathematical equation does this, for the “equal sign” in the middle of the equation tells us that even though the two sides of the equation may seem different, or have different shapes, they are nonetheless “the same data.” For example, the equation

$$\log(XY) = \log(X) + \log(Y)$$

which we encountered earlier in Chapter 10 is just two ways of looking at the same thing, yet it was the basis for much manual calculation for several centuries.

Now, the problem we have been concentrating on, in this book, has been—in effect—to understand the pattern, if we can call it that, given by the placement of prime numbers among the natural line-up of all whole numbers.

Figure 26.1: Prime Numbers up to 37

There are, of course, many ways for us to present this basic pattern. Our initial strategy was to focus attention on the staircase of primes which gives us a vivid portrait, if you wish, of the order of appearance of primes among all numbers.
As we have already hinted in the previous sections, however, there are various ways open to us to tinker with—and significantly modify—our staircase \emph{without losing the essential information it contains}. Of course, there is always the danger of modifying things in such a way that “retrieval” of the original data becomes difficult. Moreover, we had better remember every change we have made if we are to have any hope of retrieving the original data!

With this in mind, let us go back to Chapter 14 (discussing the staircase of primes) and Chapter 15, where we tinkered with the original staircase of primes—alias: the graph of \( \pi(X) \)—to get \( \psi(X) \) whose risers look—from afar—as if they approximated the 45 degree staircase.

At this point we’ll do some further carpentry on \( \psi(X) \) \emph{without destroying the valuable information it contains}. We will be replacing \( \psi(X) \) by a generalized function, i.e., a distribution, which we denote \( \Phi(t) \) that has support at all positive integral multiples of logs of prime numbers, and is zero on the complement of that discrete set. Recall that by definition, a discrete subset \( S \) of real numbers is the \emph{support} of a function, or of a distribution, if the function vanishes on the complement of \( S \) and doesn’t vanish on the complement of any proper subset of \( S \).

Given the mission of our book, it may be less important for us to elaborate on the construction of \( \Phi(t) \) than it is to (a) note that \( \Phi(t) \) contains all the valuable information that \( \psi(X) \) has and (b) pay close attention to the spike values of the trigonometric series that is the Fourier transform of \( \Phi(t) \).

For the definition of the distribution \( \Phi(t) \) see the end-note [33].

A distribution that has a discrete set of real numbers as its support—as \( \Phi(t) \) does, we sometimes like to call \textbf{spike distributions} since the pictures of functions approximating them tend to look like a series of spikes.

We have then before us a spike distribution with support at integral multiples of logarithms of prime numbers, and this generalized function retains the essential information about the placement of prime numbers among all whole numbers, and will be playing a major role in our story: knowledge of the placement of the “blips” constituting this distribution (its support), being at integral multiples of logs of prime numbers, would allow us to reconstruct the position of the prime numbers among all numbers. Of course there are many other ways to package this vital information, so we must explain our motivation for subjecting our
poor initial staircase to the particular series of brutal acts of distortion that we described, that end up with the distribution $\Phi(t)$.
Chapter 27

Going from the primes to the Riemann Spectrum

We’ve discussed the nature of the Fourier transform of (symmetrized) \( \delta \)-functions in Chapter 23. In particular, recall the “spike function"

\[
d_x(t) = (\delta_x(t) + \delta_{-x}(t))/2
\]

that has support at the points \( x \) and \( -x \). We mentioned that its Fourier transform, \( \hat{d}_x(\theta) \), is equal to \( \cos(x\theta) \) (and gave some hints about why this may be true).

Our next goal is to work with the much more interesting “spike function”

\[
\Phi(t) = e^{-t/2}\Psi'(t),
\]

which was one of the generalized function that we engineered in Chapter 26 and that has support at all nonnegative integral multiples of logarithms of prime numbers.

As with any function—or generalized function—defined for non-negative values of \( t \), we can “symmetrize it” (about the \( t \)-axis) which means that we can define it on negative real numbers by the equation

\[
\Phi(-t) = \Phi(t).
\]

Let us make that convention, thereby turning \( \Phi(t) \) into an even generalized function.
Figure 27.1: $\Phi(t)$ is a sum of Dirac delta functions at the logarithms of prime powers $p^n$ weighted by $p^{-n/2}\log(p)$ (and $\log(2\pi)$ at 0). The more blue the arrow, the smaller the weight.

We may want to think of $\Phi(t)$ as a limit of the following sequence of distributions,

$$\Phi(t) = \lim_{C \to \infty} \Phi_{\leq C}(t)$$

where $\Phi_{\leq C}(t)$ is the following finite linear combinations of (symmetrized) $\delta$-functions $d_x(t)$:

$$\Phi_{\leq C}(t) := 2 \sum_{\text{prime powers } p^n \leq C} p^{-n/2} \log(p) d_{n \log(p)}(t).$$

Since the Fourier transform of $d_x(t)$ is $\cos(t\theta)$, the Fourier transform of each $d_{n \log(p)}(t)$ is $\cos(n \log(p)\theta)$, so the Fourier transform of $\Phi_{\leq C}(t)$ is

$$\hat{\Phi}_{\leq C}(\theta) := 2 \sum_{\text{prime powers } p^n < C} p^{-n/2} \log(p) \cos(n \log(p)\theta).$$

So, following the discussion in Chapter 24 above, we are dealing with the cutoffs at finite points $C$ of the trigonometric series

$$\hat{\Phi}(\theta) := 2 \sum_{\text{prime powers } p^n} p^{-n/2} \log(p) \cos(n \log(p)\theta).$$

(This trigonometric series—whose spectral values are the logarithms of prime powers—may also be written as

$$\sum_{m=2}^{\infty} \Lambda(m) m^{-s} + \sum_{m=2}^{\infty} \Lambda(m) m^{-\bar{s}}$$

for $s = \frac{1}{2} + i\theta$, where $\Lambda(m)$ is the von-Mangoldt function.)

Here is a graph of various cutoffs of this trigonometric function (times $-1/2$).
William: (FIGURE 1) I know that you have just computed lots of new graphs but can’t log onto the cloud (it doesn’t know me or my email address) so I can’t yet suggest exactly which of the graphs to put into these figure spots, but how about putting the plot of \( \hat{\Phi}_{\leq C}(\theta) \) (times \(-1/2\)) here as a function of \( \theta \) for a few cutoffs \( C \)? For example: “fig:phic.”

Notice the risers at the red vertical lines. We’ve superimposed the graph of derivative of \( \Phi_{\leq C}(\theta) \) to focus attention more specifically on those risers.

To see this staircase phenomenon of the graph even more clearly, we will counteract the ‘ascendancy’ of the graph without destroying its staircase structure, by subtracting \( R(C, \theta) \), a certain smooth function \([34]\) of \( \theta \) and \( C \), giving us the function we call:

\[
A_C(\theta) := \hat{\Phi}_{\leq C}(\theta) + R(C, \theta)
\]

whose graph looks like:

William: (FIGURE 2) Plots of \( A_C(\theta) := \hat{\Phi}_{\leq C}(\theta) + R(C, \theta) \). Now \( R(C, \theta) \) is, I think, \(-1/2\) times the function \( T_C(\theta) := C^{1/2} \left\{ \cos(\theta \cdot \log(C)) + 2\theta \sin(\theta \cdot \log(C)) \right\} \)

of our previous draft. But we should recalculate, to make sure. Maybe again the same cutoffs \( C = 5 \) (top), 20, 50, and 500 as in NEW FIGURE 1.

To get an even picture, we can smooth out \( A_C(\theta) \) by averaging over absolutely all cutoffs ranging between the extreme cutoff at 0 and the cutoff at \( C \), giving us a function we’ll denote \( \tilde{A}_C(\theta) \). (This operation is called Césaro summability.) Its graph looks like:

William: (FIGURE 3) ) Plots of \( \tilde{A}_C(\theta) \). Maybe the same cutoffs \( C \) as in NEW FIGURE 1.

Finally, we can make the spikes look more uniform in height (for different cutoffs) by dividing by \( \log C \), giving us a function we’ll denote \( B_C(\theta) := \tilde{A}_C(\theta)/\log C \).

We have come then—again—to the expressions graphed at the beginning of Part [III] in Chapter 25.
William: (FIGURE 4) Plots of \( B_C(\theta) := \tilde{A}_C(\theta)/\log C \). Maybe the same cutoffs \( C \) as in NEW FIGURE 1.

For a theoretical discussion of this, see endnote [35].

The \( \theta \)-coordinates of the peaks of these graphs seem to be vaguely clustered about a discrete set of positive real numbers. In fact, this is already a hint of what happens to the graphs of these functions \( \frac{1}{C} \Phi_{\leq C}(\theta) \) as we allow the cut-off \( C \) to tend to infinity. These ‘spikes’ are our first glimpse of a certain infinite set of positive real numbers

\[ \theta_1, \theta_2, \theta_3, \ldots \]

which constitute the Riemann spectrum of the primes. If the Riemann Hypothesis holds, these numbers would be the key to the placement of primes on the number line.

By tabulating these peaks we can compute—at least very approximately—

\[ \begin{align*}
\theta_1 &= 14.134725 \ldots \\
\theta_2 &= 21.022039 \ldots \\
\theta_3 &= 25.010857 \ldots \\
\theta_4 &= 30.424876 \ldots \\
\theta_5 &= 32.935061 \ldots \\
\theta_6 &= 37.586178 \ldots 
\end{align*} \]

Riemann defined this sequence of numbers in his 1859 article in a manner somewhat different from the treatment we have given (in that article these \( \theta_i \) appear as the “imaginary parts of the nontrivial zeroes of his \( \zeta \)-function;” we will discuss this briefly in Chapter [30] below). Riemann wrote:

“One now finds indeed approximately this number of real roots within these limits, and it is very probable that all roots are real. Certainly one would wish for a stricter proof here; I have meanwhile temporarily put aside the search for this after some futile attempts, as it appears unnecessary for the next objective of my investigation.”

Nowadays, these mysterious numbers, these spectral lines for the staircase of primes, are known to great accuracy. Here is the smallest one, \( \theta_1 \), given with over 1,000 digits of its decimal expansion:

14.134725141734693790457251983562470270784257115699243175685567460149
9634298092567649490103931715610127792029715487974367661426914698822545
825053632394471377804133812372059705496219558658602005556672583601077
370205410982661507542780517442591306254481978651072304938725629738321
and if, by any chance, you wish to peruse the first 2,001,052 of these $\theta_i$'s calculated to an accuracy within $3 \cdot 10^{-9}$, consult Andrew Odlyzko's tables:

http://www.dtc.umn.edu/~odlyzko/zeta_tables

Since people have already computed at least 1,029.9 billion $\theta$'s (as of Feb. 18, 2005) and have never found one with multiplicity $> 1$; it is generally expected that the multiplicity of all the $\theta$'s in the Riemann Spectrum is 1. But, independent of that expectation, our convention in what follows will be that we count each of the elements in the Riemann Spectrum repeated as many times as their multiplicity. So, if it so happens that $\theta_n$ occurs with multiplicity two, we view the Riemann spectrum as being the series of numbers

$$\theta_1, \theta_2, \ldots, \theta_{n-1}, \theta_n, \theta_n, \theta_{n+1}, \ldots$$
Chapter 28

Going from the Riemann Spectrum to the primes

To justify the name *Riemann spectrum of primes* we will investigate graphically whether in an analogous manner we can use this spectrum to get information about the placement of prime numbers. We might ask, for example, if there is a trigonometric function with frequencies given by this collection of real numbers, \( \theta_1, \theta_2, \theta_3, \ldots \) that somehow pinpoints the prime powers, just as our functions

\[
\hat{\Phi}(\theta)_{\leq C} = 2 \sum_{\text{prime} \, \text{powers} \, p^n \leq C} p^{-m/2} \log(p) \cos(n \log(p) \theta)
\]

for large \( C \) pinpoint the spectrum (as discussed in the previous two chapters).

To start the return game, consider this sequence of trigonometric functions that have (zero and) the \( \theta_i \) as spectrum

\[
G_C(x) := 1 + \sum_{i < C} \cos(\theta_i \cdot x).
\]

As we’ll see presently it is best to view these functions on a logarithmic scale so we will make the substitution of variables \( x = \log(s) \) and write

\[
H_C(s) := G_C(\log(s)) = 1 + \sum_{i < C} \cos(\theta_i \cdot \log(s)).
\]

Now take a look at the graphs of our functions \( H_C(s) \) for various choices of \( C \) and ranges of \( s \).
To get a cleaner picture let us subject our functions $H_C(s)$ to Cesaro smoothing as in the previous chapter. Define:

$$\tilde{H}_C(s) := \frac{1}{C} \sum_{c=1}^{C} H_c(\theta),$$

and let’s take another look.

Figure 28.1: Illustration of $-\sum_{i=1}^{1000} \cos(\log(s)\theta_i)$, where $\theta_1 \sim 14.13, \ldots$ are the first 1000 contributions to the Riemann spectrum. The red dots are at the prime powers $p^n$, whose size is proportional to $\log(p)$.

Figure 28.2: Illustration of $-\sum_{i=1}^{1000} \cos(\log(s)\theta_i)$ in the neighborhood of a twin prime. Notice how the two primes 29 and 31 are separated out by the Fourier series, and how the prime powers $3^3$ and $2^5$ also appear.
Figure 28.3: Fourier series from 1,000 to 1,030 using 15,000 of the numbers $\theta_i$. Note the twin primes 1019 and 1021 and that $1024 = 2^{10}$. 
Part IV

Back to Riemann
Chapter 29

How to build $\pi(X)$ knowing the Spectrum (Riemann’s way)

We have been dealing in Part III of our book with $\Phi(t)$ a distribution that—we said—contains all the essential information about the placement of primes among numbers. We have given a clean restatement of Riemann’s hypothesis, the third restatement so far, in term of this $\Phi(t)$. But $\Phi(t)$ was the effect of a series of recalibrations and reconfigurings of the original untampered-with staircase of primes. A test of whether we have strayed from our original problem—to understand this staircase—would be whether we can return to the original staircase, and “reconstruct it” so to speak, solely from the information of $\Phi(t)$—or equivalently, assuming the Riemann Hypothesis as formulated in Chapter 15—can we construct the staircase of primes $\pi(X)$ solely from knowledge of the sequence of real numbers $\theta_1, \theta_2, \theta_3, \ldots$?

The answer to this is yes (given the Riemann Hypothesis), and is discussed very beautifully by Bernhard Riemann himself in his famous 1859 article cited above.

Bernhard Riemann used the spectrum of the prime numbers to provide an exact analytic formula that analyzes and/or synthesizes the staircase of primes. This formula is motivated by Fourier’s analysis of functions as constituted out of sines. Riemann started with a specific smooth function, which we will refer to as $R(X)$, a function that Riemann offered, just as Gauss offered his $\text{Li}(X)$, as a candidate smooth function approximating the staircase of primes. Recall from Chapter 12 that Gauss’s guess is $\text{Li}(X) = \int_{\frac{1}{2}}^{X} dt/\log(t)$. Riemann’s guess for a better approximation to $\pi(X)$ is obtained from Gauss’s using the Moebius
function \( \mu(n) \), which is defined by

\[
\mu(n) = \begin{cases} 
1 & \text{if } n \text{ is a square-free positive integer with an even number of distinct prime factors} \\
-1 & \text{if } n \text{ is a square-free positive integer with an odd number of distinct prime factors} \\
0 & \text{if } n \text{ is not square-free.}
\end{cases}
\]

See Figure 29.1 for a plot of the Moebius function.

Figure 29.1: The blue dots plot the values of the Moebius function \( \mu(n) \), which is only defined at integers.

Riemann’s guess is

\[
R(X) = \sum_{n=1}^{\infty} \frac{\mu(n)}{n} \text{Li}(X^{\frac{1}{n}}),
\]

where \( \mu(n) \) is the Moebius function introduced above.

In Chapter 13 we encountered the Prime Number Theorem, which asserts that \( X/\log(X) \) and \( \text{Li}(X) \) are both approximation for \( \pi(X) \), in the sense that both go to infinity at the same rate. Our first formulation of the Riemann Hypothesis (see page 44) was that \( \text{Li}(X) \) is an essentially square root accurate approximation of \( \pi(X) \). Figures 29.3–29.4 illustrate that Riemann’s function \( R(X) \) appears to be an even better approximation to \( \pi(X) \) than anything we have seen before.
Think of Riemann’s smooth curve $R(X)$ as the *fundamental* approximation to $\pi(X)$. Riemann offered much more than just a (conjecturally) better approximation to $\pi(X)$ in his wonderful 1859 article. He found a way to construct what looks like a Fourier series, but with $\sin(X)$ replaced by $R(X)$ and spectrum the $\theta_i$, which conjecturally exactly equals $\pi(X)$. He gave an infinite sequence of improved guesses,

$$R(X) = R_0(X), \quad R_1(X), \quad R_2(X), \quad R_3(X), \quad \ldots$$

and he hypothesized that one and all of them were all essentially square root approximations to $\pi(X)$, and that the sequence of these better and better approximations converge to give an exact formula for $\pi(X)$.

Thus not only did Riemann provide a “fundamental” (that is, a smooth curve that is an astoundingly close to $\pi(X)$) but he viewed this as just a starting point, for he gave the recipe for providing an infinite sequence of corrective terms—call them Riemann’s *harmonics*; we will denote the first of these “harmonics” $C_1(X)$, the second $C_2(X)$, etc. Riemann gets his first corrected curve, $R_1(X)$, from $R(X)$ by adding this first harmonic to the fundamental,

$$R_1(X) = R(X) + C_1(X),$$

he gets the second by correcting $R_1(X)$ by adding the second harmonic

$$R_2(X) = R_1(X) + C_2(X),$$
and so on

\[ R_3(X) = R_2(X) + C_3(X), \]

and in the limit provides us with an exact fit.

Figure 29.5: Riemann analytic formula for \( \pi(X) \).

The Riemann Hypothesis, if true, would tell us that these correction terms \( C_1(X), C_2(X), C_3(X), \ldots \) are all square-root small, and all the successively corrected smooth curves

\[ R(X), R_1(X), R_2(X), R_3(X), \ldots \]

are good approximations to \( \pi(X) \). Moreover,

\[ \pi(X) = R(X) + \sum_{k=1}^{\infty} C_k(X). \]

The elegance of Riemann’s treatment of this problem is that the corrective terms \( C_k(X) \) are all modelled on the fundamental \( R(X) \) and are completely described if you know the sequence of real numbers \( \theta_1, \theta_2, \theta_3, \ldots \) of the last section.

Here in the book will be where we refer explicitly to complex numbers for the first time!! The definition of Riemann’s \( C_k(X) \) requires complex numbers \( a + bi \), where \( i = \sqrt{-1} \), and requires extending the definition of the function \( \text{Li}(X) \) to make sense when given complex numbers as input. Assuming the Riemann Hypothesis, the Riemann correction terms \( C_k(X) \) are then

\[ C_k(X) = -R(X^{\frac{1}{2} + i\theta_k}), \]

where \( \theta_1 = 14.134725 \ldots, \theta_2 = 21.022039 \ldots, \text{etc.} \), are the spectrum of the prime numbers [37].

Riemann provided an extraordinary recipe that allows us to work out the harmonics,

\[ C_1(X), \ C_2(X), \ C_3(X), \ldots \]

without our having to consult, or compute with, the actual staircase of primes. As with Fourier’s modus operandi where both fundamental and all harmonics are modeled on the sine wave, but appropriately calibrated, Riemann fashioned his higher harmonics, modeling them all on a single function, namely his initial guess \( R(X) \).
The convergence of $R_k(X)$ to $\pi(X)$ is strikingly illustrated in the plots in Figures 29.6–29.11 of $R_k$ for various values of $k$.

Figure 29.6: The function $R_1$ approximating the staircase of primes up to 100

Figure 29.7: The function $R_{10}$ approximating the staircase of primes up to 100
Figure 29.8: The function $R_{25}$ approximating the staircase of primes up to 100

Figure 29.9: The function $R_{50}$ approximating the staircase of primes up to 100

Figure 29.10: The function $R_{50}$ approximating the staircase of primes up to 500
Figure 29.11: The function $\text{Li}(X)$ (top, green), the function $R_{50}(X)$ (in blue), and the staircase of primes on the interval from 350 to 400.
Chapter 30

As Riemann envisioned it, complex analysis relates the staircase of primes to its Riemann Spectrum

In the previous chapters we have described—using Riemann’s Hypothesis—how to obtain the spectrum

\[ \theta_1, \theta_2, \theta_3, \ldots \]

from the staircase of primes, and hinted at how to go back. Roughly speaking, we were performing “Fourier transformations” to make this transit. But Riemann, on the very first page of his 1859 memoir, construes the relationship we have been discussing, between spectrum and staircase, in an even more profound way.

To talk about this extraordinary insight of Riemann, one would need to do two things that might seem remote from our topic, given our discussion so far.

- We will discuss a key idea that Leonhard Euler had (circa 1740).
- To follow the evolution of this idea in the hands of Riemann, we would have to assume familiarity with basic complex analysis.

We will say only a few words here about this, in hopes of giving at least a shred of a hint of how marvelous Riemann’s idea is. We will be drawing, at this point, on some further mathematical background. For readers who wish to pursue the themes we discuss, here is a list of sources, that are our favorites among those meant to be read by a somewhat broader audience than people very advanced in the subject. We list them in order of “required background.”
1. John Derbyshire’s *Prime Obsession: Bernhard Riemann and the Greatest Unsolved Problem in Mathematics*. We have already mentioned this book in our introduction, but feel that it is so good, that it is worth a second mention here.

2. The Wikipedia entry for Riemann’s Zeta Function ([http://en.wikipedia.org/wiki/Riemann_zeta_function](http://en.wikipedia.org/wiki/Riemann_zeta_function)) It is difficult to summarize who wrote this, but we feel that it is a gift to the community in its clarity. Thanks authors!

3. Enrico Bombieri’s article [?]. To comprehend all ten pages of this excellent and fairly thorough account may require significant background, but try your hand at it; no matter where you stop, you will have seen good things in what you have read.

**Leonhardt Euler’s idea (≈1740):**

As readers of Jacob Bernoulli’s *Ars Conjectandi* (or of the works of John Wallis) know, there was already a rich mathematical theory of finite sums of (non-negative $k$-th powers) of consecutive integers. This sum,

$$S_k(N) = 1^k + 2^k + 3^k + \cdots + N^k,$$

is a polynomial in $N$ of degree $k+1$ with no constant term, a leading term equal to $\frac{1}{k+1} N^{k+1}$, and a famous linear term. The coefficient of the linear term of the polynomial $S_k(N)$ is the *Bernoulli number* $B_k$:

$$S_1(n) = 1 + 2 + 3 + \cdots + (n-1) = \frac{n(n-1)}{2} = \frac{n^2}{2} - \frac{1}{2} \cdot n,$$

$$S_2(n) = 1^2 + 2^2 + 3^2 + \cdots + (n-1)^2 = \frac{n^3}{3} + \cdots - \frac{1}{6} \cdot n,$$

$$S_3(n) = 1^3 + 2^3 + 3^3 + \cdots + (n-1)^3 = \frac{n^4}{4} + \cdots - 0 \cdot n,$$

$$S_4(n) = 1^4 + 2^4 + 3^4 + \cdots + (n-1)^4 = \frac{n^5}{5} + \cdots - \frac{1}{30} \cdot n,$$

etc. For odd integers $k > 1$ this linear term vanishes. For even integers $2k$ the Bernoulli number $B_{2k}$ is the rational number given by the coefficient of $\frac{x^{2k}}{2k!}$ in the power series expansion

$$\frac{x}{e^x - 1} = 1 - \frac{x}{2} + \sum_{k=1}^{\infty} \frac{(-1)^{k+1} B_{2k}}{2k!} x^{2k}.$$
So

\[ B_2 = \frac{1}{6}, \quad B_4 = \frac{1}{30}, \quad B_6 = \frac{1}{42}, \quad B_8 = \frac{1}{30}, \]

and to convince you that the numerator of these numbers is not always 1, here are a few more:

\[ B_{10} = \frac{5}{66}, \quad B_{12} = \frac{691}{2730}, \quad B_{14} = \frac{7}{6}. \]

If you turn attention to sums of negative \( k \)-th powers of consecutive integers, then when \( k = -1 \),

\[ S_{-1}(N) = 1 + \frac{1}{2} + \frac{1}{3} + \cdots + \frac{1}{N} \]

tends to infinity like \( \log(N) \), but for \( k < -1 \) we are facing the sum of reciprocals of powers (of exponent > 1) of consecutive whole numbers, and \( S_k(N) \) converges. This is the first appearance of the zeta function \( \zeta(s) \) for arguments \( s = 2, 3, 4, \ldots \)

So let us denote these limits by notation that has been standard, after Riemann:

\[ \zeta(s) := 1 + \frac{1}{2^s} + \frac{1}{3^s} + \cdots \]

The striking reformulation that Euler discovered was the expression of this infinite sum as an infinite product of factors associated to the prime numbers:

\[ \zeta(s) = \prod_{p \text{ prime}} \frac{1}{1 - \frac{1}{p^s}} \]

where the infinite sum on the left and the infinite product on the right both converge (and are equal) if \( s > 1 \). He also evaluated these sums at even positive integers, where—surprise—the Bernoulli numbers come in again; and they and \( \pi \) play a combined, fascinating role:

\[ \zeta(2) = \frac{1}{1^2} + \frac{1}{2^2} + \cdots = \pi^2/6 \simeq 1.65 \ldots \]

\[ \zeta(4) = \frac{1}{1^4} + \frac{1}{2^4} + \cdots = \pi^4/90 \simeq 1.0823 \ldots \]

and, in general,

\[ \zeta(2n) = \frac{1}{1^{2n}} + \frac{1}{2^{2n}} + \cdots = (-1)^{n+1} B_{2n} \pi^{2n} \cdot \frac{2^{2n-1}}{2n!}. \]

A side note to Euler’s formulas comes from the fact (only known much later) that no power of \( \pi \) is rational: do you see how to use this to give a proof that
there are infinitely many primes, combining Euler’s infinite product expansion displayed above with the formula for $\zeta(2)$, or with the formula for $\zeta(4)$, or, in fact, for the formulas for $\zeta(2n)$ for any $n$ you choose?

Pafnuty Lvovich Chebyshev’s (±1845): The second moment in the history of evolution of this function $\zeta(s)$ is when Chebyshev used the same formula as above in the extended range where $s$ is allowed now to be a real variable—not just an integer—greater than 1. Making use of this extension of the range of definition of Euler’s sum of reciprocals of powers of consecutive whole numbers, Chebyshev could prove that for large $x$ the ratio of $\pi(x)$ and $x/\log(x)$ is bounded above and below by two explicitly given constants. He also proved that there exists a prime number in the interval bounded by $n$ and $2n$ for any positive integer $n$ (this was called Bertrand’s postulate).

Riemann’s idea (1859): It is in the third step of the evolution of $\zeta(s)$ that something quite surprising happens. Riemann extended the range of Chebyshev’s sum of reciprocals of positive real powers of consecutive whole numbers allowing the argument $s$ to range over the entire complex plane $s$ (avoiding $s = 1$). Now this is a more mysterious extension of Euler’s function, and it is deeper in two ways:

- The formula
  \[
  \zeta(s) := \frac{1}{1^s} + \frac{1}{2^s} + \frac{1}{3^s} + \cdots
  \]
  does converge when the real part of the exponent $s$ is greater than 1 (i.e., this allows us to use the same formula, as Chebyshev had done, for the right upper half plane in the complex plane determined by the condition $s = x + iy$ with $x > 1$ but not beyond this). You can’t simply use the same formula for the extension.

- So you must face the fact that if you wish to “extend” a function beyond the natural range in which its defining formula makes sense, there may be many ways of doing it.

To appreciate the second point, the theory of a complex variable is essential. The uniqueness (but not yet the existence) of Riemann’s extension of $\zeta(s)$ to the entire complex plane is guaranteed by the phenomenon referred to as analytic continuation. If you are given a function on any infinite subset $X$ of the complex plane that contains a limit point, and if you are looking for a function on the entire complex plane that is differentiable in the sense of complex analysis, there may be no functions at all that have that property, but if there is one, that function is unique. But Riemann succeeded: he was indeed able to extend Euler’s function to the entire complex plane except for the point $s = 1$, thereby defining what we now call Riemann’s zeta function.

1 or to any connected open subset that contains $X$
Those ubiquitous Bernoulli numbers, by the way, reappear yet again as values of this extended zeta function at negative integers:

\[ \zeta(-n) = -\frac{B_{n+1}}{n+1} \]

so since the Bernoulli numbers indexed by odd integers \( > 1 \) all vanish, the extended zeta function \( \zeta(s) \) actually vanishes at all even integers.

The even integers \(-2, -4, -6, \ldots\) are often called the **trivial zeroes** of the Riemann zeta function. There are indeed other zeroes of the zeta function, and those other zeroes could—in no way—be dubbed “trivial,” as we shall shortly see.

It is time to consider these facts:

1. **Riemann’s zeta function codes the placement of prime powers among all numbers** The key here is to take logarithm and then the derivative of \( \zeta(s) \) (this boils down to forming \( \frac{d\zeta}{ds}(s)/\zeta(s) \)). Assuming that the real part of \( s \) is \( > 1 \), taking the logarithm of \( \zeta(s) \)—using Euler’s infinite product formulation—gives us

\[ \log \zeta(s) = \sum_{p \text{ prime}} -\log (1 - p^{-s}), \]

and we can do this term-by-term, since the real part of \( s \) is \( > 1 \). Then taking the derivative gives us:

\[ \frac{d\zeta}{ds}(s)/\zeta(s) = -\sum_{n=1}^{\infty} \Lambda(n)n^{-s} \]

where

\[ \Lambda(n) := \log p \]

if \( n = p^k \) for \( p \) a prime number and \( k > 0 \); and

\[ \Lambda(n) := 0 \]

if \( n \) is not a power of a prime number. In particular, \( \Lambda(n) \) “records” the placement of prime powers.

2. **You know lots about a function if you know its zeroes and poles** This is certainly true, for example for polynomials, or even rational functions: if someone told you, for example, that a certain function \( f(s) \) vanishes to order 1 at 0 and at \( \infty \); and that it has a double pole at \( s = 2 \) and
at all other points has finite nonzero values, then you can immediately say
that this mystery function is a nonzero constant times $s/(s - 2)^2$.

Knowing the zeroes and poles (in the complex plane) alone of the Riemann
zeta function doesn’t entirely pin it down—you have to know more about
its behavior at infinity since—for example, multiplying a function by $e^z$
doesn’t change the structure of its zeroes and poles in the finite plane.
but a complete understanding of the zeroes and poles of $\zeta(s)$ will give all
the information you need to pin down the placement of primes among all
numbers.

So here is the score:

- As for poles, $\zeta(s)$ has only one pole. It is at $s = 1$ and is of order 1
  (a “simple pole”).
- As for zeroes, we have already mentioned the trivial zeroes (at negative
even integers), but $\zeta(s)$ also has infinitely many nontrivial zeroes.
  These nontrivial zeroes are known to lie in the vertical strip

$$0 < \text{real part of } s < 1.$$  

And here is yet another equivalent statement of Riemann’s Hypothesis—this
being the formulation closest to the one given in his 1859 memoir:

**The Riemann Hypothesis (fourth formulation)**

All the nontrivial zeroes of $\zeta(s)$ lie on the vertical line in the complex
plane given by real part of $s = \frac{1}{2}$, and these zeroes are none other than
$\frac{1}{2} + i\theta_1, \frac{1}{2} + i\theta_2, \frac{1}{2} + i\theta_3, \ldots$, where $\theta_1, \theta_2, \theta_3, \ldots$ comprise the spectrum
of primes we talked about in the earlier chapters.

The zeta function, then, is the vise, that so elegantly clamps together information
about the placement of primes and their spectrum!

That a simple geometric property of these zeroes (lying on a line!) is directly
equivalent to such profound (and more difficult to express) regularities among
prime numbers suggests that these zeroes and the parade of Riemann’s corrections
governed by them—when we truly comprehend their message—may have
lots more to teach us, may eventually allow us a more powerful understanding
of arithmetic. This infinite collection of complex numbers, i.e., the nontrivial
zeroes of the Riemann zeta function, plays a role with respect to $\pi(X)$ rather
like the role the spectrum of the Hydrogen atom, plays in Fourier’s theory. Are
the primes themselves no more than an epiphenomenon, behind which there
lies, still veiled from us—a yet-to-be-discovered, yet-to-be-hypothesized, profound
conceptual key to their perplexing orneriness. Are the many innocently posed,
yet unanswered, phenomenological questions about numbers—such as in the ones
listed earlier– waiting for our discovery of this deeper level of arithmetic? Or for layers deeper still? Are we, in fact, just at the beginning?

These are not completely idle thoughts, for a tantalizing analogy relates the number theory we have been discussing to an already established branch of mathematics–due, largely, to the work of Alexander Grothendieck, and Pierre Deligne–where the corresponding analogue of Riemann’s hypothesis has indeed been proved...
Chapter 31

Companions to the zeta function

Our book, so far, has been exclusively about Riemann’s $\zeta(s)$ and its zeroes. We have been discussing how the (placement of) the zeroes of $\zeta(s)$ in the complex plane contains the information needed to understand the (placement of) the primes in the set of all whole numbers; and conversely.

It would be wrong—we think—if we don’t even mention that $\zeta(s)$ fits into a broad family of similar functions that connect to other problems in number theory.

For example—instead of the ordinary integers—consider the Gaussian integers. This is the collection of numbers 

$$\{a + bi\}$$

where $i = \sqrt{-1}$ and $a, b$ are ordinary integers. We can add and multiply two such numbers and get another of the same form. The only “units” among the Gaussian integers (i.e., numbers whose inverse is again a Gaussian integer) are the four numbers $\pm 1, \pm i$ and if we multiply any Gaussian integer $a + bi$ by any of these four units, we get the collection $\{a + bi, -a - bi, -b + ai, b - ai\}$. We measure the size of Gaussian integer by the square of its distance to the origin, i.e.,

$$|a + bi|^2 = a^2 + b^2.$$ 

This size function is called the norm of the Gaussian integer $a + bi$ and can also be thought of as the product of $a + bi$ and its “conjugate” $a - bi$. Note that the norm is a nice multiplicative function on the set of Gaussian integers, in that the norm of a product of two Gaussian integers is the product of the norms of each of them.

We have a natural notion of prime Gaussian integer, i.e., one with $a > 0$ and $b \geq 0$ that cannot be factored as the product of two Gaussian integers.
of smaller size. Given what we have just discussed, can you prove that if a Gaussian integer is a prime Gaussian integer, then its size must either be an ordinary prime number, or the square of an ordinary prime number?

Here is a plot of a large number of Gaussian primes as they display themselves amongst complex numbers:

---

To be done: Possibly a picture of the primes in the Gaussian integers. Here is an example:
http://en.wikipedia.org/wiki/File:Gauss-primes-768x768.png

---

Figure 31.1: Staircase of Gaussian primes of norm up to 14
Figure 31.2: Staircase of Gaussian primes of norm up to 100

Figure 31.3: Staircase of Gaussian primes of norm up to 1000
The natural question to ask, then, is: how are the Gaussian prime numbers distributed? Can one provide as close an estimate to their distribution and structure, as one has for ordinary primes? The answer, here is yes: there is a companion theory, with an analogue to the Riemann zeta function playing a role similar to the prototype $\zeta(s)$. And, it seems as if its “nontrivial zeroes” behave similarly: as far as things have been computed, they all have the property that their real part is equal to $\frac{1}{2}$. That is, we have companion Riemann Hypothesis.

This is just the beginning of a much larger story related to what has been come to be called the “Grand Riemann Hypotheses” and connects to analogous problems, some of them actually solved, that give some measure of evidence for the truth of these hypotheses. For example, for any system of polynomials in a fixed number of variables (with integer coefficients, say) and for each prime number $p$ there are “zeta-type” functions that contain all the information needed to count the number of simultaneous solutions in finite fields of characteristic $p$. That such counts can be well-approximated with a neatly small error term is related to the placement of the zeroes of these “zeta-type” functions. There is then an analogous “Riemann Hypothesis” that prescribes precise conditions on the real parts of their zeroes—this prescription being called the “Riemann Hypothesis for function fields.” Now the beauty of this analogous hypothesis is that it has, in fact, been proved!

Is this yet another reason to believe the Grand Riemann Hypothesis?
Chapter 32

Glossary

Here we give all the connections with the standard literature and conventional terminology that we restrained ourselves from giving in the text itself.

For the moment the list of entries is the following but it will expand.

$\pi(X) = \pi_0(X), \ Q(X) = \psi_0(X), \ \log, \ \exp, \ \delta, \ \text{distributions, RSA cryptography, Mersenne prime, Li}(x), \ \text{random walk, spectrum, harmonic, fundamental, frequency, phase, amplitude, band-pass, complex numbers, complex plane, Riemann Zeta function, zeroes of zeta.}$

Also mention Brian Conrey’s Notices article on RH as “among the best 12 pages of RH survey material that there is—at least for an audience of general mathematicians.” And mention Sarnak and Bombieri articles at the CMI website on RH.
Endnotes


[3] All of the figures were created using the free Sage mathematical software [?]. Complete Sage code is available at [http://wstein.org/rh](http://wstein.org/rh) that can be used to recreate every diagram in this book. More adventurous readers can to experiment with the parameters for the ranges of data illustrated, so as to get an even more vivid sense of how the numbers “behave.”.


[7] **How not to factor the numerator of a Bernoulli number:**
As mentioned in Chapter [30] the coefficient $B_k$ of the linear term of the polynomial

$$S_k(n) = 1^k + 2^k + 3^k + \cdots + (n - 1)^k$$

is (up to sign) the $k$-th **Bernoulli number**. These numbers are rational numbers and, putting them in lowest terms, their numerators play a role in certain problems, and their denominators in others. (This is an amazing story, which we can’t go into here!)

dealing (for various reasons) with the fraction \(-\frac{B_{200}}{400}\), where \(B_{200}\) is the two-hundredth Bernoulli number. The numerator of this fraction is quite large: it is—hold your breath—

\[
389 \cdot 691 \cdot 537005652867 \quad \text{times this 204-digit number:}
\]

\[
N := 345269032939215803146410928173696740406844815684239672101299206421451944591925694154456527606766236010874972724155570842527652727868776362959519620872735612200601036506871681124610986596878180738901486527
\]

and he incorrectly asserted that it was prime. Happily, Bartosz Naskręcki spotted this error: our 204-digit \(N\) is not prime.

How did he know this? By using the most basic test in the repertoire of tests that we have available to check to see whether a number is prime: we’ll call it the “\textit{Fermat 2-test.}” We’ll first give a general explanation of this type of test before we show how \(N\) fails the Fermat 2-test.

The starting idea behind this test is the famous result known as \textit{Fermat’s Little Theorem} where the “little” is meant to alliteratively distinguish it from you-know-what.

\textbf{Theorem 32.1} (Fermat’s Little Theorem). \textit{If} \(p\) \textit{is a prime number, and} \(a\) \textit{is any number relatively prime to} \(p\) \textit{then} \(a^{p-1} - 1\) \textit{is divisible by} \(p\).

A good exercise is to try to prove this, and a one-word hint that might lead you to one of the many proofs of it is \textit{induction}.\footnote{Here’s the proof:
\[(N + 1)^p \equiv N^p + 1 \equiv (N + 1),\] where the first equality is the binomial theorem and the second equality is induction.}

Now we are going to use this as a criterion, by—in effect—restating it in what logicians would call its \textit{contraPositive}:

\textbf{Theorem 32.2} (The Fermat \(a\)-test). \textit{If} \(M\) \textit{is a positive integer, and} \(a\) \textit{is any number relatively prime to} \(M\) \textit{such that} \(a^{M-1} - 1\) \textit{is not divisible by} \(M\), \textit{then} \(M\) \textit{is not a prime.}

Well, Naskręcki computed \(2^{N-1} - 1\) (for the 204-digit \(N\) above) and saw that it is not divisible by \(N\).\footnote{and has a residue of 3334581100595350251533969739282790317379460667738197064651672528599692566100005682 92727357926209571597827398131150654514506840724258354848985665112763692970799269354928195076056916221 73717318335512037458 after division by \(N\).} Ergo, our \(N\) fails the Fermat 2-test so is not prime.

But then, given that it is so “easy” to see that \(N\) is not prime, a natural question to ask is: what, in fact, is its prime factorization? This—it turns out—isn’t so easy to determine; Naskręcki devoted 24 hours of computer
time setting standard factorization algorithms on the task, and that was not sufficient time to resolve the issue. The factorization of the numerators of general Bernoulli numbers is the subject of a very interesting web site of Samuel Wagstaff ([http://homes.cerias.purdue.edu/~ssw/bernoulli](http://homes.cerias.purdue.edu/~ssw/bernoulli)). Linked to this web page one finds ([http://homes.cerias.purdue.edu/~ssw/bernoulli/composite](http://homes.cerias.purdue.edu/~ssw/bernoulli/composite)) which gives a list of composite numbers whose factorizations have resisted all attempts to date. The two-hundredth Bernoulli number is 12th on the list.

The page [http://en.wikipedia.org/wiki/Integer_factorization_records](http://en.wikipedia.org/wiki/Integer_factorization_records) lists record challenge factorization, and one challenge that was completed in 2009 involves a difficult-to-factor number with 232 digits; its factorization was completed by a large team of researchers and took around 2000 years of CPU time. This convinced us that with sufficient motivation it would be possible to factor \( N \), and so we asked some leaders in the field to try. They succeeded!

### Factorisation of B200

by Bill Hart on 4 Aug 05, 2012 at 07:24pm

We are happy to announce the factorization of the numerator of the 200th Bernoulli number:

\[
N = 389 \cdot 691 \cdot 5370056528687 \cdot c_{204}
\]

\[
c_{204} = p_{90} \cdot p_{115}
\]

\[
p_{90} = 149474329044343594528784250333645983079497454292
\]

\[
= 838248852612270757617561057674257880592603
\]

\[
p_{115} = 230988849487852221315416645031371036732923661613
\]

\[
= 619208811597595398791184043153272314198502348476
\]

\[
= 2629703896050377709
\]

The factorization of the 204-digit composite was made possible with the help of many people:

- William Stein and Barry Mazur challenged us to factor this number
- Sam Wagstaff maintains a table of factorizations of numerators of Bernoulli numbers at [http://homes.cerias.purdue.edu/~ssw/bernoulli/bnum](http://homes.cerias.purdue.edu/~ssw/bernoulli/bnum). According to this table, the 200th Bernoulli number is the 2nd smallest index with unfactored numerator (the first being the 188th Bernoulli number)
- Cyril Bouvier tried to factor the \( c_{204} \) by ECM up to 60-digit level, using the TALC cluster at Inria Nancy - Grand Est
- yoyo@home tried to factor the \( c_{204} \) by ECM up to 65-digit level, using the help of many volunteers of the distributed computing platform [http://www.rechenkraft.net/yoyo/](http://www.rechenkraft.net/yoyo/) after ECM was unsuccessful, we decided to factor the \( c_{204} \) by GNFS
Many people at mersenneforum helped for the polynomial selection. The best polynomial was found by Shi Bai, using his implementation of Kleinjung’s algorithm in CADO-NFS: http://www.mersenneforum.org/showthread.php?p=298264#post298264 sieving was performed by many volunteers NFS@home, thanks to Greg Childers. See http://escatter11.fullerton.edu/nfs for more details. This factorization showed that such a distributed effort might be feasible for a new record GNFS factorization, in particular for the polynomial selection. This was the largest GNFS factorization performed by NFS@home to date, the second largest being $2^{1040} + 1$ at 183.7 digits.

Two independent runs of the filtering and linear algebra were done: one by Greg Childers with msieve (http://www.boo.net/~jasonp/qs.html) using a 48-core cluster made available by Bill Hart, one by Emmanuel Thom and Paul Zimmermann with CADO-NFS (http://cado-nfs.gforge.inria.fr/), using the Grid 5000 platform.

The first linear algebra run to complete was the one with CADO-NFS, thus we decided to stop the other run.

Bill Hart

Verify the above in Sage as follows:

```
sage: p90 = 149474329044343594528784250336459830794974542928382\48852612270757617561057674257880592603
sage: p115 = 2309888494878522221315416645031371036732923661613619\2088115975953987911840431532723141985023484762629703896050377709
sage: c204 = p90 * p115
sage: 389 * 691 * 5370056528687 * c204 == -numerator(bernoulli(200))
True
sage: is_prime(p90), is_prime(p115), is_prime(c204)
(True, True, False).
```

[8] Given an integer $n$, there are many algorithms available for trying to write $n$ as a product of prime numbers. First we can apply trial division, where we simply divide $n$ by each prime $2, 3, 5, 7, 11, 13, \ldots$ in turn, and see what small prime factors we find (up to a few digits). After using this method to eliminate as many primes as we have patience to eliminate, we typically next turn to a technique called Lenstra’s elliptic curve method, which allows us to check $n$ for divisibility by bigger primes (e.g., around 10–15 digits). Once we’ve exhausted our patience using the elliptic curve method, we would next hit our number with something called the quadratic sieve, which works well for factoring numbers of the form $n = pq$, with $p$ and $q$ primes of roughly equal size, and $n$ having less than 100 digits (say, though the
100 depends greatly on the implementation). All of the above algorithms—and then some—are implemented in Sage, and used by default when you type `factor(n)` into Sage. Try typing `factor(some number, verbose=8)` to see for yourself.

If the quadratic sieve fails, a final recourse is to run the number field sieve algorithm, possibly on a supercomputer. To give a sense of how powerful (or powerless, depending on perspective!) the number field sieve is, a record-setting factorization of a general number using this algorithm is the factorization of a 232 digit number called RSA-768:

\[ n = 123018668453011775513049495838496272077285356959533479219732245215172640050726365751874520219978646938995647494277406384592519255732630345373154826850791702612214291346167042921431160222124047927473779408066535149597459569092143413 \]

which factors as \( pq \), where

\[ p = 334780716985689878604416984821269081770479498371376856891243138898288379387800228761471165231743087737814467999489 \]

and

\[ q = 367460436667995904282446337962795263227915816434308764267603228315739666511279233373417143396810270092798736308917 \]

We encourage you to try to factor \( n \) in Sage, and see that it fails. Sage does not currently (as of 2011) include an implementation of the number field sieve algorithm, though there are some free implementations currently available (see [http://www.boo.net/~jasonp/qs.html](http://www.boo.net/~jasonp/qs.html)).

[9] Nobody has ever published a proof that there is no fast way to factor integers. This is an article of “faith” among some cryptographers.


[11] We can use Sage (at [http://sagemath.org](http://sagemath.org)) to quickly compute the “hefty number” \( p = 2^{43,112,609} - 1 \). Simply type \( p = 2^{43112609} - 1 \) to instantly compute \( p \). In what sense have we computed \( p \)? Internally, \( p \) is now stored in base 2 in the computer’s memory; given the special form of \( p \) it is not surprising that it took little time to compute. Much more challenging is to compute all the base 10 digits of \( p \), which takes a few seconds: \( d = \text{str}(p) \). Now type \( d[-50:] \) to see the last 50 digits of \( p \). To compute the sum 58416637 of the digits of \( p \) type \( \text{sum}(p.digits()) \).

[12] The sequence of prime numbers we find by this procedure is discussed in more detail with references in the Online Encyclopedia of Integer Sequences [http://oeis.org/A126263](http://oeis.org/A126263).

[13] See [http://www.eff.org/press/archives/2009/10/14-0](http://www.eff.org/press/archives/2009/10/14-0) Also the 46th Mersenne prime was declared by Time Magazine to be one of the top 50 best “inventions” of 2008: [http://www.time.com/time/specials/packages/article/0,28804,1852747_1854195_1854157,00.html](http://www.time.com/time/specials/packages/article/0,28804,1852747_1854195_1854157,00.html)
[14] In contrast to the situation with factorization, testing integers of this size (e.g., the primes $p$ and $q$) for primality is relatively easy. There are fast algorithms that can tell whether or not any random thousand digit number is prime in a fraction of second. Try for yourself using the Sage command \texttt{is\_prime}. For example, if $p$ and $q$ are as in endnote 8 then \texttt{is\_prime(p)} and \texttt{is\_prime(q)} quickly output True and \texttt{is\_prime(p*q)} outputs False. However, if you type \texttt{factor(p*q, verbose=8)} you can watch as Sage tries forever and fails to factor $pq$.

[15] In Sage, the function \texttt{prime\_range} enumerates primes in a range, e.g., \texttt{prime\_range(50)} outputs the primes up to 50 and \texttt{prime\_range(50,100)} outputs the primes between 50 and 100. Typing \texttt{prime\_range(10^8)} in Sage enumerates the primes up to a hundred million in around a second. You can also enumerate primes up to a billion by typing \texttt{v=prime\_range(10^9)}, but this will use a large amount of memory, so be careful not to crash your computer if you try this. Notice that there are $\pi(10^9) = 50,847,534$ primes up to a billion by then typing \texttt{len(v)}. You can also compute $\pi(10^9)$ directly, without enumerating all primes, using the command \texttt{prime\_pi(10^9)}. This is much faster since it uses some clever counting tricks to find the number of primes without actually listing them all.

In Chapter 15 we tinkered with the staircase of primes by first counting both primes and prime powers. It turns out that there are comparatively few prime powers that are not prime. Up to $10^8$, only 1,405 of the 5,762,860 prime powers are not themselves primes. To see this, first enter \texttt{a = prime\_pi(10^8); pp = len(prime\_powers(10^8))}. Typing \texttt{(a, pp, pp-a)} then outputs the triple $(5761455, 5762860, 1405)$.

[16] For example, according to \url{http://oeis.org/A007508} there are 10,304,185,697,298 such pairs less than $10^{1000000000}$.  

[17] See \url{http://primes.utm.edu/largest.html#twin} for the top ten largest known twin primes.

[18] Hardy and Littlewood give a nice conjectural answer to such questions about gaps between primes. See Problem A8 of Guy’s book \textit{Unsolved Problems in Number Theory} (2004). Note that Guy’s book discusses counting the number $P_k(X)$ of pairs of primes up to $X$ that differ by a fixed even number $k$; we have $P_k(X) \geq \text{Gap}_k(X)$, since for $P_k(X)$ there is no requirement that the pairs of primes be consecutive.


[21] If \( f(x) \) and \( g(x) \) are real-valued functions of a real variable \( x \) such that for any \( \epsilon > 0 \) both of them take their values between \( x^{1-\epsilon} \) and \( x^{1+\epsilon} \) for \( x \) sufficiently large, then say that \( f(x) \) and \( g(x) \) are **good approximations of one another** if, for any positive \( \epsilon \) the absolute values of their difference is less than \( x^{\frac{1}{2}+\epsilon} \) for \( x \) sufficiently large. The functions \( \text{Li}(X) \) and \( R(X) \) of are good approximations of one another.


[23] This computation of \( \pi(X) \) was done by David J. Platt in 2012, and is the largest value of \( \pi(X) \) ever computed. See [http://arxiv.org/abs/1203.5712](http://arxiv.org/abs/1203.5712) for more details.

[24] In fact, \(| \text{Li}(X) - \pi(X) | \leq \sqrt{X} \log(X) \) for all \( X \geq 2.01 \). See Section 1.4.1 of Crandall-Pomerance’s book *Prime numbers, a computational perspective*.

[25] For a proof of this here’s a hint. Compute the difference between the derivatives of \( \text{Li}(x) \) and of \( x/\log x \). The answer is \( 1/\log^2(x) \). So you must show that the ratio of \( \int_2^X dx/\log^2(x) \) to \( \text{Li}(x) = \int_2^X dx/\log(x) \) tends to zero as \( x \) goes to infinity, and this is a good Calculus exercise.


[27] We have

\[
\psi(X) = \sum_{p^n \leq X} \log p
\]

where the summation is over prime powers \( p^n \) that are \( \leq X \).


[29] Discuss some good readable article on the Fast Fourier Transform algorithm; there are probably many such articles.

[30] See [http://en.wikipedia.org/wiki/Distribution_%28mathematics%](http://en.wikipedia.org/wiki/Distribution_%28mathematics%) for more about distributions. Also, Schwartz’s explains on page 238 of his autobiography: “Why did we choose the name distribution? Because, if \( \mu \) is a measure, i.e., a particular kind of distribution, it can be considered as a distribution of electric charges in the universe. Distributions give more general types of electric charges, for example dipoles and magnetic distributions. If we consider the dipole placed at the point \( a \) having magnetic moment \( M \), we easily see that it is defined by the distribution \(-D_M\delta(a)\). These objects occur in physics. Deny’s thesis, which he defended
shortly after, introduced electric distributions of finite energy, the only ones which really occur in practice; these objects really are distributions, and do not correspond to measures. Thus, distributions have two very different aspects: they are a generalization of the notion of function, and a generalization of the notion of distribution of electric charges in space. [...] Both these interpretations of distributions are currently used.”.

[31] From Wikipedia:

“Generalized functions” were introduced by Sergei Sobolev in 1935. They were independently introduced in the late 1940s by Laurent Schwartz, who developed a comprehensive theory of distributions.

[32] If the Riemann Hypothesis holds they are precisely the imaginary parts of the “nontrivial” zeroes of the Riemann zeta-function.

[33] The construction of \( \Phi(t) \) from \( \phi(X) \):

Succinctly, for positive real numbers \( t \),

\[
\Phi(t) := e^{-t/2} \Psi'(t),
\]

where \( \Psi(t) = \phi(e^t) \), and \( \Psi' \) is the derivative of \( \Psi(t) \) viewed as distribution. We extend this function to all real arguments \( t \) by requiring \( \Phi(t) \) to be an even function of \( t \), i.e., \( \Phi(-t) = \Phi(t) \). But, to review this in a more leisurely pace,

1. Distort the \( X \)-axis of our staircase by replacing the variable \( X \) by \( e^t \) to get the function

\[
\Psi(t) := \psi(e^t).
\]

No harm is done by this for we can retrieve our original \( \psi(X) \) as

\[
\psi(X) = \Psi(\log(X)).
\]

Our distorted staircase has risers at (0 and) all positive integral multiples of logs of prime numbers.

2. Now we’ll do something that might seem a bit more brutal: take the derivative of this distorted staircase \( \Psi(t) \). This derivative \( \Psi'(t) \) is a generalized function with support at all nonnegative integral multiples of logs of prime numbers.
Figure 32.1: $\Psi'(t)$ is a sum of Dirac delta functions at the logarithms of prime powers $p^n$ weighted by $\log(p)$ (and $\log(2\pi)$ at 0). The more red the arrow, the larger the weight.

3. Now—for normalization purposes—multiply $\Psi'(t)$ by the function $e^{-t/2}$ which has no effect whatsoever on the support.

Figure 32.2: Illustration of the staircase $\psi(x)$ constructed in Chapter 15 that counts weighted prime powers.

So what happens when we take the derivative—in the sense of distributions—of a complicated staircase? For example, see Figure 32.2. Well, we would have blip-functions (alias: Dirac $\delta$-functions) at each point of discontinuity of $\Psi(x)$; that is, at $x$ = any power of a prime.

Let us denote by $\Phi(t)$ the generalized function that resulted from the above carpentry:

$$\Phi(t) = e^{-t/2} \Psi'(t)$$
[34] **William:** Here we should put Andrew Granville’s write-up if he sends one. Also, we should include a picture of Granville!

[35] The theoretical story behind the phenomena that we have seen graphically in this chapter is yet another manifestation of Riemann’s explicit formula. Modern references for this are [Iwaniec-Kowalski] and [Davenport: Multiplicative Number Theory]. (See also the bibliography in these books.) Many ways of seeing the explicit relationship are given in Chapter 5 of [I-K]. For example, consider Exercise 5 on page 109.

\[ \sum \hat{\phi}(\rho) = -\sum_{n \geq 1} \Lambda(n) \phi(n) + I(\phi), \]

where

- \( \phi \) is any smooth complex-valued function on the reals with compact support,
- \( \hat{\phi} \) is its Mellin transform:
  \[ \hat{\phi}(s) := \int_0^\infty \phi(x)x^{s-1}dx, \]

- the last term on the right, \( I(\phi) \), is just
  \[ I(\phi) := \int_1^\infty (1 - \frac{1}{x^3} - \frac{1}{x}) \phi(x)dx \]
  (coming from the pole at \( s = 1 \) and the “trivial zeroes”).
- The more serious summation on the left hand side of the equation is over the nontrivial zeroes \( \rho \), noting that if \( \rho \) is a nontrivial zero so is \( \bar{\rho} \).

Of course, this ‘explicit formulation’ is not *immediately* applicable to the graphs we are constructing since we cannot naively take \( \hat{\phi} \) to be a function forcing the left hand side to be \( G_C(x) \).

See also Exercise 7 on page 112, which discusses the sums

\[ x - \sum_{|\theta| \leq C} \frac{x^{1+i\theta} - 1}{\frac{1}{2} + i\theta}. \]

[36] People who know that these correction terms are indexed by the nontrivial zeroes of the Riemann zeta-function may well ask how we propose to order them if RH is false; the following prescription will do: order them in terms of (the absolute value of) their imaginary part, and in the unlikely situation that there is more than one zero with the same imaginary part, order zeroes of the same imaginary part by their real parts, going from right to left.