



UNIVERSIDAD DE GRANADA

DE NOVO GENOME ASSEMBLY USING QUANTUM ANNEALING

JOSÉ ANTONIO ÁLVAREZ OCETE

Trabajo Fin de Grado

Doble Grado en Ingeniería Informática y Matemáticas

Tutores

Carlos Cano

Antonio Lasanta

FACULTAD DE CIENCIAS

E.T.S. INGENIERÍAS INFORMÁTICA Y DE TELECOMUNICACIÓN

Granada, a 28 de julio de 2021

ÍNDICE GENERAL

1.	QUANTUM MECHANICS MODEL	4
1.1.	Postulate 1: State Space	4
1.1.1.	Bra-ket notation	4
1.1.2.	Inner product and Hilbert Spaces	5
1.1.3.	Postulate 1 statement	6
1.1.4.	Quantum Computation Perspective: The Quantum Bit	7
1.2.	Postulate 2: Measurement	8
1.2.1.	Outer product	8
1.2.2.	Unitary and Hermitian operators	10
1.2.3.	Postulate 2 statement	11
1.2.4.	Real life qubit examples	14
1.3.	Postulate 3: Evolution	15
1.3.1.	Eigenvalues and eigenvectors	15
1.3.2.	Postulate 3 statement	16
1.3.3.	Quantum Computing perspective: Quantum Gates	19
1.4.	Postulate 4: Composite systems	20
1.4.1.	Tensor product	20
1.4.2.	Postulate 4 statement	23
1.4.3.	Projective measurement	24
1.4.4.	Quantum Computing perspective: Multiple qubits	26
1.5.	The no-cloning theorem	28
2.	THE D-WAVE QUANTUM COMPUTER	30
2.1.	Quantum Annealing	30
2.2.	Adiabatic Evolution	31
2.2.1.	Avoided crossing	32
2.2.2.	Solving problems using adiabatic evolution	34
2.3.	QUBO and Ising models	35
2.3.1.	The Max Cut problem	37
2.3.2.	General translation method	38
2.3.3.	Satisfiability problems: Max 2-SAT	40
2.3.4.	Graph Coloring	43
2.3.5.	The Travelling Salesman problem	46
2.4.	D-Wave Systems	54
2.4.1.	Quantum Annealing in D-Wave	54
	Bibliography	58

RESUMEN

TODO: resumen

TODO: que en el resumen en inglés no ponga en la cabecera Resumen.^{en} español.

QUANTUM MECHANICS MODEL

Quantum Mechanics are a mathematical framework in which quantum physics are developed. In this section, we will develop a quantum mechanics model in order to understand quantum computing. The Quantum Postulates will be our guidance. They provide a connection between the physical world and the mathematical formalization. We will provide context and formalization for each postulate, so both the mathematical precision and intuition notions are developed at the same time. This development is based on [1], [2] and [3].

1.1 POSTULATE 1: STATE SPACE

The first postulate sets the environment in which we will operate: The State Space. It will be a Hilbert space associated to a physical system. Let us rigorously define the necessary concepts using the Bra-ket notation. We will start by revisiting the required linear algebra.

1.1.1 *Bra-ket notation*

Let V be a complex vector space. That is, a vector space over \mathbb{C} . We will restrict our study to finite complex vector spaces. If z is a vector in V , we will denote its coordinates either as $z = (z_1, z_2, \dots, z_n)$ or by column notation:

$$z = \begin{pmatrix} z_1 \\ z_2 \\ \vdots \\ z_n \end{pmatrix}$$

Since V is a vector space we have two basic operations: (vector) addition and scalar multiplication.

In quantum mechanics, the usual notation is the Dirac's, also known as *bra-ket* notation. In this context, vectors in a complex vector space are denoted as $|\varphi\rangle$ and are known as *kets*. The only exception to this is the zero vector, which will be denoted as $0 = (0, \dots, 0)$ instead of $|0\rangle$ since $|0\rangle$ will be used as something completely different. A *vector subspace* W of V is a subset of W closed for addition and scalar multiplication.

A *base* of a vector space is a set of vectors $|v_1\rangle, \dots, |v_n\rangle$ such that they are linearly independent and any given vector $|v\rangle$ can be written as a linear combination of them: $|v\rangle = \sum_{i=1}^n \alpha_i |v_i\rangle$. The *dimension* of a vector space is the number of elements in any of its bases, which is independent from the chosen base.

Definition 1. Given two complex vector spaces V and W , a *linear operator* is an application $M : V \rightarrow W$ that is linear in its inputs:

$$M(\alpha|u\rangle + \beta|w\rangle) = \alpha M(|u\rangle) + \beta M(|w\rangle)$$

If V to W have dimensions n and m respectively, there is a bijection between the operators from V to W and the n by m matrices. Given an operator M , the obtained matrix M' is called the *matrix representation* of the linear operator. Furthermore, $M(|u\rangle) = M' \cdot |u\rangle$, so we usually denote the linear operator and its matrix representation by the same letter, and $M(|u\rangle)$ simply as $M|u\rangle$.

We will usually refer to linear operators simply as *operators*.

1.1.2 Inner product and Hilbert Spaces

Let us define another operation within the complex vector spaces.

Definition 2. Let V be a complex vector space. An inner product $\langle \cdot | \cdot \rangle : V^2 \rightarrow \mathbb{C}$ is a function such that:

1) $\langle \cdot | \cdot \rangle$ is sesquilinear. That is,

1.1) $\langle \cdot | \cdot \rangle$ is conjugate symmetric: for all u, v in V , $\langle u | v \rangle = \overline{\langle v | u \rangle}$.

1.2) $\langle \cdot | \cdot \rangle$ is linear on the second variable: for all u, v, w in V and α, β in \mathbb{C} :

$$\langle u | \alpha v + \beta w \rangle = \alpha \langle u | v \rangle + \beta \langle u | w \rangle$$

2) $\langle \cdot | \cdot \rangle$ is definite positive. That is, for all u in V , $\langle u | u \rangle \geq 0$ and $\langle u | u \rangle = 0 \iff u = 0$.

Given this properties it can easily be proven that $\langle \cdot | \cdot \rangle$ is also conjugate linear on the first variable. That is, for all u, v, w in V and α, β in \mathbb{C} :

$$\langle \alpha u + \beta v | w \rangle = \overline{\alpha} \langle u | w \rangle + \overline{\beta} \langle v | w \rangle$$

We will sometimes denote the inner product $\langle \cdot | \cdot \rangle$ as (\cdot, \cdot) to simplify notation.

Two vectors are said to be *orthonormal* if their inner product is zero. We define the norm of a vector $|v\rangle$ by:

$$\| |v\rangle \| = \sqrt{\langle v|v \rangle}$$

A *unit vector* is a vector $|v\rangle$ such that $\| |v\rangle \| = 1$. We also say that $|v\rangle$ is *normalized*, and we can normalize any vector except the zero vector by dividing it by its norm.

A base $|v_1\rangle, \dots, |v_n\rangle$ is said to be *orthonormal* if every vector is a unit vector and they are pairwise orthogonal. That is, $\langle v_i | v_j \rangle = \delta_{ij}$ where

$$\delta_{ij} = \begin{cases} 1 & \text{if } i = j \\ 0 & \text{if } i \neq j \end{cases}$$

Definition 3. An *inner product space* is a vector space with an associated inner product. A **Hilbert Space** is an inner product space that is also complete.

Hausdorff's Theorem states that every finite normed space is complete, therefore every finite inner product space over \mathbb{C} is a Hilbert space [4]. Again, by Hausdorff's theorem, we know that every n dimensional Hilbert space is isomorphic to \mathbb{C}^n . Thus, \mathbb{C}^n is the canonical n dimensional Hilbert space. Our study will be focused on these spaces.

Let $\alpha = a + i \cdot b \in \mathbb{C}$. We define the *conjugate*, $\bar{\alpha}$, as $\bar{\alpha} = a - i \cdot b$. The canonical inner product in \mathbb{C}^n is:

$$\langle u | v \rangle = \sum_{i=1}^n \bar{u}_i v_j$$

where $u = (u_1, \dots, u_n)$ and $v = (v_1, \dots, v_n)$, for every u, v in \mathbb{C}^n .

1.1.3 Postulate 1 statement

The reader should be familiar by now with the notation and the necessary linear algebra to formulate the first postulate.

Postulate 1. Associated to any isolated physical system is a complex vector space with an inner product (that is, a Hilbert space) known as the *state space* of the system. The system is completely described by its *state vector*, which is a unit vector in the system's state space.

An important concern with this postulate is that it does not tell us which is the state space of a given system, nor its state vector. Although up to this point we cannot formally assure this, in quantum computing the state space will be fixed: \mathbb{C}^{2^n} for an n -qubits system. Our evolving state vector will be a vector 2^n -vector.

Let us start by modulating a simpler system: a single qubit system.

Definition 4. A state vector of the state space \mathbb{C}^2 is called a **qubit**. Thus, \mathbb{C}^2 may be called a single qubit state space.

Suppose $|0\rangle, |1\rangle$ form an orthonormal basis of a 2-dimensional Hilbert space. Then, any state vector in this state space may be described as:

$$|\varphi\rangle = \alpha|0\rangle + \beta|1\rangle$$

where α, β are complex numbers called *amplitudes*. Thus, the condition that $|\varphi\rangle$ is a unit vector, $\langle\varphi|\varphi\rangle = 1$ is equivalent to $|\alpha|^2 + |\beta|^2 = 1$. This is known as the *normalization condition*.

We will always think of $|0\rangle, |1\rangle$ as a previously fixed orthonormal base. A linear combination of state vectors $\sum_i a_i|\varphi_i\rangle$ is called a *superposition* of the states $|\varphi_i\rangle$ with amplitudes a_i respectively. For example, the state

$$\frac{|0\rangle - |1\rangle}{\sqrt{2}}$$

is a superposition of the states $|0\rangle$ and $|1\rangle$ with amplitudes $1/\sqrt{2}$ and $-1/\sqrt{2}$ respectively.

1.1.4 Quantum Computation Perspective: The Quantum Bit

The bit is the minimum measure of information on classical computation and classical information theory. Everything in these fields is built from scratch based on bits. Likewise, quantum computing and quantum information theory are built upon the **qubit**.

We have reached the qubit definition from quantum physics and pure mathematics, describing the qubit as a mathematical object independent of its physical implementation. By describing them as mathematical entities we will be able to explore their properties mathematically without having to worry about the physics underneath. This allows us to construct the quantum computing and quantum information theories independently of the physical implementation.

So, intuitively, what is a qubit? Just like the classical bit, a qubit has a state. For the bit, the two only possible states are either 0 or 1. A qubit can take the states $|0\rangle$ and

$|1\rangle$ -corresponding to the classical states 0 and 1- or it can be in a *linear combination* of them:

$$|\varphi\rangle = \alpha|0\rangle + \beta|1\rangle$$

Where α and β are complex numbers. Thus, we can describe a qubit as a vector in a two-dimensional complex Hilbert space (the canonical \mathbb{C}^2), where $|0\rangle$ and $|1\rangle$ form an orthonormal basis called the *computational basis*. $|0\rangle$ and $|1\rangle$ will be called *computational basis states*.

At this point, the reader may ask themselves if a qubit may even physically exist, not just as a mathematical entity. After all, the first postulate states that given a *physical system*, there is an associated state space and vector states that describe the system. However, we define a qubit from state space (\mathbb{C}^2) without considering a physical system.

The answer is positive: there are numerous physical systems such that their associated state spaces are (\mathbb{C}^2). Thus, modeling a qubit. More intuitive examples are provided in section 1.2.4 and precise physical implementations are briefly discussed in section [TODOref].

1.2 POSTULATE 2: MEASUREMENT

The second postulate describes how states are ‘measured’, that is, how an outside observer may look inside the system. In classical physics, consider a simple system of a moving particle. ‘Measuring’ would be recording, for instance, the particle mass and speed at a given time. That is, someone **outside** the system would look **into** the system to record some information. Lastly, in the classical computation model measuring a bit is simply retrieving its content.

In quantum physics, measuring has some unexpected and sometimes counter-intuitive properties. Let us provide some linear algebra context before formulating the second postulate.

1.2.1 Outer product

Definition 5. Let V, W be two vector spaces and $|v\rangle \in V, |w\rangle \in W$. We define the *outer product* between $|v\rangle$ and $|w\rangle$, $|w\rangle\langle v|$, as the only linear operator such that for any $|v'\rangle \in V$,

$$(|w\rangle\langle v|) |v'\rangle = |w\rangle\langle v|v'\rangle = \langle v|v'\rangle |w\rangle$$

These identities provide a dual interpretation: the already known product of a complex value $\langle v|v' \rangle$ with a vector $|w\rangle$, and the application of the new operator, the outer product $|w\rangle\langle v|$ to the vector $|v'\rangle$. The outer product is defined so that this duality occurs.

Let us consider linear combinations of outer products. By definition, $\sum_i a_i |w_i\rangle\langle v_i|$ is the operator that transforms $|v'\rangle$ into $\sum_i a_i |w_i\rangle\langle v_i|v'\rangle = \sum_i a_i \langle v_i|v'\rangle |w_i\rangle$.

The most important result concerning outer products is the *completeness relation*:

Proposition 1 (Completeness relation). *Let $|i\rangle$ be any orthonormal basis of a finite vector space V . Then:*

$$\sum_i |i\rangle\langle i| = I$$

Proof. Let $|v\rangle \in H$. $|v\rangle$ can be expressed as $\sum_i v_i |i\rangle$ for some complex numbers v_i . Notice that $\langle i|v\rangle = v_i$. Therefor:

$$|v\rangle = \sum_i v_i |i\rangle = \sum_i \langle i|v\rangle |i\rangle = \sum_i |i\rangle\langle i|v\rangle = \left(\sum_i |i\rangle\langle i| \right) |v\rangle$$

Since $|v\rangle$ was arbitrary, this proves that $\sum_i |i\rangle\langle i| = I$. \square

Corollary 1 (Cauchy-Schwarz inequality). *For any two vectors $|v\rangle, |w\rangle$ in a Hilbert space,*

$$|\langle v|w\rangle|^2 \leq \langle v|w\rangle\langle w|w\rangle$$

where the equality occurs if and only if $|v\rangle$ and $|w\rangle$ are linearly dependant

Proof. We provide a proof supposed that our Hilbert space is finite.

Let $|v\rangle, |w\rangle$ be two vectors of a finite Hilbert space H . Since H is finite, using the Gram-Schmidt procedure we may obtain a basis $|i\rangle$ where the first vector is $|w\rangle / \sqrt{\langle w|w\rangle}$. Using the completeness relation $\sum_i |i\rangle\langle i| = I$:

$$\langle v|v\rangle\langle w|w\rangle = \langle v|I|v\rangle\langle w|w\rangle = \langle v| \sum_i (|i\rangle\langle i|) |v\rangle\langle w|w\rangle = \sum_i \langle v|i\rangle\langle i|v\rangle\langle w|w\rangle$$

The sum of a list of positive numbers is obviously greater than its first element, so:

$$\sum_i \langle v|i\rangle\langle i|v\rangle\langle w|w\rangle \geq \frac{\langle v|w\rangle\langle w|v\rangle}{\langle w|w\rangle} \langle w|w\rangle = \langle v|w\rangle\langle w|v\rangle = |\langle w|v\rangle|^2$$

Lastly, the equality occurs if and only if $\langle v|i\rangle = 0$ for every $|i\rangle \neq |w\rangle / \sqrt{\langle w|w\rangle}$. But since $|i\rangle$ is a base, this means $|v\rangle$ and $|w\rangle$ are linearly dependent. \square

1.2.2 Unitary and Hermitian operators

Another way of looking at the inner product is the *adjoint*.

Definition 6. Let A be an operator between \mathbb{C}^n and \mathbb{C}^m , finite dimensional Hilbert spaces. That is, $A \in \mathcal{M}_{n \times m}(\mathbb{C})$. Then, its *adjoint* or *conjugate transpose* A^\dagger is defined by:

$$(A^\dagger)_{ij} = \bar{A}_{ji}$$

If $|v\rangle$ is a vector, we can compute its adjoint by seeing it as a matrix. By convention, we will denote $|v\rangle^\dagger = \langle v|$. Adjoints of vector are usually called *bras*, making given sense to the *bra-ket* notation since $\langle v| \cdot |v\rangle = \langle v|v\rangle$, where \cdot denotes the dot product.

Some useful algebraic identities associated to adjoints are:

- Given an operator $A \in \mathcal{M}_n(\mathbb{C})$, A^\dagger is the only operator such that for any two vectors $|u\rangle, |v\rangle \in \mathbb{C}^n$: $(|u\rangle, A|v\rangle) = (A^\dagger|u\rangle, |v\rangle)$
- For any two operators A, B , $(AB)^\dagger = B^\dagger A^\dagger$.
- As a corollary, for any vector $|v\rangle$ and for any operator A , $(A|v\rangle)^\dagger = \langle v|A^\dagger$.

Definition 7. An operator A is said to be *normal* if $AA^\dagger = A^\dagger A$.

Characterization of normal operators is provided in Theorem 1. There are two particular cases of normal operators that will be of special interest to us:

Definition 8. An operator A is said to be *Hermitian* if its adjoint is itself: $A^\dagger = A$.

Definition 9. A matrix U is said to be *unitary* if $UU^\dagger = I$. Similarly, an operator is said to be *unitary* if $UU^\dagger = I$. A unitary operator U also fulfills that $U^\dagger U = I$.

Clearly, Hermitian and unitary operators are also normal. The importance of unitary matrices and operators in quantum computing lies in the following

Proposition 2. *Unitary operators preserve inner product between vectors. Thus, they also preserve the norm of a vector.*

Proof. Let $|u\rangle, |v\rangle \in \mathbb{C}^n$ and $U \in \mathcal{M}_n(\mathbb{C})$ be a unitary operator. Then:

$$(U|u\rangle, U|v\rangle) = \langle u|U^\dagger U|v\rangle = \langle u|I|v\rangle = \langle u|v\rangle = (|u\rangle, |v\rangle)$$

Which proves the proposition. □

An important type of Hermitian operators is the projectors.

Definition 10. Let W be a k -dimensional subspace of the d -dimensional vector space V . Let $|1\rangle, \dots, |d\rangle$ be an orthonormal base of V where $|1\rangle, \dots, |k\rangle$ is an orthonormal base of W . The *projector* onto the subspace W is defined by:

$$P = \sum_{i=1}^k |i\rangle\langle i|$$

It can easily be shown that this definition is independent from the chosen base $|1\rangle, \dots, |k\rangle$. Since $|v\rangle\langle v|$ is Hermitian for any vector $|v\rangle$, P is also Hermitian: $P = P^\dagger$. We will often refer to the subspace onto which P projects simply as P for commodity. The *orthonormal completement* of P is $Q \equiv I - P$. It can be verified the vector subspace Q is spanned by the base $|k+1\rangle, \dots, |d\rangle$.

As the reader may already imagine, n-qubits systems will be represented as vector states or *certain* state spaces. That is, as unitary vectors of certain Hilbert spaces. Thus, our definition of transformations on qubits must preserve their norm. These will be the qubits gates, which will be represented as unitary operators.

On the other hand, Hermitian operators will be key in order to study how a quantum system evolves with time using the Schrodinger equation in the third Postulate.

Finally, projectors will be used on as particular way of measurement and will be further discussed in section 1.4.3, once systems with multiple qubits have been introduced.

1.2.3 Postulate 2 statement

As previously discussed, the second postulate describes how a quantum system may be measured.

Postulate 2. Quantum measurement are described by a collection $\{M_m\}$ of *measurement operators*. These act on the state space associated to the physical system being measured. The index m refers to the measurement outcomes that may occur. That is, if $|\varphi\rangle$ is the vector state before measure, then the probability of the result m occurring is:

$$p(m) = \langle \varphi | M_m^\dagger M_m | \varphi \rangle$$

and the state of the system after the measurement is:

$$\frac{M_m |\varphi\rangle}{\sqrt{p(m)}}$$

Finally, the measurement operator satisfy the *completeness equation*:

$$\sum_m M_m^\dagger M_m = I$$

The completeness equation is equivalent to the fact that the probabilities of the different possible outcomes add up to one:

$$\sum_m p(m) = \sum_m \langle \varphi | M_m^\dagger M_m | \varphi \rangle = \langle \varphi | \sum_m (M_m^\dagger M_m) | \varphi \rangle = \langle \varphi | \varphi \rangle = 1$$

which holds for every state vector since they are unitary. Reciprocally, this equation occurring for every state vector $|\varphi\rangle$ implies the completeness equation.

The reader may have already noticed a huge difference between classic and quantum measurement: The measured state **changes** after the measurement. This means that we interfere with the system by merely looking into it! It will ultimately translate into huge differences between classical and quantum computing, such that we will generally not be able to clone a qubit (see theorem 3).

Let us look at an important example: *measurement of a qubit on the computational basis*. That is, measuring a qubit with two possible outcomes: $|0\rangle$ and $|1\rangle$. Although this is a particular case of projective measurement (further explained in section 1.4.3), it is worth introducing it now to deepen our understanding of qubits.

To obtain such results we use the operators $M_0 = |0\rangle\langle 0|$ and $M_1 = |1\rangle\langle 1|$. In order for $\{M_0, M_1\}$ to be a correct collection of measurement operators they must satisfy the completeness equation. Observe that each operator is Hemitian: $M_i^\dagger = (|i\rangle\langle i|)^\dagger = (\langle i|)^\dagger(|i\rangle)^\dagger = |i\rangle\langle i| = M_i$. Furthermore, $M_i^2 = |i\rangle\langle i|i\rangle\langle i| = |i\rangle\langle i| = M_i$.

Finally, the computational basis is an orthonormal basis and therefore the completeness relation tells us that:

$$\sum_i |i\rangle\langle i| = 1$$

We can see that the completeness equation holds:

$$\sum_i M_i^\dagger M_i = \sum_i M_i^2 = \sum_i M_i = \sum_i |i\rangle\langle i| = 1$$

Let's measure using this operators, also called *measure in the computational basis*, to better understand the measurement. Suppose the state being measured is $|\varphi\rangle = \alpha|0\rangle + \beta|1\rangle$. Then, the probability of obtaining the outcome 0 is:

$$p(0) = \langle \varphi | M_0^\dagger M_0 | \varphi \rangle = \langle \varphi | M_0 | \varphi \rangle = |\alpha|^2$$

Similarly, the probability of obtaining the outcome 1 is $p(1) = |\beta|^2$. Naturally, $|\alpha|^2 + |\beta|^2 = 1$. What happens after measuring? The post-measurement state will be, respectively if we measured 0 or 1:

$$\frac{M_0|\varphi\rangle}{|a|} = \frac{a}{|a|}|0\rangle$$

$$\frac{M_1|\varphi\rangle}{|b|} = \frac{b}{|b|}|1\rangle$$

Factors like $a/|a|$ are known as *phases*. An important result of quantum mechanics assures that multiplying by factors like these does not affect the state vector [1], so we virtually obtained $|0\rangle$ and $|1\rangle$. We may appreciate now how dividing by $\sqrt{p(m)}$ in the post-measurement state is only done so the resulting vector is a unit vector.

So if we measured a 0, the post-measurement vector will be $|0\rangle$ and vice-versa. Any measurements performed after the first one will yield exactly the same result. This behavior is called *qubit collapsing*.

We defined measurement to be independent of any basis so we may measure in the most convenient basis at each point. For example, we may use the following other basis:

$$|+\rangle = \frac{1}{\sqrt{2}}|0\rangle + \frac{1}{\sqrt{2}}|1\rangle = \frac{|0\rangle + |1\rangle}{\sqrt{2}}$$

$$|-\rangle = \frac{1}{\sqrt{2}}|0\rangle - \frac{1}{\sqrt{2}}|1\rangle = \frac{|0\rangle - |1\rangle}{\sqrt{2}}$$

It can easily be proven that $\{|+\rangle, |-\rangle\}$ is a basis and therefore it satisfies the completeness equation. In fact, the operators $M_+ = |+\rangle\langle +|$ and $M_- = |-\rangle\langle -|$ satisfy the completeness equation. Using this operators, measuring the qubit $|\varphi\rangle = \alpha|0\rangle + \beta|1\rangle$ will output + with probability:

$$\begin{aligned} p(+) &= \langle\varphi|M_+^\dagger M_+|\varphi\rangle = \langle\varphi|M_+|\varphi\rangle = \langle\varphi|+ \rangle\langle +|\varphi\rangle = \langle\varphi|+ \rangle^2 = \\ &= \left(\frac{\alpha}{\sqrt{2}}\langle 0|0\rangle + \frac{\alpha}{\sqrt{2}}\langle 1|0\rangle + \frac{\beta}{\sqrt{2}}\langle 0|1\rangle + \frac{\beta}{\sqrt{2}}\langle 1|1\rangle \right)^2 = \frac{(\alpha + \beta)^2}{2} \\ p(-) &= \frac{(\alpha - \beta)^2}{2} \end{aligned}$$

And, naturally, $p(+) + p(-) = 1$. So $|\varphi\rangle$ will collapse to $|+\rangle$ with probability $(\alpha + \beta)^2/2$ when measured with this operators. Observe that the post-measurement state will never be $|0\rangle$ nor $|1\rangle$ in this case, we made the qubit collapse to the chosen state basis.

Finally, there is a technical fineness between the first and second postulates worth mentioning. The first postulate was stated for an *isolated* physical system, but by measuring the system we interfere with it. However, measuring devices are also quantum systems, so together the measured and the measuring system form a larger isolated system (it may be necessary to include more quantum systems, but this can be done).

1.2.4 Real life qubit examples

In classical computation, we may know the state of a bit by consulting it. That is, what can simply retrieve that information from the bit. The first difficulty we find in quantum computing is that once we *measure* a qubit it *collapses* to either $|0\rangle$ with probability $|\alpha|^2$, or to $|1\rangle$ with probability $|\beta|^2$. The obtained output reflects the qubit state *after* it has collapsed to either one of these states, so the outcome may only be either $|0\rangle$ or $|1\rangle$. This means that we may never retrieve directly the values α and β . Thus, being unable to clone a qubit. This is the main idea behind the no-cloning theorem 3.

We can, however, initialize qubits in a certain state and apply some operations to them in order to alter their coefficients, thus knowing their exact value. However, once a single measurement is done, the qubit collapses and the α and β values are 'lost'.

Superposition and collapsing might be counter-intuitive concepts, so let us look at them with an analogy. We can think of a perfect coin being tossed as the following qubit:

$$|+\rangle = \frac{1}{\sqrt{2}}|0\rangle + \frac{1}{\sqrt{2}}|1\rangle$$

This does **not** represent a coin that has landed somehow on its side, but a spinning coin that has not landed yet. Upon measuring it, we 'make the coin land' and see the result: either heads or tails, and neither of the states in between. This example also describes the qubit collapse: once the coin has landed, we will see the same result every time we look at it -obviously-, just like every time a qubit is measured after the first measurement, the outcome will be the same since it has already collapsed. We will return to this state, also known as the Bell state, in section 1.4.2.

On the other hand, this was quite an inaccurate example since the system is not really isolated, although it was interesting intuition-wise. One of the first (accurate) qubit models ever proposed was the Schrodinger's Cat [5] [6]. In this hypothetical experiment, a cat would be locked in a room for an hour with a device that during that hour would *perhaps* trigger, killing the cat. On the other hand, with equal probability, it would not trigger at all. After the whole hour elapses, the cat would be alive and dead with equal probability, ending up in a halfway state. In this case, our computational bases would be the states alive and dead, and we achieve the state $|+\rangle$ after

that hour. Once we open the room and check on the cat, our qubit collapses to either state and stays on it until further disturbance.

Although physical implementations are discussed in section [TODOref chapter], we cannot proceed any further without providing a more accurate and reproducible description of a qubit than 'a coin being tossed' and such a hypothetical cat experiment. A possible realization of a qubit is an electron in a single atom's orbit, as seen in Figure 1. An electron in an orbit may be in the so-called *ground* and *excited* states, $|0\rangle$ and $|1\rangle$ respectively, depending on its energy. By shining light to the electron with a certain energy and for a certain amount of time, one can make the electron move from the ground state to the excited state and vice versa. But most interestingly, one can apply the light to the electron during a smaller amount of time, moving the electron somehow 'halfway' between both states.

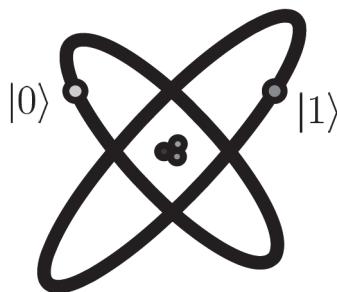


Figure 1: Qubit represented by two electron orbits in an atom, [1].

1.3 POSTULATE 3: EVOLUTION

The third postulate of quantum mechanics modulates the evolution of a quantum system. That is, the evolution of the state vector that describes the system. In order to properly formalize it, the concepts of eigenvectors, eigenvalues, and Hermitian operators are required.

1.3.1 Eigenvalues and eigenvectors

Definition 11. Let V be a vector space and A an operator on V . An *eigenvector* is a non zero vector $|v_\lambda\rangle$ such that $A|v_\lambda\rangle = \lambda|v_\lambda\rangle$ for a complex value λ called the associated *eigenvalue*.

Eigenvalues and their associated eigenvectors will usually be denoted with the same letter for simplicity: λ and $|\lambda\rangle$. We assume the reader is familiar with eigenvectors and values basic notions. For instance, that they may be calculated using the *characteristic equation*: $|I - \lambda A| = 0$.

Definition 12. A *diagonal representation* of an operator A is a representation $\sum_i \lambda_i |\lambda_i\rangle\langle\lambda_i|$ where the $|\lambda_i\rangle$ form an orthonormal set of A 's eigenvectors and λ_i are the respective eigenvalues. An operator is said to be *diagonalizable* if it allows a diagonal representation. The *spectrum* of an operator is the set of its eigenvalues.

Example 1. As an example of this, let us consider the following matrix:

$$Z = \begin{pmatrix} 1 & 0 \\ 0 & -1 \end{pmatrix}$$

This matrix is called the *Z Pauli matrix*. It is relevant for quantum computing and it will be introduced later on along with the rest of the Pauli matrices. For now, Let's compute its diagonalizable representation. Since it is already diagonal we can infer that its eigenvalues are $\{1, -1\}$. Computing the diagonal representation we realize that a pair orthonormal eigenvectors are $\{|0\rangle, |1\rangle\}$ respectively. Therefore:

$$Z = \begin{pmatrix} 1 & 0 \\ 0 & -1 \end{pmatrix} = |0\rangle\langle 0| - |1\rangle\langle 1|$$

Normal operators have significant relevancy thanks to the following result:

Theorem 1 (Spectral Decomposition Theorem). *An operator A is normal if and only if it is diagonalizable.*

Proof. TODO: To be copied, Box 2.2, page 72, Nielschenchen. □

Since Hermitian and unitary operators are normal, it follows the next

Corollary 2. *Any Hermitian operator is diagonalizable. Any unitary operator is diagonalizable.*

1.3.2 Postulate 3 statement

Postulate 3. The evolution of a *closed* quantum system is described by a *unitary transformation*. That is, the state $|\varphi\rangle$ of the system at time t_1 is related to the state $|\varphi'\rangle$ of the system at time t_2 by a unitary operator U which depends only on the times t_1 and t_2 ,

$$|\varphi'\rangle = U|\varphi\rangle$$

Just like the first postulate does not provide the state space or state vector of the system, the third postulate does not provide the unitary transformation that concretes this evolution. For our quantum computing case, we will be the ones to define the

unitary transformation to the system. That is, the quantum circuit that transforms our qubit.

Example 2. Let's consider the *X Pauli matrix*, also known as the *bit flip* matrix:

$$\sigma_1 = X = \begin{pmatrix} 0 & 1 \\ 1 & 0 \end{pmatrix}$$

It is called the bit flip matrix because it takes $|0\rangle$ to $|1\rangle$ and vice-versa:

$$X|0\rangle = \begin{pmatrix} 0 & 1 \\ 1 & 0 \end{pmatrix} \begin{pmatrix} 1 \\ 0 \end{pmatrix} = \begin{pmatrix} 0 \\ 1 \end{pmatrix} = |1\rangle$$

This product that we just computed is precisely what postulate 3 states: the evolution of our state vector following the unitary operator X : $X|0\rangle = |1\rangle$. Even though for an arbitrary system we do not know the specific unitary transformation the system follows, we can create systems that follow certain desired transformations. These are the basics of quantum gates and quantum circuits.

The description of the system evolution provided by Postulate 3 only bears information for those fixed times t_1 and t_2 . A continuous time-description of this evolution is provided by the Schrodinger equation, which provides a redefinition of the second postulate.

Postulate 3'. The time evolution of the state of a *closed* quantum system is described by the Schrodinger equation:

$$i\hbar \frac{d|\varphi\rangle}{dt} = H|\varphi\rangle$$

where \hbar is *Planck's constant*, i is the imaginary unit and H is a fixed Hermitian operator known as the *Hamiltonian*.

There are several notes to make about this postulate. First, the Hamiltonian is fixed for the given system and it is not be confused with the *Hadamard quantum gate*, also represented by an H . Second, \hbar is a physical constant that can be absorbed into the Hamiltonian for our purposes, simplifying the equation. Finally, this is a differential equation, so by knowing the initial state space of the system and the exact Hamiltonian we may know the exact evolution of the system.

Let us study the Hamiltonian in general. Since it is a Hermitian operator, it allows a spectral decomposition by theorem 1:

$$H = \sum_E E|E\rangle\langle E|$$

where E are the eigenvalues and $|E\rangle$ the respective normalized eigenvectors. the states $|E\rangle$ are usually referred to as *energy eigenstates* or *stationary states*, and E is the *energy* of the state $|E\rangle$. Furthermore, the lowest energy is called the *ground energy state* while the corresponding eigenstate is called the *ground state*.

Example 3. Suppose a single qubit system has the following Hamiltonian:

$$H = \hbar\omega X = \hbar\omega \begin{pmatrix} 0 & 1 \\ 1 & 0 \end{pmatrix}$$

Where X is the first Pauli matrix, and ω is a positive parameter. The eigenenergy states are the same as X eigenstates: $(|0\rangle + |1\rangle)/2$ and $(|0\rangle - |1\rangle)/2$ with respective energies $\hbar\omega$ and $-\hbar\omega$. Thus, the ground state is $(|0\rangle - |1\rangle)/2$ with ground state energy $-\hbar\omega$.

Let us deduce the connection between the Hamiltonian perspective of dynamics, Postulate 3', and the unitary operator perspective, postulate 3. We can solve the Schrödinger equation, which can be proven to be:

$$|\varphi(t_2)\rangle = \exp\left(\frac{-iH(t_2 - t_1)}{\hbar}\right) |\varphi(t_1)\rangle$$

We define:

$$U(t_1, t_2) \equiv \exp\left(\frac{-iH(t_2 - t_1)}{\hbar}\right)$$

Which is a unitary operator. In fact, any unitary operator U may be expressed as $U = \exp(iK)$ for some Hermitian operator K . Thus, we obtained:

$$|\varphi(t_2)\rangle = U(t_1, t_2) |\varphi(t_1)\rangle$$

Following this procedure, we have proven that there is a one to one correspondence between the continuous time-varying postulate 3' using the Hamiltonian and the more stationary discrete-time version using the unitary operator. Although the discrete-time vision is usually used in quantum computing, Quantum Annealing -the specific application of quantum mechanics used in this thesis- rests mostly on the Hamiltonian point of view. We will deepen in this architecture on section [TODOref].

It is worth mentioning that both versions of this postulate assume our physical system to be *closed*. That is, there is no interaction with the system coming from the exterior. In reality, the only real closed system is the universe as a whole. However, we may recreate sufficiently closed systems so that they can be described with approximations as being closed.

Furthermore, this severely interferes with postulate 2, where an outsider to the system may interfere with it by measuring it. The quantum system will evolve following postulate 3 until measurement is applied. Then they will evolve following the behavior described on postulate 2.

In practice, obtaining the Hamiltonian for a given quantum system is a really laborious work and usually needs experimental data [TODO: add evidence? stated as such in Niel.]. However, for our computational purposes, we will be the ones designing the Hamiltonian such that our system evolves as desired. In particular, chapter [TODOref to QUBO chapter] describes in detail the construction of Hamiltonians for QUBO problems.

1.3.3 Quantum Computing perspective: Quantum Gates

Although quantum annealing does not make use of quantum gates, they are the basis of quantum computing and they fully rely on postulate 3. Because of their importance, a brief overview of quantum gates is provided ojin this section. However, a profound understanding of them will not be necessary to understand the rest of the thesis.

Quantum computing is built upon the most simple operation we can compute on single qubits: quantum gates. Since state vectors are unit two-dimensional vectors, the operations we apply to them must preserve the norm. Thus, quantum gates will be represented by 2×2 unitary matrices. We have already introduce some of the most famous gates, the Pauli matrices:

$$X = \begin{pmatrix} 0 & 1 \\ 1 & 0 \end{pmatrix}; \quad Y = \begin{pmatrix} 0 & -i \\ i & 0 \end{pmatrix}; \quad Z = \begin{pmatrix} 1 & 0 \\ 0 & -1 \end{pmatrix}$$

Some of the others most important gates are the Hadamard gate (denoted as H), the phase gate (denoted as S), and the $\pi/8$ gate (denoted as T):

$$H = \frac{1}{\sqrt{2}} \begin{pmatrix} 1 & 1 \\ 1 & -1 \end{pmatrix}; \quad S = \begin{pmatrix} 1 & 0 \\ 0 & i \end{pmatrix}; \quad T = \begin{pmatrix} 1 & 0 \\ 0 & e^{i\pi/4} \end{pmatrix}$$

Although we have not described the fundamentals of multiple qubits systems (postulate 4 is required for this), it is worth mentioning a simple two-qubits gate for the purpose of this section. Just like in classical computing, conditional operations are essential for the construction of complex quantum algorithms. The most basic conditional operation is the *controlled-NOT* gate, also referred to as the CNOT gate. This gate takes two bits $|c\rangle|t\rangle$, a control qubit, and a target qubit. It flips the target qubit if the first qubit is set to $|1\rangle$, producing the operation $|c\rangle|t\rangle \rightarrow |c\rangle|c \oplus t\rangle$, where \oplus denotes the exclusive-or operation. Therefore, if the control qubit is $|c\rangle = \alpha|0\rangle + \beta|1\rangle$,

it will virtually flip the target qubit with probability $|\beta|^2$ and leave it how it is with probability $|\alpha|^2$.

One of the most important results of classical computation theory is that any boolean function may be constructed by only using AND, OR, and NOT gates, or by simply using NAND gates [7]. There is an equivalent result concerning quantum gates [3].

Theorem 2. *Any unitary matrix can be approximated by a combination of the Hadamard, CNOT, and $\pi/8$ gates.*

In this case, $\{H, CNOT, T\}$ is called a *universal gate set*. In practice, is it efficient to build such any gate using only this gate set? The answer to the question is positive, as stated by the **Solovay-Kitaev theorem** [8]. However, this is out of the scope of this project. We provide a sketch proof of the previous theorem from [3].

Proof. TODO: copy proof from Bayens, Theorem 4.2 □

Further study of quantum circuits falls out of the scope of this thesis since we will use the D-Wave architecture, which relies on quantum annealing instead of quantum circuits.

1.4 POSTULATE 4: COMPOSITE SYSTEMS

In this section, we introduce the last postulate, which lets us understand the state space associated with a physical system composed of other minor systems. This will allow us to study multiple-qubits systems and state some of the most important results in quantum mechanics and quantum computing: The *Heisenberg uncertainty principle* and the no-cloning theorem.

1.4.1 Tensor product

Let V and W be complex vector spaces with dimensions m and n respectively. Then, $V \otimes W$, read ' V tensor W ', is a mn complex vector space. Let $|v\rangle$ and $|w\rangle$ be vectors in V and W respectively. Then, $|v\rangle \otimes |w\rangle$ is in $V \otimes W$. Furthermore, any element of $V \otimes W$ may be expressed as a linear combinations of tensor products $|v\rangle \otimes |w\rangle$ of elements from V and W . We may describe elements of $V \otimes W$ using the following equivalent notations: $|v\rangle \otimes |w\rangle$, $|v\rangle|w\rangle$, $|v,w\rangle$ and even $|vw\rangle$.

Let $|i\rangle$ and $|j\rangle$ be basis for V and W respectively. Then, $|i\rangle \otimes |j\rangle = |ij\rangle$ is a basis for $V \otimes W$. For example, consider the complex vector space \mathbb{C}^2 . Then

$$|0\rangle \otimes |0\rangle + |1\rangle \otimes |1\rangle = |00\rangle + |11\rangle$$

is an element in $\mathbb{C}^2 \otimes \mathbb{C}^2 \cong \mathbb{C}^4$.

Although the tensor product is generally described as an abstract construct, we may provide a more visual perspective using the *Kronecker product*. This is a matrix representation for the tensor product of finite vector spaces. Suppose $A = \{a\}_{ij}$ is a $n \times m$ matrix and $B = \{b\}_{ij}$ is an $p \times q$ matrix. Then $A \otimes B$ is a $np \times mq$, matrix with the following representation:

$$A \otimes B = \begin{pmatrix} a_{11}B & a_{12}B & \dots & a_{1m}B \\ a_{21}B & a_{22}B & \dots & a_{2m}B \\ \vdots & \vdots & \vdots & \vdots \\ a_{n1}B & a_{n2}B & \dots & a_{nm}B \end{pmatrix}$$

For example:

$$\begin{pmatrix} 0 & 1 \\ 2 & 3 \end{pmatrix} \otimes \begin{pmatrix} 10 & 20 \\ 30 & 40 \end{pmatrix} = \begin{pmatrix} 0 \times \begin{pmatrix} 10 & 20 \\ 30 & 40 \end{pmatrix} & 1 \times \begin{pmatrix} 10 & 20 \\ 30 & 40 \end{pmatrix} \\ 2 \times \begin{pmatrix} 10 & 20 \\ 30 & 40 \end{pmatrix} & 3 \times \begin{pmatrix} 10 & 20 \\ 30 & 40 \end{pmatrix} \end{pmatrix} =$$

$$\begin{pmatrix} 0 \times 10 & 0 \times 20 & 1 \times 10 & 1 \times 20 \\ 0 \times 30 & 0 \times 40 & 1 \times 30 & 1 \times 40 \\ 2 \times 10 & 2 \times 20 & 3 \times 10 & 3 \times 20 \\ 2 \times 30 & 2 \times 40 & 3 \times 30 & 3 \times 40 \end{pmatrix} = \begin{pmatrix} 0 & 0 & 10 & 20 \\ 0 & 0 & 30 & 40 \\ 20 & 40 & 30 & 60 \\ 60 & 80 & 90 & 120 \end{pmatrix}$$

By definition, the tensor product satisfies:

- Let α be a complex number. For any $|v\rangle$ in V and $|w\rangle$ in W ,

$$\alpha(|v\rangle \otimes |w\rangle) = (\alpha|v\rangle) \otimes |w\rangle = |v\rangle \otimes (\alpha|w\rangle)$$

- For any $|v_1\rangle, |v_2\rangle$ in V and $|w\rangle$ in W ,

$$(|v_1\rangle + |v_2\rangle) \otimes |w\rangle = |v_1\rangle \otimes |w\rangle + |v_2\rangle \otimes |w\rangle$$

- For any $|v\rangle$ in V and $|w_1\rangle, |w_2\rangle$ in W ,

$$|v\rangle \otimes (|w_1\rangle + |w_2\rangle) = |v\rangle \otimes |w_1\rangle + |v\rangle \otimes |w_2\rangle$$

The inner products of the spaces V and W may be used to extend and natural inner product the tensor space. Define:

$$\left(\sum_i a_i |v_i\rangle \otimes |w_i\rangle, \sum_j b_j |v'_j\rangle \otimes |w'_j\rangle \right) \equiv \sum_{ij} \bar{a}_i b_j \langle v_i | v'_j \rangle \langle w_i | w'_j \rangle$$

With this product and the previous properties in mind, it can easily be proven that the tensor product of Hilbert spaces is a Hilbert space.

Proposition 3. Let H_1 be and H_2 be two Hilbert spaces with respective orthonormal bases $B_1 = \{|v_i\rangle\}_{i=1,\dots,n}$ and $B_2 = \{|w_j\rangle\}_{j=1,\dots,m}$. Then, $H_1 \otimes H_2$ is a Hilbert space with inner product given by

$$(|v\rangle \otimes |w\rangle, |v'\rangle \otimes |w'\rangle) = \langle vw|v'w'\rangle = \langle v|v'\rangle \langle w|w'\rangle$$

For any $|v\rangle, |v'\rangle$ in V and $|w\rangle, |w'\rangle$ in W . Furthermore, $B_1 \otimes B_2 = \{|v_i w_j\rangle\}_{i,j}$ is an orthonormal basis of $H_1 \otimes H_2$.

Proof. By definition, $H_1 \otimes H_2$ is a complex vector space with $\dim(H_1 \otimes H_2) = \dim(H_1) \cdot \dim(H_2)$. Furthermore, $B_1 \otimes B_2 = \{|v_i\rangle \otimes |w_j\rangle\}_{i,j} = \{|v_i w_j\rangle\}_{i,j}$ is a base of it. Let $|v_i w_j\rangle, |v_k w_l\rangle \in B_1 \otimes B_2$:

$$\langle v_i w_j | v_k w_l \rangle = \langle v_i | v_k \rangle \langle w_j | w_l \rangle = \delta_{ik} \delta_{jl}$$

which equals one if and only if $i = k$ and $j = l$, and zero otherwise. Thus, proving that $B_1 \otimes B_2$ is orthonormal.

Finally, for $H_1 \otimes H_2$ to be a Hilbert space the product defined above needs to be, in fact, an inner product. It suffies to prove that it is definite positive. Let $|v\rangle \in H_1 \otimes H_2$. Since $B_1 \otimes B_2$ is an orthonormal basis, there exist $\alpha_{ij} \in \mathbb{C}$ such that:

$$|v\rangle = \sum_{i,j} \alpha_{ij} |v_i w_j\rangle$$

By linearity:

$$\langle v | v \rangle = \sum_{i,j,k,l} \overline{\alpha_{ij}} \alpha_{kl} \langle v_i w_j | v_k w_l \rangle = \sum_{i,j,k,l} \overline{\alpha_{ij}} \alpha_{kl} \delta_{ik} \delta_{jl} = \sum_{i,j} |\alpha_{ij}|^2 \geq 0$$

where the equality holds if and only if $\alpha_{ij} = 0 \forall i, j$. □

From the inner product, the tensor space $H_1 \otimes H_2$ naturally inherits the notions of adjoint, unitary, normality and Hermicity. We will denote a vector $|\varphi\rangle$ tensored with itself n times, $|\varphi\rangle \otimes \dots \otimes |\varphi\rangle$, as $|\varphi\rangle^{\otimes n}$, and equivalently with Hilbert spaces: $H \otimes \dots \otimes H$, as $H^{\otimes n}$. For our quantum computing purposes, the following case holds particular relevance.

Corollary 3. The tensor product of \mathbb{C}^2 with itself n times, $\mathbb{C}^2 \otimes \dots \otimes \mathbb{C}^2$, is isomorphic to \mathbb{C}^{2^n} .

Proof. Let $H = \mathbb{C}^2 \otimes \dots \otimes \mathbb{C}^2$. Since $\dim(H) = \dim(\mathbb{C}^2)^n = 2^n$, H is a Hilbert space with dimension 2^n . Thus, by Hausdorff's Theorem [4], it is isomorphic to the canonical complex vector space of dimension 2^n , \mathbb{C}^{2^n} . □

In the same way that the inner product was naturally extended to the tensor space, we may extend operators ensuring linearity. Let $A : V \rightarrow V'$ and $B : W \rightarrow W'$ two operators between Hilbert spaces. Then, we define $A \otimes B : V \otimes W \rightarrow V' \otimes W'$ using the following equation:

$$(A \otimes B)(|v\rangle \otimes |w\rangle) \equiv A|v\rangle \otimes B|w\rangle$$

We extend this definition to every element of $V \otimes W$ ensuring linearity:

$$(A \otimes B)\left(\sum_i a_i |v_i\rangle \otimes |w_i\rangle\right) \equiv \sum_i a_i A|v_i\rangle \otimes B|w_i\rangle$$

It can be shown that $A \otimes B$ is a well-defined linear operator. In fact, any linear operator mapping $V \otimes W$ to $V' \otimes W'$ may be expressed as a linear combination of tensor product of linear operators mapping V to V' and W to W' :

$$C = \sum_i a_i A_i \otimes B_i$$

where by definition:

$$\left(\sum_i a_i A_i \otimes B_i\right)|v\rangle \otimes |w\rangle = \sum_i a_i A_i|v\rangle \otimes B_i|w\rangle$$

1.4.2 Postulate 4 statement

Reaching the core of this section, suppose we consider two (or more) distinct physical systems. In order to describe the state of the composite system we need to make use of tensor product.

Postulate 4. The state space of a composite physical system is the tensor product of the state spaces of the component physical systems. Moreover, if we have systems numbered 1 through n , and system number i is prepared in the state $|\varphi_i\rangle$, then the joint state of the total system is $|\varphi_1\rangle \otimes \dots \otimes |\varphi_n\rangle$.

Let us provide an intuitive notion of why is the tensor product used in order to describe composite systems. After all, we would expect that there exists a *somehow canonical way* of describing the product composition of systems, just like the cartesian product is used with vector spaces. Let us refer again to the so-called *superposition principle of quantum mechanics*. Let $|\varphi\rangle$ and $|\psi\rangle$ be two vector states of a system. Then, any superposition of the states, $\alpha|\varphi\rangle + \beta|\psi\rangle$, should also be another vector state, where $|\alpha|^2 + |\beta|^2 = 1$. Suppose we now have two systems X and Y , with two vector states

$|\varphi_x\rangle$ and $|\varphi_y\rangle$. Then, we could describe the state of the composite system XY as some state $|\varphi_x\rangle|\varphi_y\rangle$. Studying the superposition principle on composite systems leads us naturally to the notion of tensor product. It is important to note that this development is not thorough, since we are not taking the superposition principle as a fundamental component of our description of quantum mechanics. It does help up to build an intuitive notion on why this tensor product is used.

We now introduce the *Bell State* or *EPR pair*:

$$|+\rangle = \frac{|00\rangle + |11\rangle}{\sqrt{2}}$$

Although it may seem harmless at first glance, this state has been responsible for many surprises during the development of quantum physics [9]. Let us have a first look into it, although we will come back to it later on.

The first thing to notice about this state is that it can be expressed as a single tensor product of state vector: $|+\rangle \neq |a\rangle|b\rangle$. This means that this state is not describing two independent physical systems put together with their respective vector states independently, it describes a **relation** between both systems too. It means that those systems are somehow interfering with each other. So we may affect one by disrupting the other.

This phenomenon is called *entanglement*. A state is called *entangled* when it cannot be expressed as the product of states of its component systems. This concept rests at the heart of the disparity between classic physics and quantum physics and it is key in quantum computing. It was deeply studied first by Einstein, Podolsky, and Rosen (EPR) [10] and second by John Bell [9].

1.4.3 Projective measurement

In this section, we explain an important particular case of measurement. In fact, this method of projective measurement is equivalent to Postulate 2, when they are combined with the capacity to perform unitary operations. It has huge relevancy in quantum computing and it is explained here because of its relation with composite systems. Let us state the alternative postulate:

Projective Measurement. A projective measurement is described by an *observable*, M , a Hermitian operator on the state space of the system being described. The observable has a spectral decomposition:

$$M = \sum_m m P_m$$

where P_m is the projector onto the eigenspace with eigenvalue m . The possible outcomes of the measurement correspond to the eigenvalues, m , of the observable. Upon measuring the state $|\varphi\rangle$, the probability of the result m occurring is:

$$p(m) = \langle \varphi | P_m | \varphi \rangle$$

and the state of the system after the measurement, given that outcome m occurred, is:

$$\frac{P_m |\varphi\rangle}{\sqrt{p(m)}}$$

Projective measurements can be seen as a particular case of Postulate 2. Suppose the measurement operators from such postulate, in addition to satisfying the completeness relation $\sum_m M_m M_m^\dagger = I$, also fulfill that M_m are *orthogonal projectors*. That is, M_m is Hermitian and $M_m M_{m'} = \delta_{mm'} M_m \forall m, m'$. With these additional restrictions it holds that $M_m M_m^\dagger = M_m M_m = M_m$. By choosing $P_m = M_m$ we see that projective measurements is a particular case of Postulate 2.

Thanks to the way projective measurements are expressed, they have nice and handy properties. Let us compute the average value of a projective measurement:

$$\begin{aligned} E(M) &= \sum_m m p(m) \\ &= \sum_m m \langle \varphi | P_m | \varphi \rangle \\ &= \langle \varphi | \left(\sum_m m P_m \right) | \varphi \rangle \\ &= \langle \varphi | M | \varphi \rangle \end{aligned} \tag{1}$$

The average value of an operator is usually written as $\langle M \rangle \equiv \langle \varphi | M | \varphi \rangle$. The standard deviation associated to observations of M is then written as:

$$\begin{aligned} [\Delta(M)]^2 &= \langle (M - \langle M \rangle)^2 \rangle \\ &= \langle M^2 \rangle - \langle M \rangle^2 \end{aligned} \tag{2}$$

This means that if we prepare our state $|\varphi\rangle$ multiple times and measure it, the measures will follow a normal distribution with expected value $\langle M \rangle$ and standard deviation $\Delta(M) = \sqrt{\langle M^2 \rangle - \langle M \rangle^2}$. This formulation of standard deviation using observables

provides us with an elegant proof of perhaps the most famous quantum mechanics result: The *Heisenberg uncertainty principle*. Such proof can be found in annex [TODOref anexo con la demostracion, copy from Box 2.4 in Niel.].

There are two additional notations worth mentioning regarding projective measurement. Instead of providing an observable M , we will sometimes provide a complete set of orthogonal projectors. That is, a set P_m such that $\sum_i P_m = I$ and $P_m P_{m'} = \delta_{mm'} P_m$. The corresponding implicit observable used in this case is $M = \sum_m m P_m$. Finally, another widely used notation is to 'measure in a basis $|m\rangle$ ', where $|m\rangle$ form an orthonormal basis. This simply means that the chosen projectors are $P_m = |m\rangle\langle m|$ and the observable is $M = \sum_m m |m\rangle\langle m|$. This is the usual notation in quantum computing.

1.4.4 Quantum Computing perspective: Multiple qubits

Suppose we have a pair of qubits. In the classical case, two bits can be in four possible states: 00, 01, 10, and 11. Similarly, the two qubits computational basis states are $|00\rangle$, $|01\rangle$, $|10\rangle$ and $|11\rangle$. We arrive at this natural notation by knowing that each qubit has \mathbb{C}^2 as their associated state space and using the tensor product. Just like in the single qubit case, our two qubits system may be in a superposition of these four states:

$$|\varphi\rangle = \alpha_{00}|00\rangle + \alpha_{01}|01\rangle + \alpha_{10}|10\rangle + \alpha_{11}|11\rangle$$

Correspondingly, the measurement of this system will result in either 00, 01, 10 or 11 since we are using the observable $\sum_{x \in \{0,1\}^2} x|x\rangle\langle x|$, where $\{0,1\}^2$ are the strings of length two where each character is either 0 or 1. In fact, it will yield state x with probability $|\alpha_x|^2$, being α_x the coefficient associated with the state $|x\rangle$. The condition of the probabilities adding up to one is also called the *normalization condition* and can be expressed as $\sum_{x \in \{0,1\}^2} |\alpha_x|^2 = 1$ for the two qubits case.

The fundamental differences with the single qubit case start on measurement. Of course, we can measure both qubits at the same time, but we could also measure only one of them. In order to achieve this we can use the projectors $P_0 = |00\rangle\langle 00| + |01\rangle\langle 01|$ and $P_1 = |10\rangle\langle 10| + |11\rangle\langle 11|$. Upon measuring, the system will collapse to $P_0|\varphi\rangle / \sqrt{p_0}$ with probability:

$$p_0 = \langle \varphi | P_0 | \varphi \rangle = |\alpha_{00}|^2 + |\alpha_{01}|^2$$

Since these are the coefficients associated with the first qubit being 0, this is the probability of first qubit being 0. Furthermore, our system will collapse to:

$$|\varphi'\rangle = \frac{\alpha_{00}|00\rangle + \alpha_{01}|01\rangle}{\sqrt{|\alpha_{00}|^2 + |\alpha_{01}|^2}}$$

Note the normalization term $\sqrt{|\alpha_{00}|^2 + |\alpha_{01}|^2}$, which appears so the post-measurement state still satisfies the normalization condition. Naturally, after obtaining 0 in the first qubit we can still obtain either 0 or 1 in the second qubit, with probabilities

$$\frac{|\alpha_{00}|^2}{|\alpha_{00}|^2 + |\alpha_{01}|^2} \text{ and } \frac{|\alpha_{01}|^2}{|\alpha_{00}|^2 + |\alpha_{01}|^2}$$

respectively, adding up to 1. Correspondingly, the first qubit being measured will yield 1 with probability $p_1 = |\alpha_{10}|^2 + |\alpha_{11}|^2$.

Additionally, the first qubit independently should satisfy the normalization condition. That is, its probabilities of being 0 and 1 upon measurement must add up to 1. But those are p_0 and p_1 , which add up to one because of the normalization condition for $|\varphi\rangle$, as expected.

Let us come back to the previously introduced *Bell State*:

$$|+\rangle = \frac{|00\rangle + |11\rangle}{\sqrt{2}}$$

When measuring using the previously defined projectors P_0 and P_1 we obtain that the first qubit will collapse to either 0 or 1 with equal probability $p_0 = p_1 = 1/2$. Most importantly, in the first case the qubit will collapse to the state:

$$\frac{|00\rangle / \sqrt{2}}{\sqrt{p_0}} = |00\rangle$$

Meaning that the second qubit also collapsed to 0. Note that we only affected the first qubit and the second one was disturbed. These are the consequences of entanglement, key in the appearance of impressive phenomenons and discussions such as the EPR Paradox [10], [9].

Let us finally consider the more general case. In an n-qubits system our computational (orthonormal) basis would consist of the states $|x_1x_2\dots x_n\rangle$, where $x_i \in \{0,1\}$. As we already know, in a single qubit system we have two amplitudes α_0 and α_1 . We have four amplitudes for a 2-qubits system, eight for a 3-qubits system... And 2^n for an n-qubits system. This means that the number of amplitudes grows exponentially as we add qubits to the system. An immense increment compared to the classical case where the quantity of information that our system holds grows linearly with the numbers of bits. Of course, it is not that simple, since there are huge limitations on how we may access this information in the quantum realm such as how a qubit collapses upon measurement and the no-cloning theorem. However, we can already glimpse the power of quantum computing versus the classical one.

1.5 THE NO-CLONING THEOREM

A first exposure to the basis of quantum computing is fundamentally incomplete without the no-cloning theorem. It lets us understand one of the fundamental inconveniences of quantum computing: the impossibility of cloning arbitrary states.

We proceed by formalizing the copying mechanism. Let H be a state space and $|\varphi\rangle \in H$ the state to be copied. Given another state $|\varphi'\rangle \in H$, we would like to copy $|\varphi\rangle$ into $|\varphi'\rangle$. Since we can prepare $|\varphi'\rangle$ at will before the copying occurs, let $|\varphi'\rangle = |x_0\rangle$ be a fixed state. Then, our copying operation will take both states and produce the following result:

$$(|\varphi\rangle \otimes |x_0\rangle) \xrightarrow{U} |\varphi\rangle \otimes |\varphi\rangle$$

Operations between n-qubits (n-qubits gates) must preserve their norm, thus they must be unitary $n \times n$ matrices. In our case, U is a $2\dim(H) \times 2\dim(H)$ unitary matrix. Let us now state and prove the theorem.

Theorem 3 (No-cloning Theorem). *There is no unitary operator $U : H^{\otimes 2} \longrightarrow H^{\otimes 2}$ and state $|x_0\rangle \in H$ such that for any arbitrary state $|\varphi\rangle \in H$ it holds*

$$U|\varphi\rangle|x_0\rangle = |\varphi\rangle|\varphi\rangle$$

Proof. Suppose there exists such unitary operator U and such state $|x_0\rangle$. Let $|\varphi\rangle, |\psi\rangle \in H$. We may apply the copying operators to both of them:

$$U|\varphi\rangle|x_0\rangle = |\varphi\rangle|\varphi\rangle$$

$$U|\psi\rangle|x_0\rangle = |\psi\rangle|\psi\rangle$$

Taking the inner product of both equations and using that U is unitary results in:

$$\begin{aligned} \left(U|\varphi\rangle|x_0\rangle, U|\psi\rangle|x_0\rangle \right) &= \left(|\varphi\rangle|\varphi\rangle, |\psi\rangle|\psi\rangle \right) \iff \\ \left(|\varphi\rangle|x_0\rangle, |\psi\rangle|x_0\rangle \right) &= \langle\varphi|\psi\rangle\langle\varphi|\psi\rangle \iff \\ \langle\varphi|\psi\rangle\langle x_0|x_0\rangle &= \langle\varphi|\psi\rangle^2 \end{aligned}$$

Since $|x_0\rangle$ is a normalized vector, $\langle x_0|x_0\rangle = 1$, and we obtain:

$$\langle\varphi|\psi\rangle = \langle\varphi|\psi\rangle^2$$

Which only holds if $\langle \varphi | \psi \rangle$ equals to either 0 or 1, which means $|\varphi\rangle$ and $|\psi\rangle$ are either equal or orthonormal. But those vectors were arbitrary, thus such a general cloning operator is impossible. \square

2

THE D-WAVE QUANTUM COMPUTER

2.1 QUANTUM ANNEALING

Simulated annealing (SA) is a general optimization technique that was first introduced in 1983 by Kirkpatrick *et al.* [11]. The main idea is to mimic how thermal fluctuations work to let the system escape from local minimums in the cost function. The *temperature* of the system dictates the probability with which it is allowed to jump to worse solutions (higher values of the cost function). The *annealing schedule* is the decrease rate of the temperature, controlled by the programmer. Under the appropriate one, the system would be able to escape from local minimums to reach the global one. If the temperature decreases too quickly, the system converges prematurely to a local minimum. If it decreases too slowly, the technique transforms into a random walker, reaching the global minimum at some point but being worthless time-wise.

Inspiration in thermal fluctuations are used in SA so the system may escape from local minima. Similarly, *quantum tunneling* is used in Quantum Annealing (QA) in order to escape from local minimums. This new technique was introduced in its present form by two similar proposals [12] [13].

Quantum tunneling is a quantum mechanical phenomenon where a wavefunction (i.e. a quantum state) may propagate through a potential barrier [14]. The probability of this event occurring depends on the height and width of the barrier (see figure 2)

The main advantage of quantum tunneling compared to classical thermal fluctuations is that the probability of the system shifting depends not only on the height of the potential barrier but also on its width. Thus, application of QA may potentially outperform SA in problems where the energy (cost) landscape consists on very high and thin barriers surrounding local minimums.

Particularly, as studied in [15], let Δ be the height of a barrier and ω its width. The probability of a thermal transition occurring is dictated by $\exp\{-\frac{\Delta}{k_B T}\}$ where T is the temperature of the system and k_B is the Boltzmann constant. On the other hand, the probability of a quantum tunneling transition occurring through the same barrier

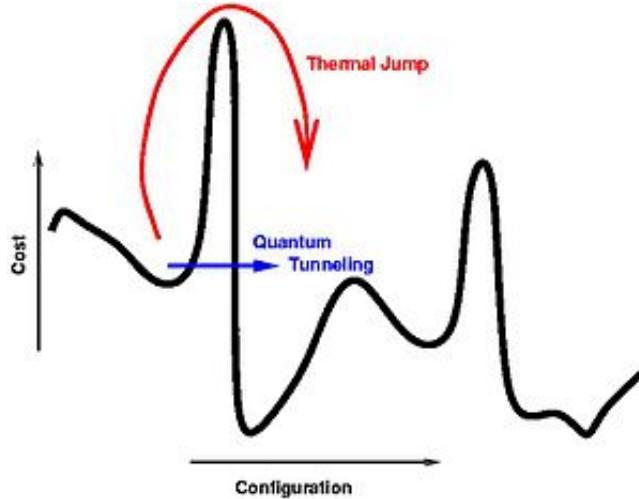


Figure 2: Thermal jumps vs quantum tunneling.

(assuming isolation) is $\exp\{-\frac{\sqrt{\Delta\omega}}{\Gamma}\}$, where Γ is the tunneling field. Thus, the probability is much higher for the quantum tunneling effect on a high and thin barriers.

We may also observe looking at the previous expressions how the tunneling field and the temperature play a similar roles in the different methods. Thus, the annealing schedule of the QA system is controlled using the tunneling field just like the temperature is used in SA.

There are many examples in the literature of QA simulations using classical computers, both theoretical [16] and numerical -most of them using Quantum Monte Carlo (QMC) methods [17] [18]-. This evidence suggests outperformance of SA by simulated QA under certain conditions. It is reasonable to wonder what happens if instead of simulating QA we could somehow encode our problems in a real physical system that tends to its lower energy state (ground state), thus using quantum tunneling naturally. This is precisely what the D-Wave system does.

Let us further deepen into the technicalities of quantum annealing in order to understand how it is implemented in the D-Wave system.

2.2 ADIABATIC EVOLUTION

The alternative form of the third postulate of quantum mechanics (section 1.3.2) stated that the evolution of a quantum system is described by the Schrodinger equation:

$$i\hbar \frac{d|\varphi\rangle}{dt} = H|\varphi\rangle$$

By better understanding the Hamiltonian H we may control this evolution. Suppose that the evolution of a given quantum system is described by a Hamiltonian $\hat{H}(t)$.

Suppose that at some initial time t_0 our quantum system is in an eigenstate $|\varphi(t_0)\rangle$ of $\hat{H}(t_0)$. Since the evolution is continuous, at some other time $t_1 > t_0$ we could expect our system to be in the corresponding eigenstate $|\varphi(t_1)\rangle$ of $\hat{H}(t_1)$. This fact critically depends on the time $\tau = t_1 - t_0$ during which the modification takes place, as stated by the adiabatic theorem, first proposed by Max Born and Vladimir Fock (1928) [19].

Theorem 4 (Adiabatic Theorem). *A physical system remains in its instantaneous eigenstate if a given perturbation is acting on it slowly enough and if there is a gap between the eigenvalue and the rest of the Hamiltonian's spectrum.*

Thus, we may define diabatic and adiabatic processes depending on how they adapt to the system changes [20].

Definition 13. A *diabatic process* is a process where rapidly changing conditions prevent the system from adapting its configuration during the process. Typically there is no eigenstate of the final Hamiltonian with the same functional form as the initial state. The system ends in a linear combination of the states.

Definition 14. An *adiabatic process* is a process where gradually changing conditions allow the system to adapt its configuration. If the system starts in an eigenstate of the initial Hamiltonian, it will end in the corresponding eigenstate of the final Hamiltonian.

The 'gap condition' that appears in theorem 4 refers to an additional requirement: the spectrum of $\hat{H}(t)$ is *nondegenerate*, meaning there are no two equal eigenvalues at fixed time t . This condition is also called the *no-crossing* condition, and allows us to sort the eigenstates using the eigenvalues without any ambiguity.

As a classical analogy, consider a simple pendulum. If the pendulum's support point is moved, the oscillation of the pendulum may change. In fact, violent changes on the support point will dramatically affect the pendulum's movement, changing it completely. However, if the support point is moved slowly enough, the pendulum will remain unchanged. These are examples of diabatic and adiabatic processes respectively.

2.2.1 Avoided crossing

Let us consider an important physical example known as the *avoided crossing*. Consider a two-state quantum system, i.e. a qubit. Suppose its Hamiltonian to be:

$$H = \begin{pmatrix} E_1 & 0 \\ 0 & E_2 \end{pmatrix}$$

Whose eigenvalues are E_1 and E_2 , and eigenvectors,

$$|0\rangle = \begin{pmatrix} 1 \\ 0 \end{pmatrix}, \quad |1\rangle = \begin{pmatrix} 0 \\ 1 \end{pmatrix}$$

which will be used as our base. Thus, a state vector describing the system may be written as a superposition of both states:

$$|\varphi\rangle = \alpha(t)|0\rangle + \beta(t)|1\rangle$$

Supposing adiabatic evolution, if the system is prepared in either one of the eigenstates it will remain as such as long as the gap condition is suffied: $E_1 \neq E_2$. However, if $E_1 = E_2$, any superposition of states will be an eigenstate, thus remaining unchanged. Hence, independently of E_1 and E_2 , any system prepared in an eigenstate will remain as such, supposed adiabatic evolution.

Let us consider a P perturbation into our original system. For simplicity we only consider perturbations with degenerated diagonal. Our new Hamiltonian will be the following:

$$H' = H + P = \begin{pmatrix} E_1 & 0 \\ 0 & E_2 \end{pmatrix} + \begin{pmatrix} 0 & \omega \\ \bar{\omega} & 0 \end{pmatrix} = \begin{pmatrix} E_1 & \omega \\ \bar{\omega} & E_2 \end{pmatrix}$$

where $\omega \in \mathbb{C}$. Note that by setting ω , the other value in the antidiagonal is fixed since H' must be Hermitian. By using the characteristic equation we may compute the eigenvalues of H' :

$$\begin{aligned} E_+ &= \frac{1}{2}(E_1 + E_2) + \frac{1}{2}\sqrt{(E_1 - E_2)^2 + 4|\omega|^2} \\ E_- &= \frac{1}{2}(E_1 + E_2) - \frac{1}{2}\sqrt{(E_1 - E_2)^2 + 4|\omega|^2} \end{aligned}$$

A continuación quiero añadir una imagen, seguramente como la de [wikipedia](#) (figura 3):

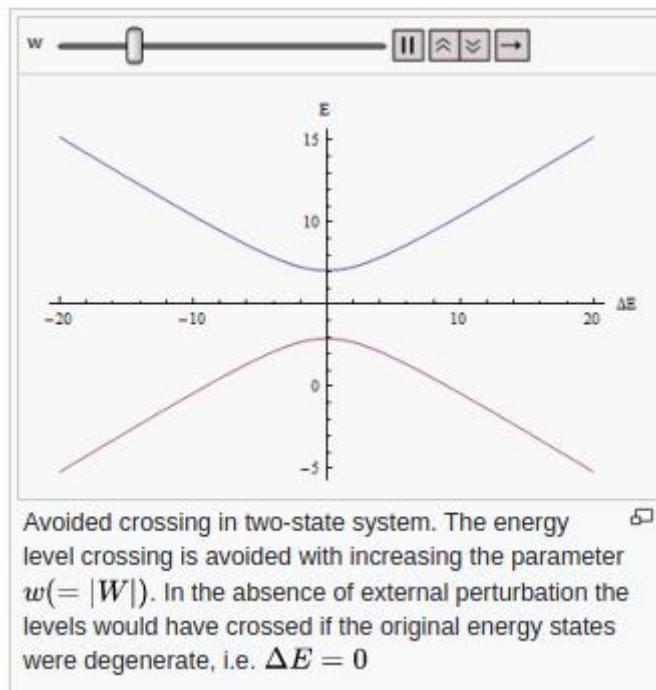


Figure 3: Figura de wikipedia

He hecho una versión propia en Geogebra, tomando $E_1 = -E_2$ (luego $E_1 + E_2 = 0$ y $E_1 - E_2 = 2E$, he llamado $x \equiv 2E$). En la simulación, podemos ajustar el valor de w , que he llamado a utilizando el slider, reproduciendo el efecto del gif de wikipedia (figura 3):

<https://www.geogebra.org/graphing/svkru3s>

Sin embargo, me salen las ramas de las hiperbolas giradas 90 grados. Esperaba que la rama superior fuese E_+ y la inferior E_- pero cada rama tiene parte de las dos. ¿Cuál es mi fallo al pintar las gráficas?

2.2.2 Solving problems using adiabatic evolution

In order to use adiabatic evolution to solve optimization problems we will configure our own Hamiltonian for the system. In particular, the Hamiltonian will be a linear combination of two Hamiltonians:

$$H(t) = f(t) \cdot H_{initial} + g(t) \cdot H_{final} \quad \forall t \in [t_{initial}, t_{final}]$$

where $f, g : [t_{initial}, t_{final}] \rightarrow \mathbb{R}_0^+$. The initial Hamiltonian, $H_{initial}$, will be an easy configuration so we may prepare our system to start in the ground state of $H_{initial}$.

The final Hamiltonian, H_{final} , will encapsulate the cost function of our problem such that the minimum value is reached in the ground state.

Functions f and g fulfil that $g(t_{initial}) = f(t_{final}) = 0$. When times elapses and supposing adiabatic evolution, the system will stay on the ground state of H and by $t = t_{final}$, the system will be in the H_{final} 's ground state, codifying the minimum of our cost function.

However, we cannot always grant the necessary conditions for perfect adiabatic evolution. In some cases, the gap condition does not suffice at the system will jump with certain probability to other eigenstates, providing suboptimal solutions but still useful ones. This depends of the eigenvalues of the final Hamiltonian, which at the same time entirely depends on the problem being studied.

The only question remaining is how to encode our cost function in a Hamiltonian.

2.3 QUBO AND ISING MODELS

The QUBO and Ising problems can be easily represented with a Hamiltonian. In practice, non-optimization problems are transformed into optimization problems and then codified as either QUBO or Ising to be solved using quantum annealing. Let us explore this type of problems and some examples of these transformations.

Let $B = \{0, 1\}$ and $f : \mathbb{B} \rightarrow \mathbb{R}$ be a quadratic polynomial over binary variables:

$$f_Q(x) = \sum_{i=1}^n \sum_{j=1}^i q_{ij} x_i x_j$$

where $x_i \in \mathbb{B}$ for $i \in \{1, \dots, n\}$ and the coefficients $q_{ij} \in \mathbb{R}$ for $1 \leq j \leq i \leq n$. A **quadratic unconstrained binary optimization (QUBO) problem** consists of finding a binary vector x' such that x' is a minimum of f :

$$x' = \arg \min_{x \in \mathbb{B}^n} f(x)$$

QUBO problems can be formulated in its more compact matrix form:

$$f_Q(x) = x^T Q x$$

where Q is a $n \times n$ symmetric matrix containing q_{ii} and its diagonal and $q_{ij}/2$ in position (i, j) where $i \neq j$. Let us explore some simple properties of QUBO problems:

- Multiplying the coefficients q_{ij} by a constant $a \in \mathbb{R}^+$ scales the output by exactly a . Thus, x' remains the minimum:

$$f_{aQ}(x) = \sum_{i < j} (aq_{ij})x_i x_j = a \sum_{i < j} q_{ij}x_i x_j = af_Q(x)$$

- Flipping the sign of the coefficients flips the sign of $f_Q(x)$. Thus, x' is the maximum of $f_{-Q}(x)$

$$f_{-Q}(x) = \sum_{i < j} (-q_{ij})x_i x_j = - \sum_{i < j} q_{ij}x_i x_j = -f_Q(x)$$

- If all coefficients are positive the minimum is $x = (0, \dots, 0)$. Analogously, if all coefficients are negative, the minimum is $x = (1, \dots, 1)$.

The **Ising models**, on the other hand, find their inspiration in physics. They are named after the physicist Ernst Ising who solved the one-dimensional model in his 1924 thesis [21]. They are stated using a Hamiltonian function $H : \{-1, 1\}^n \rightarrow \mathbb{R}$:

$$H(\sigma) = - \sum_{\langle i j \rangle} J_{ij}\sigma_i\sigma_j - \mu \sum_j h_j\sigma_j$$

with parameters $h_j, J_{ij}, \mu \in \mathbb{R}$. The *spin variables* σ_i are in $\{-1, 1\}$ instead of \mathbb{B} . Moreover, the spin variables are arranged in a graph, typically a *lattice*, where the local structure repeats periodically. Only pair of variables $\langle i j \rangle$ which are neighbor nodes in the graph may have non-zero coefficients J_{ij} . Let us see the connection between the QUBO and Ising models by using the identity $\sigma \mapsto 2x - 1$:

$$\begin{aligned} f(x) &= \sum_{\langle i j \rangle} -J_{ij}(2x_i - 1)(2x_j - 1) - \sum_j \mu h_j(2x_j - 1) \\ &= \sum_{\langle i j \rangle} -4J_{ij}x_i x_j + 2J_{ij}x_i + 2J_{ij}x_j - J_{ij} + \sum_j 2\mu h_j x_j - \mu h_j \quad \text{using } x_j = x_i x_j \\ &= \sum_{i=1}^n \sum_{j=1}^i q_{ij}x_i x_j + C \end{aligned}$$

where

$$q_{ij} = \begin{cases} -4J_{ij} & \text{if } i \neq j \\ \sum_{\langle k i \rangle} 2J_{ki} + \sum_{\langle i l \rangle} 2J_{il} + 2\mu h_i & \text{if } i = j \end{cases}$$

$$C = - \sum_{\langle i j \rangle} J_{ij} - \mu \sum_j h_j$$

Since adding a constant does not change the minimum x' , it can be omitted during optimization and only be used for transforming one type of problem into another.

In practice, a given problem would be translated into either QUBO or Ising. It is useful to learn both type of problems since they both are popular among the literature. Let us deepen it the translation of different NP problems. The resolution of QUBO and Ising models will not be study, but only the translations into these models. In particular, these examples of QUBO transformations are mainly from [22] and [23], unless stated otherwise. For an extense literature review on QUBO and Ising translations of NP problems refer to [24].

2.3.1 The Max Cut problem

One of the first famous problems that was encoded as a QUBO was the max-cut problem due its natural translation. It reads as follows: Given an undirected graph $G(V, E)$ with vertex set V and edge set E , seek a partition of V such that the number of edges between both sets is a large as possible.

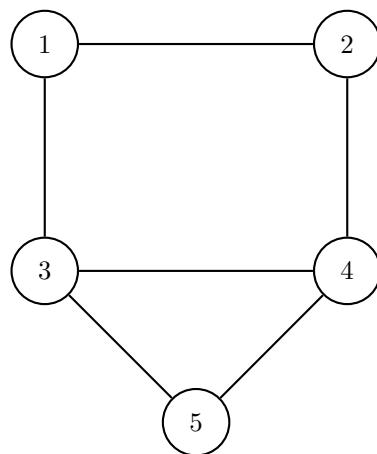
The max-cut problem can be modeled by setting variable $x_j = 1$ if vertex j is in one set and $x_j = 0$ if it is in the other one, for every vertex j in V . The quantity $(x_i + x_j - 2x_i x_j)$ marks whether the edge (i, j) is in the cut. That is, $(x_i + x_j - 2x_i x_j) = 1$ if the vertices x_i and x_j are in different sets, and 0 otherwise.

Thus, maximazing the number of edges in the cut is encoded as a QUBO as:

$$\text{Maximize } f(x) = \sum_{(i,j) \in E} (x_i + x_j - 2x_i x_j)$$

NUMERICAL EXAMPLE

Consider the following graph:



We would have five variables: x_1, \dots, x_5 . The cost associated to edge $(1, 2)$ is $x_1 + x_2 - 2x_1x_2$. By repeating this process for every edge in the graph we obtain the cost function:

$$\begin{aligned} \text{Maximize } f(x) = & (x_1 + x_2 - 2x_1x_2) + (x_1 + x_3 - 2x_1x_3) + (x_2 + x_4 - 2x_2x_4) \\ & + (x_3 + x_4 - 2x_3x_4) + (x_3 + x_5 - 2x_3x_5) + (x_4 + x_5 - 2x_4x_5) \end{aligned}$$

or equivalently:

$$\begin{aligned} \text{Maximize } f(x) = & 2x_1 + 2x_2 + 3x_3 + 3x_4 + 2x_5 \\ & - 2x_1x_2 - 2x_1x_3 - 2x_2x_4 - 2x_3x_4 - 2x_3x_5 - 2x_4x_5 \end{aligned}$$

We can transform the problem using the following symmetric matrix:

$$Q = \begin{pmatrix} 2 & -1 & -1 & 0 & 0 \\ -1 & 2 & 0 & -1 & 0 \\ -1 & 0 & 3 & -1 & -1 \\ 0 & -1 & -1 & 3 & -1 \\ 0 & 0 & -1 & -1 & 2 \end{pmatrix}$$

Yielding the following compact form:

$$\text{Maximize } f_Q(x) = x^T Q x$$

The minimum value of $f_Q(x)$ is $x = (0, 1, 1, 0, 0)$. Thus, vertices 2 and 3 are in one set and vertices 1, 4 and 5 are in the other one.

2.3.2 General translation method

The previous problem was naturally transform into QUBO due to its particular constraints. However, can any quadratic optimization problem be encoded as a QUBO?

The answer is positive. Furthermore, any quadratic optimization problem with linear constraints may be transformed into a QUBO. Let us first study a simple case. Consider the following problem:

$$\begin{aligned} \text{Minimize } f(x_1, x_2) \\ \text{subject to the constraint} \\ x_1 + x_2 \leq 1 \end{aligned}$$

where $x_1, x_2 \in \mathbb{B}$. We may transform our linear constraint into a penalty $P(x_1 x_2)$ for some positive scalar P . This term equals 0 if and only if our constraint is satisfied by the variables. Thus, by adding the penalty to our cost function, for P large enough, the problem

$$\text{Minimize } f(x_1, x_2) + Px_1 x_2$$

will minimize $f(x_1, x_2)$ and, at the same time, satisfy the constraint. Namely, the constraint translates into adding a non-negative value to the cost function. Some other useful translations can be found in table 1. In general, if the cost function is being minimized we may substitute constraints by non-negative terms that only equal zero if they are satisfied, multiplied by a large-enough positive value P . Thus, we may remove the associated constraints by simply adding these terms to the cost function.

Classical Constraint	Equivalent Penalty
$x + y \leq 1$	$P(xy)$
$x + y \geq 1$	$P(1 - x - y + xy)$
$x + y = 1$	$P(1 - x - y + 2xy)$
$x \leq y$	$P(x - xy)$
$x_1 + x_2 + x_3 \leq 1$	$P(x_1 x_2 + x_1 x_3 + x_2 x_3)$
$x = y$	$P(x + y - 2xy)$

Table 1: Constraints and penalties relations

Specifically, let us study the more general case:

$$\text{Minimize } f(x) = x^T C x$$

subject to the constraints

$$Ax = b$$

where $x \in \mathbb{B}^n$, k is the number of constraints, $b \in \mathbb{R}^k$, C is a $n \times n$ matrix and A is a $k \times n$ matrix. This accommodates to both linear and quadratic cost function since elements in the diagonal are $c_{ii}x_i^2 = c_{ii}x_i$. For a positive scalar P , the quadratic penalty $P(Ax - b)^T(Ax - b)$ is added to the cost function to obtain:

$$\begin{aligned} f'(x) &= x^T C x + P(Ax - b)^T(Ax - b) \\ &= x^T C x + P(x^T A^T A x - x^T A^T b - b^T A x + b^T b) \\ &= x^T C x + x^T D x + c \\ &= x^T Q x + c \end{aligned}$$

where $D = P(A^T A)$, $c = P(-x^T A^T b - b^T A x + b^T b)$ and $Q = C + D$. Since the constant c does not affect where the minimum is reached, f and f' share this minimum. Therefore, we may simply solve:

$$\text{Minimize } f(x) = x^T Q x$$

which is a QUBO problem.

In this development we only considered equality constraints. There is also a general method to transform inequality constraints into equality ones by introducing *slack variables* [25]. In practice, the general case will not be necessary since the constraints are usually quite simple.

Finally, we may also consider satisfiability problems. That is, given a set of constraints for the variables, find a configuration of them such that all the constraints are satisfied. If the constraints are linear, these problems may be transformed into QUBO by setting the cost function as constant and following the same method. Suppose the constraints over x can be written as $Ax = b$. The equivalent optimization problem is written as:

$$\begin{aligned} \text{Minimize } f(x) &= 0 \\ \text{subject to the constraints} \\ Ax &= b \end{aligned}$$

Any solution satisfying the constraint will yield 0 on the penalties, thus minimizing the transformed cost function $f'(x) = x^T Q x$.

Let us explore some more complicated examples to expand this ideas.

2.3.3 Satisfiability problems: Max 2-SAT

This example is based on [22] and [18].

In the previous section we mentioned satisfiability problems, also known simply as SAT. In this section we study these problems from its traditional logic perspective.

A *clause* is statement depending on a finite set of binary variables, $C = C(x_1, \dots, x_k)$, that be either True or False. A *literal* is a variable or its negation: x and $\neg x$. Let C_1, \dots, C_n be a set of *Clauses* depending on variables x_1, \dots, x_m . The satisfiability problem associated to this problem is to find a configuration for the variables x_1, \dots, x_m such that every clause is true:

$$C_1 \wedge \dots \wedge C_n$$

SAT was the first problem proven to be NP complete. This result is known as the Cook-Levin theorem, named after Stephen Cook and Leonid Levin [26]. Its relevance in computer science is astronomical. There has been many variations of SAT problems. Let us explore the Max 2-SAT variation and transform it into a QUBO model.

Max 2-SAT problem is not an optimization problem. Its main restriction is that each clause may depend on at most two literals and it is true if and only if at least one literal is true. The problem consist on finding a configuration of variables such that the number of true clauses is maximized. Each clause $C(x_C, y_C)$ is associated with an energy function depending on the variables (x_C, y_C) :

$$E_C(x_C, y_C) = \begin{cases} 0, & \text{if } (x_C, y_C) \text{ satisfies clause } C \\ 1, & \text{if } (x_C, y_C) \text{ violates clause } C \end{cases}$$

The energy of the whole system, which will work as our cost function, is defined as the sum of all the energies:

$$E = \sum_C E_C$$

Clearly $E \geq 0$ and $E = 0$ if and only if every clause is true. Furthermore, the minimum energy of the system E_0 will be reached where the maximum number of clauses are satisfied: exactly every clause but E_0 will be true. Note that this perspective does not depend on the Max 2-SAT problem. In fact, the same process is used to transform other satisfiability problems.

The remaining question is how to transform our clauses into energy functions. For the Max 2-SAT problem this is quite straight forward. Each clause may depend on at most two literals. Thus, there are either zero, one or two negations in each clause. We may study this cases separately, as seen in table 2.

Clause	Traditional constraint	Energy function
$x \vee y$	$x + y \geq 1$	$(1 - x - y + xy)$
$\neg x \vee y$	$\neg x + y \geq 1$	$(x - xy)$
$\neg x \vee \neg y$	$\neg x + \neg y \geq 1$	(xy)

Table 2: Clause and energy functions relations

Since every energy function is quadratic, their will also be so. This is the cost function used in our QUBO model:

$$\text{Minimize } E(x)$$

Note that, interestingly enough, our QUBO model does not depend on the number of clauses but only in the number of variables. This means that a problems with 30,000 clauses and 200 variables will only use 200 variables in the associated QUBO model.

NUMERICAL EXAMPLE

Given the following the set of clauses found in table 3, find a solution to the Max 2-SAT problem associated to them.

Clause #	Clause	Energy function
1	$x_1 \vee x_2$	$(1 - x_1 - x_2 + x_1x_2)$
2	$x_1 \vee \neg x_2$	$(x_2 - x_1x_2)$
3	$\neg x_1 \vee x_2$	$(x_1 - x_1x_2)$
4	$\neg x_1 \vee \neg x_2$	(x_1x_2)
5	$\neg x_1 \vee x_3$	$(x_1 - x_1x_3)$
6	$\neg x_1 \vee \neg x_3$	(x_1x_3)
7	$x_2 \vee \neg x_3$	$(x_3 - x_2x_3)$
8	$\neg x_2 \vee x_3$	$(x_2 - x_2x_3)$
9	$\neg x_2 \vee \neg x_3$	(x_2x_3)
10	$x_2 \vee x_4$	$(1 - x_2 - x_4 + x_2x_4)$
11	$x_3 \vee x_4$	$(1 - x_3 - x_4 + x_3x_4)$
12	$\neg x_3 \vee \neg x_4$	(x_3x_4)

Table 3: Clause and energy functions relations

A quick look at the first four clauses tells us that we cannot possibly find a configuration of variables that satisfy every clause. In fact, at most three of those four will be true. Adding the energies associated to every clause gives us the energy of the system, and therefore our QUBO model:

$$\text{Minimize } E(x) = 3 + x_1 - 2x_4 - x_2x_3 + x_2x_4 + 2x_3x_4$$

or equivalently:

$$\text{Minimize } E(x) = 3 + x^T Q x$$

where

$$Q = \begin{pmatrix} 1 & 0 & 0 & 0 \\ 0 & 0 & -1/2 & 1/2 \\ 0 & -1/2 & 0 & 1 \\ 0 & 1/2 & 1 & -2 \end{pmatrix}$$

Solving this QUBO gives $x_1 = x_2 = x_3 = 0$ and $x_4 = 1$ with $E(0, 0, 0, 1) = 1$, meaning that all clauses but one are satisfied.

As a final remark, this method of solving Max 2-SAT problems has proven to be computationally useful up to hundreds of variables of thousands of clauses [27].

2.3.4 Graph Coloring

Consider the following problem: Given a graph G and K -colors, assign a single color to each vertex in V such that no two adjacent vertices have the same color. Let x_{ij} be a binary variable that is set to 1 if and only if the i -th node is assigned the color j -th. We may develop the following two constraints. First, since all nodes must be colored exactly once:

$$\sum_{j=1}^K x_{ij} = 1 \quad \forall i \in \{1, \dots, n\}$$

where n is the number of nodes. Moreover, the condition that two adjacent nodes cannot have the same color is translated into:

$$x_{ic} + x_{jc} \leq 1 \quad \forall c \in \{1, \dots, K\}$$

for every pair of adjacent nodes (i, j) in G .

Thus, by finding a vector configuration for the variables x_{ij} that satisfy this constraint, we find a solution of the graph coloring problem. This is what we previously introduced as a satisfiability problem. Therefore, it may be written as:

$$\text{Minimize } f(x) = 0$$

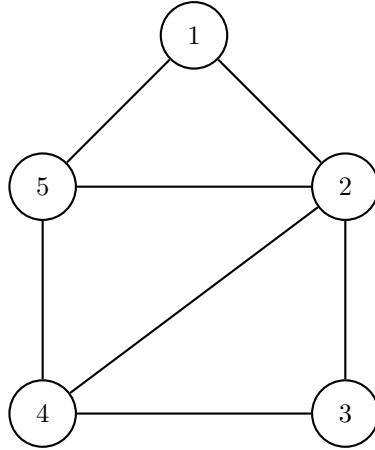
subject to the constraints

$$\sum_{j=1}^K x_{ij} = 1 \quad \forall i \in \{1, \dots, n\} \quad \text{and}$$

$$x_{ic} + x_{jc} \leq 1 \quad \forall c \in \{1, \dots, K\} \quad \text{for all adjacent nodes } i \text{ and } j$$

NUMERICAL EXAMPLE

Given the following graph, find a coloring using three colors.



For this particular graph there are 15 variables x_{ij} for all $i \in \{1, \dots, 5\}$ and for all $j \in \{1, 2, 3\}$. The associated constraints are:

$$x_{i1} + x_{i2} + x_{i3} = 1 \quad \forall i \in \{1, \dots, 5\}$$

$$x_{ic} + x_{jc} \leq 1 \quad \forall c \in \{1, 2, 3\} \quad \text{for all adjacent nodes } i \text{ and } j$$

a total of 26: 5 for assigning one and only one color to each node, and 3 additional ones per edge in the graph (7) for no adjacent equally colored nodes. First, let us rename our variables from 1 to 15 as follows:

$$(x_{11}, x_{12}, x_{13}, x_{21}, x_{22}, x_{23}, x_{31}, \dots, x_{52}, x_{53}) = (x_1, x_2, x_3, x_4, x_5, x_6, x_7, \dots, x_{15}, x_{16})$$

The first set of constraints may be substituted by $P(x_{i1} + x_{i2} + x_{i3} - 1)^2$, which only equals 0 if the constraint is satisfied. For instance, for the first node $P(x_1 + x_2 + x_3 - 1)^2$ is transformed to $P(-x_1 - x_2 - x_3 + 2x_1x_2 + 2x_1x_3 + 2x_2x_3) + P$. By choosing a value for $P = 4$ and dropping the constant we can insert these penalties into an under developed Q matrix:

$$Q = \begin{pmatrix} -4 & 4 & 4 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 4 & -4 & 4 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 4 & 4 & -4 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & -4 & 4 & 4 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 4 & -4 & 4 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 4 & 4 & -4 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & -4 & 4 & 4 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 4 & -4 & 4 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 4 & 4 & -4 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & -4 & 4 & 4 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 4 & -4 & 4 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 4 & -4 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & -4 & 4 & 4 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 4 & -4 & 4 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 4 & 4 \end{pmatrix}$$

Secondly, in order to transform $x + y \leq 1$ we can refer to table 1 and use $P(xy)$. We add these penalties to the previous matrix in the proper places:

$$Q = \begin{pmatrix} -4 & 4 & 4 & 2 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 2 & 0 & 0 \\ 4 & -4 & 4 & 0 & 2 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 2 \\ 4 & 4 & -4 & 0 & 0 & 2 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 2 \\ 2 & 0 & 0 & -4 & 4 & 4 & 2 & 0 & 0 & 2 & 0 & 0 & 2 & 0 & 0 \\ 0 & 2 & 0 & 4 & -4 & 4 & 0 & 2 & 0 & 0 & 2 & 0 & 0 & 2 & 0 \\ 0 & 0 & 2 & 4 & 4 & -4 & 0 & 0 & 2 & 0 & 0 & 2 & 0 & 0 & 2 \\ 0 & 0 & 0 & 2 & 0 & 0 & -4 & 4 & 4 & 2 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 2 & 0 & 4 & -4 & 4 & 0 & 2 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 2 & 4 & 4 & -4 & 0 & 0 & 2 & 0 & 0 & 0 \\ 0 & 0 & 0 & 2 & 0 & 0 & 2 & 0 & 0 & -4 & 4 & 4 & 2 & 0 & 0 \\ 0 & 0 & 0 & 0 & 2 & 0 & 0 & 2 & 0 & 4 & -4 & 4 & 0 & 2 & 0 \\ 0 & 0 & 0 & 0 & 0 & 2 & 0 & 0 & 2 & 4 & 4 & -4 & 0 & 0 & 2 \\ 2 & 0 & 0 & 2 & 0 & 0 & 0 & 0 & 0 & 2 & 0 & 0 & -4 & 4 & 4 \\ 0 & 2 & 0 & 0 & 2 & 0 & 0 & 0 & 0 & 0 & 2 & 0 & 4 & -4 & 4 \\ 0 & 0 & 2 & 0 & 0 & 2 & 0 & 0 & 0 & 0 & 0 & 2 & 4 & 4 & -4 \end{pmatrix}$$

Since the matrix above incorporate all the constraints, finding a graph coloring is equivalent to finding a minimum of:

$$\text{Minimize } f(x) = x^T Qx$$

Solving this model yields $x_2 = x_4 = x_9 = x_{11} = x_{15} = 1$, and 0 for the rest of the variables. Meaning, nodes 1 and 4 are assigned color number 2, node 2 is assigned color number 1, and nodes 3 and 5 are assigned color number 3.

As a final remark, this method of graph coloring has proven to be computationally useful for graphs with up to 450 nodes and 9757 edges [27].

2.3.5 The Travelling Salesman problem

This example is mainly based on [22].

In order to formulate our next problem, a couple of extra definitions are needed. Given a graph, a *Hamiltonian path* is a path -a sequence of connected nodes- such that it only passes through each node exactly once. A *cycle*, also called *tour*, is a path that starts and ends in the same node. A *Hamiltonian cycle* is a cycle that passes through each node a single time, except for the first/last node which appears twice in the cycle. Additionally, a weighted graph $G = (V, E, W)$ is a graph with a series of weights $w_e \in \mathbb{R}$ associated to each edge $e \in E$.

The Travelling Salesman problem is the following: Given a directed weighted graph, find a Hamiltonian cycle with minimum total weight. The problem is sometimes stated with a Hamiltonian path instead of Hamiltonian cycle, but as it will be seen in section [TODOref], for our purposes the cycle-statement will be of better use.

We may suppose without lossing generality that the graph is complete. That is, there is always an edge from every two nodes in every direction. If the given graph is not complete, we may add the remaining edges with a high enough weight so that they are never used.

Since solutions to this problem are a series of $n \equiv |V|$ nodes -the number of nodes in the graph-, we will encode n -nodes cycles as $(v_{k_0}, \dots, v_{k_{n-1}})$, where nodes v_{k_0} and $v_{k_{n-1}}$ are also connected. Thus, our problem may originally be stated as:

$$\text{Minimize } \sum_{i=0}^{n-1} w_{k_i, k_{i+1}}$$

subject to the constraint that $(v_{k_0}, \dots, v_{k_{n-1}})$ is a Hamiltonian cycle

where the $k_n \equiv k_0$. The binary variables $x_{i,p}$ will be used, where $i, p \in \{0, \dots, n - 1\}$. $x_{i,p}$ will equal 1 if and only if node i is in position p in our cycle. Thus, we will have n^2 variables under the following constraints:

1. Every node must be assigned. Thus, node assignment will be rewarded with a non-positive weight (favorable bias) called **self-bias** with an associated non-positive penalty:

$$ax_{i,p} \quad \forall i, p \in \{0, \dots, n - 1\}$$

where $a \leq 0$. In practice, this will not be necessary for every problem. We may remove the self-bias by setting $a = 0$. Adding these penalties for every node and position yields:

$$a \sum_{i=0}^{n-1} \sum_{p=0}^{n-1} x_{i,p} \tag{3}$$

2. Every node must be assigned to only one position. This constraint will be called **repetition**:

$$\sum_{p=0}^{n-1} x_{i,p} = 1 \quad \forall p \in \{0, \dots, n - 1\}$$

Resulting in the following penalty:

$$b \left(\sum_{p=0}^{n-1} x_{i,p} - 1 \right)^2 \quad \forall p \in \{0, \dots, n - 1\}$$

where b is a positive parameter. Adding these n penalties we obtain:

$$b \sum_{i=0}^{n-1} \left(\sum_{p=0}^{n-1} x_{i,p} - 1 \right)^2 \tag{4}$$

3. Every position must be assigned only one node. This constraint will be called **colocation**:

$$\sum_{i=0}^{n-1} x_{i,p} = 1 \quad \forall i \in \{0, \dots, n - 1\}$$

Resulting in the following penalty:

$$c \sum_{p=0}^{n-1} \left(\sum_{i=0}^{n-1} x_{i,p} - 1 \right)^2 \tag{5}$$

where c is a positive parameter.

Finally, we may re-write the cost function in terms of the binary variables. Weight w_{ij} must be added in the cost function if and only if nodes i and j follow each other in the path. That is, if and only if $x_{i,p}x_{j,p+1} = 1$ for any position p . In fact, it may be needed to add it multiple times if the previous equality holds for multiple positions. Thus, the penalty associated to weight $w_{i,j}$ is:

$$w_{i,j} \sum_{p=0}^{n-1} x_{i,p}x_{j,p+1}$$

Adding up the penalties associated to every weight yields the cost function:

$$\text{Minimize } \sum_{i=0}^{n-1} \sum_{j=0}^{n-1} w_{i,j} \sum_{p=0}^{n-1} x_{i,p}x_{j,p+1}$$

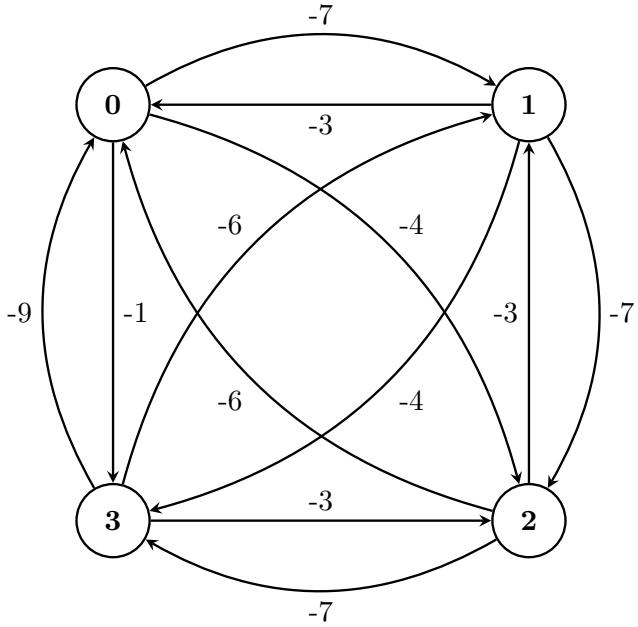
The final QUBO model is obtained by adding the penalties 3, 4 and 5 to the cost function:

$$\begin{aligned} \text{Minimize } & \sum_{i=0}^{n-1} \sum_{j=0}^{n-1} w_{i,j} \sum_{p=0}^{n-1} x_{i,p}x_{j,p+1} \\ & + a \sum_{i=0}^{n-1} \sum_{p=0}^{n-1} x_{i,p} \quad (\text{self-bias}) \\ & + b \sum_{i=0}^{n-1} \left(\sum_{p=0}^{n-1} x_{i,p} - 1 \right)^2 \quad (\text{repetition}) \\ & + c \sum_{p=0}^{n-1} \left(\sum_{i=0}^{n-1} x_{i,p} - 1 \right)^2 \quad (\text{colocation}) \end{aligned} \tag{6}$$

Since this is a quadratic unconstraint function, this is already a QUBO. The conversion to matrix is straight-forward. Let us see this in an example.

NUMERICAL EXAMPLE

Given the following graph:



Find a Hamiltonian cycle with minimum cost.

First of all, let us take a deeper look into our possible solutions. There are $(n - 1)!$ cycles in a graph with n nodes. For our 4-nodes graph there are $3! = 6$ type of cycles:

- Type A: $0 \rightarrow 1 \rightarrow 2 \rightarrow 3 \rightarrow 0$
- Type B: $0 \rightarrow 3 \rightarrow 2 \rightarrow 1 \rightarrow 0$
- Type C: $0 \rightarrow 2 \rightarrow 1 \rightarrow 3 \rightarrow 0$
- Type D: $0 \rightarrow 3 \rightarrow 1 \rightarrow 2 \rightarrow 0$
- Type E: $0 \rightarrow 1 \rightarrow 3 \rightarrow 2 \rightarrow 0$
- Type F: $0 \rightarrow 2 \rightarrow 3 \rightarrow 1 \rightarrow 0$

We may see this graphically in table 4.

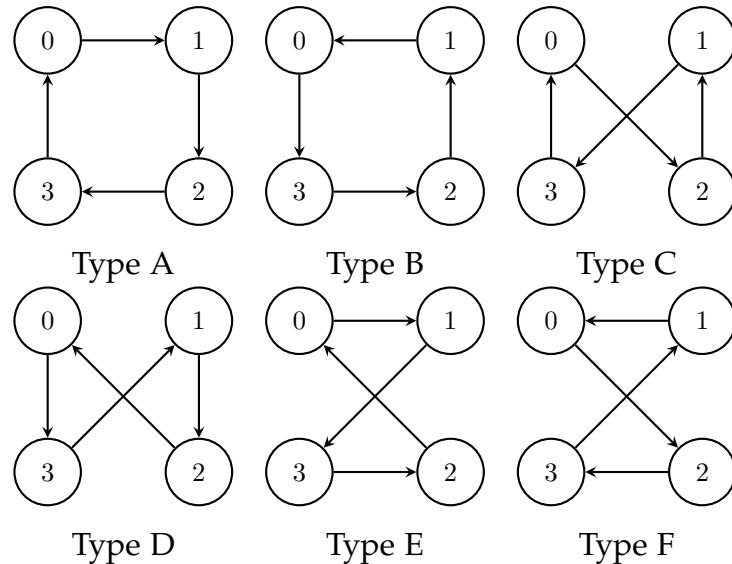


Table 4: Type of cycles in a 4-nodes graph

These six are the only possible cycles through our graph and of them will be the one with minimum cost. Since our binary encoding additionally depends on the node position, there are 4 solutions of our encoding that represent the same cycle. For example, $(0 \rightarrow 1 \rightarrow 2 \rightarrow 3 \rightarrow 0)$ and $(1 \rightarrow 2 \rightarrow 3 \rightarrow 0 \rightarrow 1)$ both represent cycle type A. In this case, type A will be the tour with minimum cost.

A graphical representation of the different penalties interactions is shown in figure 5, based on the constraints defined above.

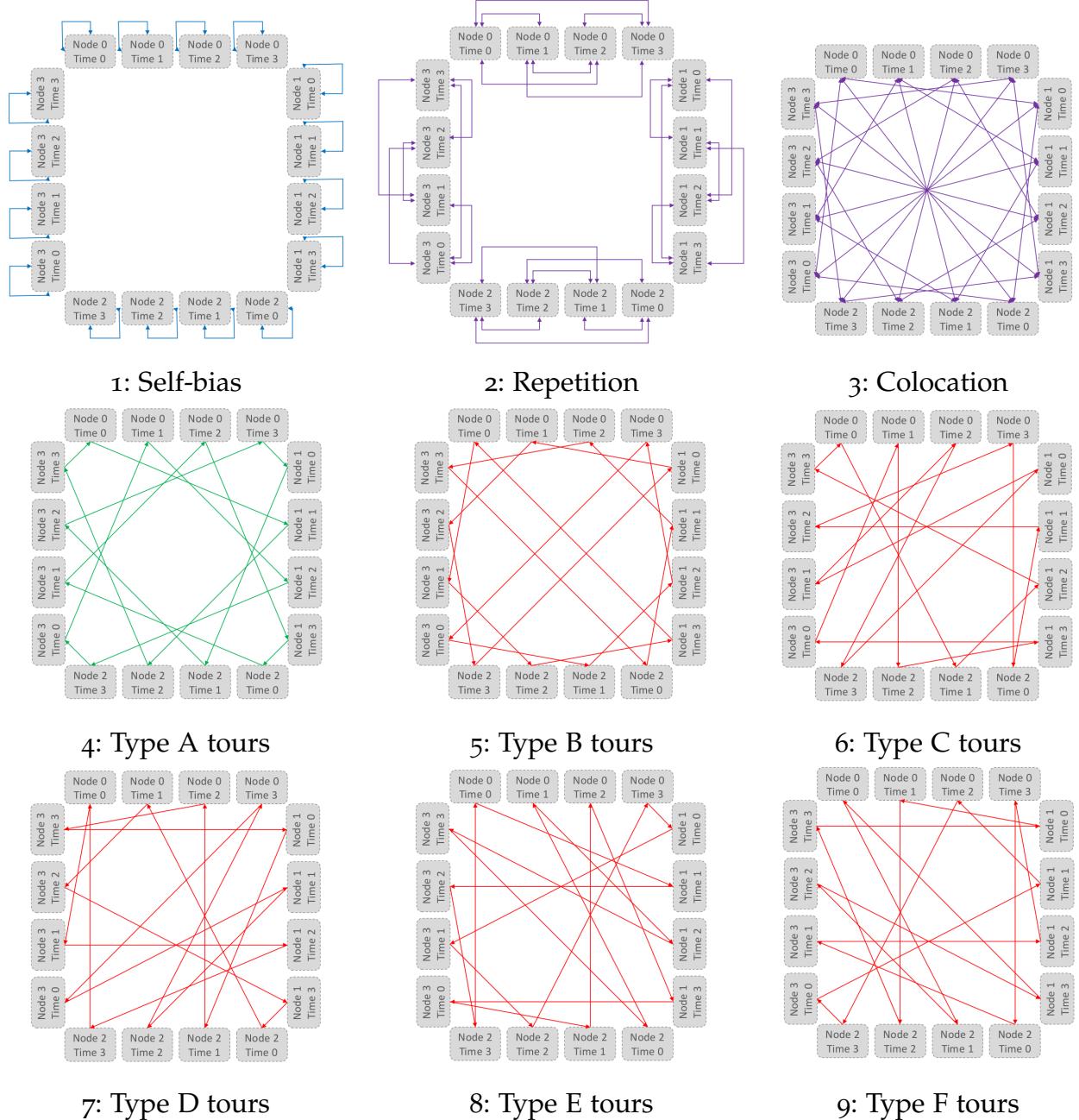


Table 5: Graphical representation of penalties between interactions [28]

We start the problem refactoring by renaming the variables:

$$(x_{0,0}, x_{0,1}, x_{0,2}, x_{0,3}, x_{1,0}, x_{1,1}, x_{1,2}, \dots, x_{3,2}, x_{3,3}) = (x_1, x_2, x_3, x_4, x_5, x_6, x_7, \dots, x_{15}, x_{16})$$

Each of the terms in the cost function 6 may be transformed into a matrix. The term associated to the weights is additionally splitted using the different types of cycles in which they appear. We thus obtain nine adjacency matrices, found on figure 4.

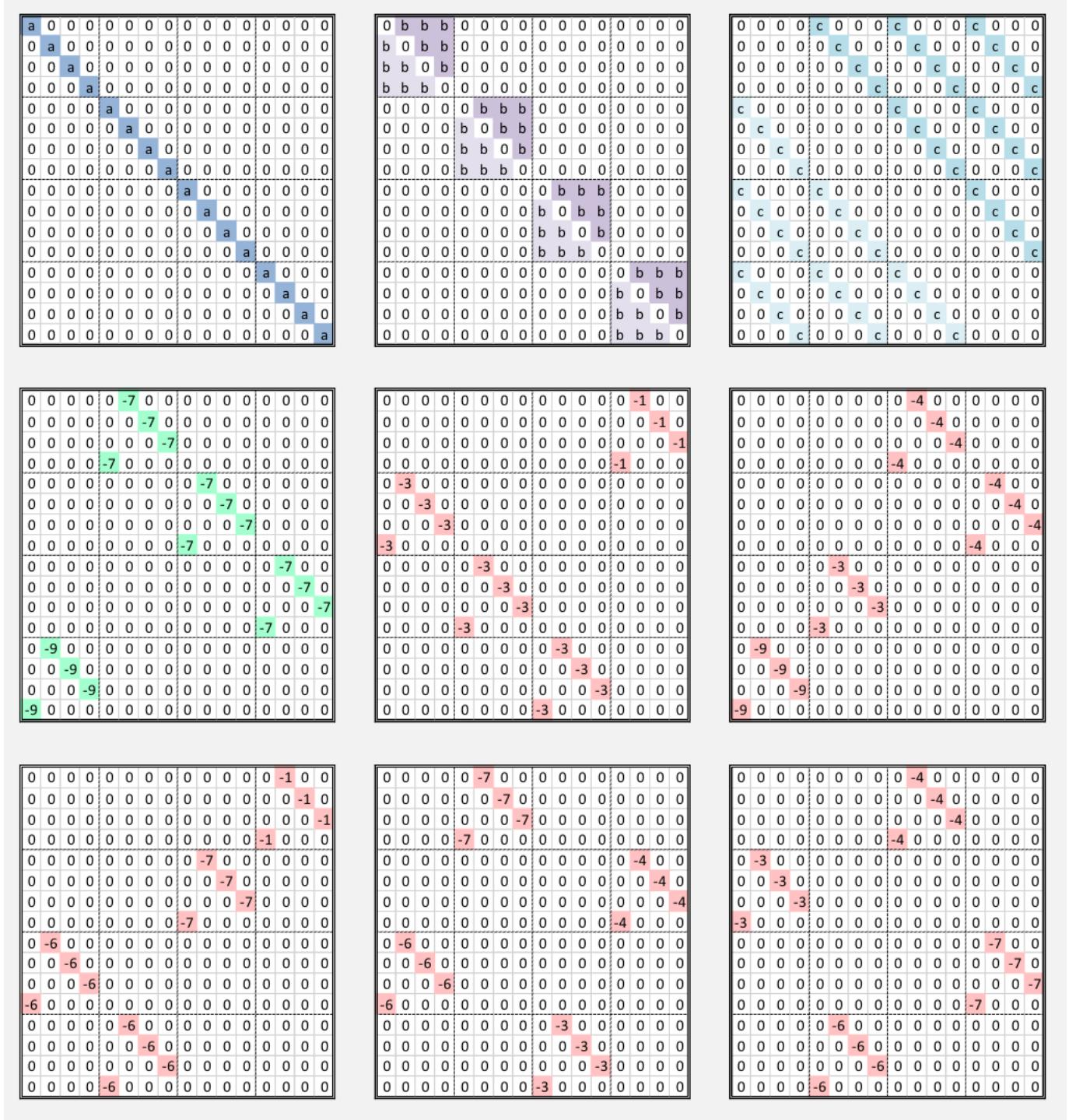


Figure 4: Term by term penalty matrices [28]

Note that the matrices associated with the constraints are all symmetric. This happens because they only depend on the graph topology and not on the weights. On the other hand, since the cost of going back and forth between a pair of nodes is not the same, the matrices associated with the cycles are never symmetric.

The final Q matrix for our problem is obtained by adding these matrices. Experimental evidence suggests that setting $a = 0$ and $b = c = 13$ and applying quantum annealing yields optimal solutions of the cost function [28]. Using these values for the penalties, the final Q matrix is seen in figure 5.

0	13	13	13	13	13	-14	0	0	13	-8	0	0	13	-2	0	0
13	0	13	13	13	0	13	-14	0	0	13	-8	0	0	13	-2	0
13	13	0	13	13	0	0	13	-14	0	0	13	-8	0	0	13	-2
13	13	13	0	-14	0	0	13	-8	0	0	13	-2	0	0	0	13
13	-6	0	0	0	0	13	13	13	13	-14	0	0	13	-8	0	0
0	13	-6	0	13	0	13	13	13	0	13	-14	0	0	13	-8	0
0	0	13	-6	13	13	0	13	0	0	13	-14	0	0	0	13	-8
-6	0	0	13	13	13	13	0	-14	0	0	13	-8	0	0	0	13
13	0	0	0	13	-6	0	0	0	13	13	13	13	13	-14	0	0
-12	13	0	0	0	13	-6	0	13	0	13	13	0	0	13	-14	0
0	-12	13	0	0	0	13	-6	13	13	0	13	0	0	0	13	-14
-12	0	-12	13	-6	0	0	13	13	13	0	-14	0	0	0	13	0
13	-18	0	0	13	0	0	0	13	-6	0	0	0	13	13	13	13
0	13	-18	0	-12	13	0	0	0	13	-6	0	13	0	13	13	13
0	0	13	-18	0	-12	13	0	0	0	13	-6	13	13	0	13	13
-18	0	0	13	-12	0	-12	13	-6	0	0	13	13	13	13	13	0

Figure 5: Final Q-matrix [28]

Finally transforming our problem into a QUBO in its matrix form, using the previous matrix:

$$\text{Minimize } f(x) = x^T Q x$$

Finding one of the following minimas:

- $x^T = [1000010000100001]$, equivalent to $x_{0,0} = x_{1,1} = x_{2,2} = x_{3,3} = 1$ and 0 otherwise. Or in path notation: $(0 \rightarrow 1 \rightarrow 2 \rightarrow 3 \rightarrow 0)$
- $x^T = [0100001000011000]$, equivalent to $x_{0,1} = x_{1,2} = x_{2,3} = x_{3,0} = 1$ and 0 otherwise. Or in path notation: $(1 \rightarrow 2 \rightarrow 3 \rightarrow 0 \rightarrow 1)$
- $x^T = [0010000110000100]$, equivalent to $x_{0,2} = x_{1,3} = x_{2,0} = x_{3,1} = 1$ and 0 otherwise. Or in path notation: $(2 \rightarrow 3 \rightarrow 0 \rightarrow 1 \rightarrow 2)$
- $x^T = [0001100001000010]$, equivalent to $x_{0,3} = x_{1,0} = x_{2,1} = x_{3,2} = 1$ and 0 otherwise. Or in path notation: $(3 \rightarrow 0 \rightarrow 1 \rightarrow 2 \rightarrow 3)$

Which corresponds to the four ways of encoding type A cycles with our notation.

2.4 D-WAVE SYSTEMS

D-Wave Systems Inc. is a Canadian company dedicated to quantum computing. In 2011, they announced the first commercial quantum computer system, D-Wave One. During the following two decades, the D-Wave team have developed a series of computing computers dedicated to quantum annealing. The last one being the Advantage System (figure 6).



Figure 6: AdvantageTM system

The quantum processing unit (QPU) of this system consist of 5640 qubits and 40,484 *couplers* (links between qubits that allow entanglement between a pair of qubits). The number of couplers is especially relevant. It tells us that not every qubit can be entangled with every other qubit. We will come back to this in section [TODOref]. Since the QPU must be isolated to operate, it is encapsulated in a system at temperatures below 15 mK. In addition, radio frequency (RF)-shielded enclosur and magnetic shieldings are used to protect it from electromagnetic interference [29].

D-Wave provides an easy-to-use software enviroment to solve problems using quantum annealing. We will explain in depth in section [TODOref].

2.4.1 Quantum Annealing in D-Wave

For this section we refer to the D-Wave documentation, which explains how quantum annealing is implemented and may be exploited in the D-Wave systems [30].

<https://journals.aps.org/prl/pdf/10.1103/PhysRevLett.117.180402>

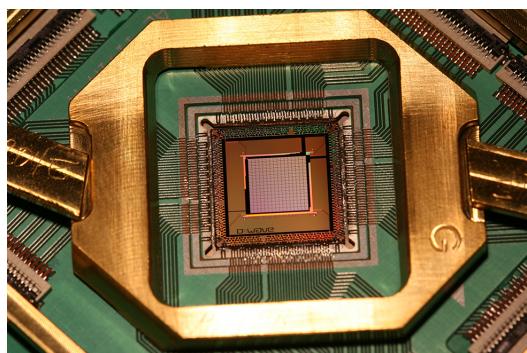


Figure 7: D-Wave QPU

BIBLIOGRAPHY

- [1] Michael A. Nielsen, Isaac Chuang, and Lov K. Grover. *Quantum Computation and Quantum Information*, volume 70. 2002. ISBN 9781107002173. doi: 10.1119/1.1463744.
- [2] Daniel Manzano. A short introduction to the Lindblad master equation. *AIP Adv.*, 10(2), feb 2020. ISSN 21583226. doi: 10.1063/1.5115323.
- [3] Pablo Bayens. *Modelos de computación cuánticos*. PhD thesis, University of Granada, 2019. URL <https://github.com/mx-psi/tfg/blob/master/tfg.pdf>.
- [4] Rafael Payá. Apuntes de Análisis Funcional. In Rafael Payá, editor, *Apunt. Análisis Func.*, chapter 5, pages 53–62. Granada, 2020. URL https://www.ugr.es/~rpaya/documentos/Funcional/2020-21/Apuntes_05.pdf.
- [5] E. Schrödinger. Die gegenwärtige Situation in der Quantenmechanik. *Naturwissenschaften*, 23(48):807–812, nov 1935. ISSN 00281042. doi: 10.1007/BF01491891.
- [6] John D. Trimmer. The Present Situation in Quantum Mechanics: A Translation of Schrödinger's "Cat Paradox" Paper. *Proc. Am. Philos. Soc.*, 124(5):323–338, 1980. URL <https://archive.is/20121204184041/http://www.tuhh.de/rzt/rzt/it/QM/cat.html#sect5>.
- [7] Dietlinde Lau. *Function Algebras on Finite Sets*. 2006. doi: 10.1007/3-540-36023-9.
- [8] Christopher M. Dawson and Michael A. Nielsen. The Solovay-Kitaev algorithm. *Quantum Inf. Comput.*, 6(1):081–095, jan 2006. ISSN 15337146. doi: 10.26421/qic6.1-6. URL <https://arxiv.org/abs/quant-ph/0505030v2>.
- [9] J. S. BELL. on the Einstein Podolsky Rosen Paradox. Technical Report 3, 1995.
- [10] A. Einstein, B. Podolsky, and N. Rosen. Can quantum-mechanical description of physical reality be considered complete? *Phys. Rev.*, 47(10):777–780, may 1935. ISSN 0031899X. doi: 10.1103/PhysRev.47.777. URL <https://journals.aps.org/pr/abstract/10.1103/PhysRev.47.777>.
- [11] S. Kirkpatrick, C. D. Gelatt, and M. P. Vecchi. Optimization by simulated annealing. *Science (80-)*, 220(4598):671–680, may 1983. ISSN 00368075. doi: 10.1126/science.220.4598.671. URL <https://science.scienmag.org/content/220/4598/671> <https://science.scienmag.org/content/220/4598/671.abstract>.

- [12] A. B. Finnila, M. A. Gomez, C. Sebenik, C. Stenson, and J. D. Doll. Quantum annealing: A new method for minimizing multidimensional functions. *Chem. Phys. Lett.*, 219(5-6):343–348, mar 1994. ISSN 00092614. doi: 10.1016/0009-2614(94)00117-0.
- [13] Tadashi Kadowaki and Hidetoshi Nishimori. Quantum annealing in the transverse Ising model. *Phys. Rev. E - Stat. Physics, Plasmas, Fluids, Relat. Interdiscip. Top.*, 58(5):5355–5363, 1998. ISSN 1063651X. doi: 10.1103/PhysRevE.58.5355.
- [14] G Nimtz and A Haibel. Zero time space: how quantum tunneling broke the light speed barrier. *Choice Rev. Online*, 46(02):46–0955–46–0955, 2008. ISSN 0009-4978. doi: 10.5860/choice.46-0955. URL https://inis.iaea.org/search/search.aspx?orig_q=RN:39067461.
- [15] P Ray, B. K. Chakrabarti, and Arunava Chakrabarti. Sherrington-Kirkpatrick model in a transverse field: Absence of replica symmetry breaking due to quantum fluctuations. *Phys. Rev. B*, 39(16):11828–11832, 1989. ISSN 01631829. doi: 10.1103/PhysRevB.39.11828.
- [16] Satoshi Morita and Hidetoshi Nishimori. Mathematical foundation of quantum annealing. *J. Math. Phys.*, 49(12):125210, 2008. ISSN 00222488. doi: 10.1063/1.2995837. URL <https://api.semanticscholar.org/CorpusID:13992889#id-name=S2CID>.
- [17] Sergei V Isakov, Guglielmo Mazzola, Vadim N Smelyanskiy, Zhang Jiang, Sergio Boixo, Hartmut Neven, and Matthias Troyer. Understanding quantum tunneling through quantum Monte Carlo simulations. *Phys. Rev. Lett.*, 117(18), 2016. ISSN 10797114. doi: 10.1103/PhysRevLett.117.180402.
- [18] Edward Farhi, Jeffrey Goldstone, Sam Gutmann, and Michael Sipser. Quantum Computation by Adiabatic Evolution. 2000. URL <http://arxiv.org/abs/quant-ph/0001106>.
- [19] M. Born and V. Fock. Beweis des Adiabatensatzes. *Zeitschrift für Phys.*, 51(3-4):165–180, mar 1928. ISSN 14346001. doi: 10.1007/BF01343193. URL <https://ui.adsabs.harvard.edu/abs/1928ZPh...51..165B/abstract>.
- [20] Tosio Kato. On the Adiabatic Theorem of Quantum Mechanics. *J. Phys. Soc. Japan*, 5(6):435–439, dec 1950. ISSN 13474073. doi: 10.1143/JPSJ.5.435. URL <https://journals.jps.jp/doi/abs/10.1143/JPSJ.5.435>.
- [21] Ernst Ising. Beitrag zur Theorie des Ferromagnetismus D. *Phys. ZS*, 6(4):920, 1924. URL <https://link.springer.com/article/10.1007%2FBF02980577>.
- [22] Fred Glover, Gary Kochenberger, and Yu Du. Quantum Bridge Analytics I: a tutorial on formulating and using QUBO models. *4OR*, 17(4):335–371, 2019. ISSN 16142411. doi: 10.1007/s10288-019-00424-y. URL <https://www.nap.edu/catalog/25196/quantum-computing-progress-and-prospects>.

- [23] Andrew Lucas. Ising formulations of many NP problems. *Front. Phys.*, 2:1–14, feb 2014. ISSN 2296424X. doi: 10.3389/fphy.2014.00005. URL www.frontiersin.org.
- [24] Gary Kochenberger, Jin Kao Hao, Fred Glover, Mark Lewis, Zhipeng Lü, Haibo Wang, and Yang Wang. The unconstrained binary quadratic programming problem: A survey. *J. Comb. Optim.*, 28(1):58–81, 2014. ISSN 15732886. doi: 10.1007/s10878-014-9734-0.
- [25] David G. Hull. Constrained Minimization: Inequality Constraints. pages 59–75, 2003. doi: 10.1007/978-1-4757-4180-3_4. URL https://link.springer.com/chapter/10.1007/978-1-4757-4180-3_4.
- [26] Ronald V. Book. Book Review: Computers and intractability: A guide to the theory of NP-completeness. *Bull. Am. Math. Soc.*, 3(2):898–905, 1980. ISSN 0273-0979. doi: 10.1090/s0273-0979-1980-14848-x. URL <https://archive.org/details/computersintract0000gare>.
- [27] Gary A Kochenberger, Fred Glover, Bahram Alidaee, and Cesar Rego. An unconstrained quadratic binary programming approach to the vertex coloring problem. *Ann. Oper. Res.*, 139(1):229–241, 2005. ISSN 02545330. doi: 10.1007/s10479-005-3449-7.
- [28] Aritra Sarkar, Zaid Al-Ars, and Koen Bertels. QuASeR Quantum Accelerated De Novo DNA Sequence Reconstruction. *arXiv*, pages 1–24, 2020. ISSN 23318422.
- [29] Welcome to D-Wave — D-Wave System Documentation documentation. URL https://docs.dwavesys.com/docs/latest/c_gs_1.html.
- [30] What is Quantum Annealing? — D-Wave System Documentation documentation, 2021. URL https://docs.dwavesys.com/docs/latest/c_gs_2.html.