Properties of Entanglement Monotones for Three-Qubit Pure States

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Various parameterizations for the orbits under local unitary transformations of three-qubit pure states are analyzed. The interconvertibility, symmetry properties, parameter ranges, calculability and behavior under measurement are looked at. It is shown that the entanglement monotones of any multipartite pure state uniquely determine the orbit of that state under local unitary transformations. It follows that there must be an entanglement monotone for three-qubit pure states which depends on the Kempe invariant defined in [2]. A form for such an entanglement monotone is proposed. A theorem is proved that significantly reduces the number of entanglement monotones that must be looked at to find the maximal probability of transforming one multipartite state to another.

I. INTRODUCTION

Entanglement is at the heart of the studies of quantum computation and quantum information theory. It is what separates these studies from their classical counterparts. If we are to understand what new phenomena occur when we look at the true quantum mechanical description of nature as opposed to the approximations of classical mechanics then we must understand how the quantum mechanical description differs from the classical description. Entanglement is a measure of this difference. While entanglement between two parties is quite well understood [2] [3] [4] [5], the entanglement within a quantum algorithm or in a state shared between many parties involves multipartite entanglement which is just beginning to be understood [6] [7] [8].

An integral part of the study of entanglement is determining the probability of transforming one pure state into another by Local Operations and Classical Communication (LOCC). For two part systems this problem is solved, or at least reduced to the problem of finding the eigenvalues of a hermitian matrix, by [2] [3] [4]. For a \( N \times M \) pure state the Schmidt decomposition tells us we can write

\[
|\psi\rangle = \sum_{i=1}^{n} \sqrt{\lambda_i^\dagger} |i\rangle |i'\rangle
\]

(1)

where the \( \lambda_i^\dagger \) are in increasing order, \( \sum_i \lambda_i^\dagger = 1 \), the \( |i\rangle \) and \( |i'\rangle \) are an orthonormal set of vectors in space \( A \) and \( B \) respectively, and \( n = \min(N, M) \). If we define

\[
E_k(|\psi\rangle) = \sum_{i=1}^{k} \lambda_i^\dagger \quad k = 1, \ldots, n - 1
\]

(2)

then the highest attainable probability of transforming \( |\psi\rangle \) to \( |\phi\rangle \), \( P(|\psi\rangle \rightarrow |\phi\rangle) \), is given by

\[
P(|\psi\rangle \rightarrow |\phi\rangle) = \min_{k} \frac{E_k(|\psi\rangle)}{E_k(|\phi\rangle)}
\]

(3)

The proof of this theorem is constructive so we can actually write down the transformation that gives us \( |\phi\rangle \) from \( |\psi\rangle \). For pure states of more than two parts no such nice theorem is known. The question of whether two three-qubit pure states can be transformed into each other with non-zero probability by LOCC has been solved by Dür et al. [5] but just getting a reasonable upper bound on that probability when it is a non-zero is unsolved. In this paper I attempt to make some progress towards solving this problem for three-qubit pure states and hopefully shed some light on how we might solve it for larger dimensional spaces and more parts.

One way to find \( P(|\psi\rangle \rightarrow |\phi\rangle) \) is to look at the entanglement monotones \( E(|\psi\rangle) \) for the two states. For the duration of the paper “state” will refer to a pure state unless explicitly called a mixed state. An entanglement monotone, EM, is defined as a function that goes from states to positive real numbers and does not increase unless explicitly called a mixed state. An entanglement duration of the paper “state” will refer to a pure state.

\[
P(\rho \rightarrow \rho') = \min_{E} \frac{E(\rho)}{E(\rho')}
\]

(4)

where the minimization is taken over the set of all EMs [10]. This can be seen by considering \( P(\rho \rightarrow \rho') \) as an EM for \( \rho \). The problem is that this minimization is difficult to take since there is no known way to characterize all the entanglement monotones for multipartite states. We would like a “minimal set” of EMs similar to the \( E_k \) for the bipartite case in order to take the minimization.

The situation for three or more parts is somewhat different than for bipartite pure states. Firstly, generic \( M \times M \) bipartite states have a stabilizer (i.e. the set of unitaries that takes a state to itself) of dimension \( M - 1 \) isomorphic to \( U(1)^{\otimes M-1} \) while pure states with more parts generically have a discrete stabilizer. States whose parts are not of the same dimension may have larger stabilizers but bipartite states are the only ones that always have a continuous stabilizer. Secondly, the generalized
Schmidt decomposition, however you choose to generalize it [11] [12], has complex coefficients for pure states with 3 or more parts. This implies that generically these states are not local unitarily equivalent to their complex conjugate states (i.e., the state with each of its coefficients complex conjugated). Also, for bipartite pure states all the local unitary (LU) invariants can be calculated from the eigenvalues of the reduced density matrices but this does not hold for more parts. I will go into more detail about LU invariants in the next section.

The structure of the paper is as follows, in section II the interconvertibility, behavior under measurement, symmetry properties, parameter ranges and calculability of two generalizations of the Schmidt decomposition of equation (4) and the polynomial invariants (defined below) are looked at. In section III it is shown that the entanglement monotones uniquely determine the orbit of multipartite pure states and this is used to show that there must be an EM algebraically independent of the known EMs. A form for this EM is proposed and studied. Section IV discusses other monotones that must exist and their properties. Lastly in section V a theorem is proved that significantly reduces the number of EMs that must be minimized over to get \( P(\rho \rightarrow \rho') \) of equation (3).

II. DECOMPOSITIONS AND INVARIANTS OF THREE-QUBIT PURE STATES

Let \( |\psi\rangle \) be a multipartite state in \( \mathcal{H}_1 \otimes \mathcal{H}_2 \ldots \otimes \mathcal{H}_n \) and let \( A_k^{(i)} : \mathcal{H}_i \rightarrow \mathcal{H}_i' \) be Krauss operators for an operation on the hilbert space \( \mathcal{H}_i \) with \( \sum_k A_k^{(i)\dagger} A_k^{(i)} = I_i \) and \( I_i \) is the identity acting on \( \mathcal{H}_i \). A (non-increasing) EM is a real valued function \( E(|\psi\rangle) \) such that

\[
E(|\psi\rangle) \geq \sum_k p_k E\left(\frac{I_1 \otimes \ldots \otimes A_k^{(i)} \otimes \ldots \otimes I_n |\psi\rangle}{\sqrt{p_k}}\right)
\]  

(5)

for any state \(|\psi\rangle\), operation \(A_k^{(i)}\), and space \(i\) where

\[
p_k = ||I_1 \otimes \ldots \otimes A_k^{(i)} \otimes \ldots \otimes I_n |\psi\rangle||^2.
\]  

(6)

This definition for pure states is taken from the definition for a general state in [11]. One can always transform a state into product states and a product state cannot be transformed into anything but another product state so the value of an EM for a product state is chosen to be zero and all other states must have a non-negative value for the EM. Since \( A_k^{(i)} \) can be a unitary operator or the inverse of that operator, equation (6) implies that all EMs must be invariant under LU. Hence, a first step to understanding the EMs is to look at the LU invariants that parameterize the set of orbits.

There are many ways to find LU invariants for three-qubit states [13] [12] [11] [13] [14] some of which can be generalized to more parts and larger spaces but for now I will concentrate on the three-qubit case. The three sets of invariants I will look at in this section are the polynomial invariants [13], what I will call the diagonalization decomposition [11] and what I will call the maximization decomposition [12].

A. The Polynomial Invariants

A general polynomial invariant \( P_{\sigma,\tau}(|\psi\rangle) \) for a state of the form

\[
|\psi\rangle = \sum_{i,j,k=0}^1 t_{ijk}|ijk\rangle
\]  

(7)

is written as

\[
P_{\sigma,\tau}(|\psi\rangle) = \sum t_{i_1 j_1 k_1} \ldots t_{i_n j_n k_n} \tilde{t}_{i_1 j_1 k_1} \ldots \tilde{t}_{i_n j_n k_n} \]

(8)

where \( \sigma \) and \( \tau \) are permutations on \( n \) elements, repeated indices are summed and \( \tilde{t} \) stands for the complex conjugate of \( t \). If one applies a unitary to any of the qubits in \(|\psi\rangle\) and explicitly writes out \( P_{\sigma,\tau}(|\psi\rangle) \) again it becomes apparent that \( P_{\sigma,\tau}(|\psi\rangle) \) is invariant. Of course, any polynomial in terms of the polynomial invariants \( P_{\sigma,\tau}(|\psi\rangle) \) is another polynomial invariant. In fact, it can be shown that all the polynomial invariants are of this form.

We know from [11] that generic three-qubit states have a discrete stabilizer so the number of independent polynomial invariants is given by

\[
\dim[C^2 \otimes C^2 \otimes C^2] - 3 \dim[SU(2)] - \dim[U(1)] - 1 = 5
\]  

(9)

where the last \(-1\) is due to the fact that we are using normalized states. The 5 independent continuous invariants are

\[
I_1 = P_{(12)}
I_2 = P_{(12),e}
I_3 = P_{(12),(12)}
I_4 = P_{(12),(132)}
I_5 = \left| \sum t_{i_1 j_1 k_1} t_{i_2 j_2 k_2} t_{i_3 j_3 k_3} t_{i_4 j_4 k_4} \times \epsilon_{i_1 i_2} \epsilon_{i_3 i_4} \epsilon_{j_1 j_2} \epsilon_{j_3 j_4} k_1 k_3 k_4 k_2 \right|^2
\]

(10)

where \( \epsilon_{00} = \epsilon_{11} = 0, \) and \( \epsilon_{01} = -\epsilon_{10} = 1 \) and again repeated indices are summed. \( I_4 \) is the Kempe invariant referred to in the abstract. If one writes out \( I_5 \) and uses the identity \( \epsilon_{ij} \epsilon_{rs} = \delta_{ir} \delta_{js} - \delta_{is} \delta_{jr} \), it can be shown that \( I_5 \) is just the sum and difference of 64 polynomials of the form in equation (8). With one more discrete invariant,

\[
I_6 = \text{sign} \left( \text{Im} \left( P_{(34)(56),(13524)} \right) \right)
\]

(11)
the LU orbit of a three-qubit state is determined uniquely [12, 17]. In this paper I will define \(\text{sign}[x]\) as 1 for non-negative numbers and \(-1\) otherwise. The polynomial invariants have the advantage of being easy to compute for any state and the four previously known independent EMs [3] are the following simple functions of \(I_1, I_2, I_3\) and \(I_5\)

\[
\begin{align*}
\tau_{(AB)C} &= 2(1 - I_1) \\
\tau_{(AC)B} &= 2(1 - I_2) \\
\tau_{(BC)A} &= 2(1 - I_3) \\
\tau_{ABC} &= 2 \sqrt{I_5}.
\end{align*}
\]

\[\tag{12}\]

B. The Diagonalization Decomposition

The diagonalization decomposition, DD, introduced by Acin et al. [12] is accomplished by first defining matrices \((T_0)_{j,k} = t_{0,j,k}\) and \((T_1)_{j,k} = t_{1,j,k}\), then finding a unitary operation on space A that makes \(T_0\) singular, finding unitaries on space B and C that make \(T_0\) diagonal and using the remaining phase freedom to get rid of as many phases as possible. What is left is a state of the form

\[
|\psi_{DD}\rangle = \sqrt{\mu_0} |000\rangle + \sqrt{\mu_1} e^{i\phi} |100\rangle + \sqrt{\mu_2} |101\rangle + \sqrt{\mu_3} |110\rangle + \sqrt{\mu_4} |111\rangle
\]

\[\tag{13}\]

where \(\mu_i \geq 0, \mu_0 + \mu_1 + \mu_2 + \mu_3 + \mu_4 = 1\) and \(0 \leq \phi \leq \pi\). Note that generally there are two unitaries that will make \(T_0\) singular but it can be shown that only one will lead to \(\phi\) between 0 and \(\pi\). If there is another solution, with \(\phi\) between \(\pi\) and \(2\pi\) exclusive, it is referred to as the dual state of \(|\psi_{DD}\rangle\). Some nice properties of DD are that there is a 1 to 1 correspondence with the orbits and there are a set of invertible functions between the parameters of the decomposition and the set of polynomial invariants given above. Namely,

\[
\begin{align*}
I_1 &= 1 - 2\mu_0(\mu_2 + \mu_4) - 2\Delta \\
I_2 &= 1 - 2\mu_0(\mu_3 + \mu_4) - 2\Delta \\
I_3 &= 1 - 2\mu_0(\mu_2 + \mu_3 + \mu_4) \\
I_4 &= 1 - 3!(\mu_2 + \mu_3)(\mu_0 - \mu_4) + \mu_4(1 - \mu_4) \\
&\quad - \mu_2\mu_3\mu_0 + (1 - \mu_0)(\Delta - \mu_1\mu_4) \\
I_5 &= 4\mu_0^2\mu_1^2 \\
I_6 &= \text{sign} [\text{sign}[\sin(\phi)]\mu_0^2\sqrt{\mu_1\mu_2\mu_3\mu_4}] \\
&\quad \times (\Delta - \mu_4(1 - 2\mu_0 + \mu_1) - \mu_2\mu_3))
\end{align*}
\]

\[\tag{14}\]

where \(\Delta = \mu_1\mu_4 + \mu_2\mu_3 - 2\sqrt{\mu_1\mu_2\mu_3\mu_4}\cos(\phi)\) and if we define

\[
\begin{align*}
J_1 &= \frac{1}{4} \left( 1 - I_1 - I_2 + I_3 - 2\sqrt{I_5} \right) \\
J_2 &= \frac{1}{4} \left( 1 - I_1 + I_2 - I_3 - 2\sqrt{I_5} \right) \\
J_3 &= \frac{1}{4} \left( 1 + I_1 - I_2 - I_3 - 2\sqrt{I_5} \right)
\end{align*}
\]

\[\tag{15}\]

then the coefficients are given by

\[
\begin{align*}
\mu_0^\pm &= \frac{J_4 + J_5 \pm \sqrt{\gamma}}{2(J_1 + J_4)} \\
\mu_i^\pm &= \frac{J_i}{\mu_0^\mp}, \quad i = 2, 3, 4 \\
\mu_1^\pm &= 1 - \mu_0^\pm \pm \frac{J_2 + J_3 + J_4}{\mu_0^\pm} \\
\cos(\phi^\pm) &= \frac{\mu_1^\pm \mu_4^\pm + \mu_2^\pm \mu_3^\pm - J_3}{2\sqrt{\mu_1^\pm \mu_2^\pm \mu_3^\pm \mu_4^\pm}} \\
\text{sign} [\sin(\phi^\pm)] &= I_6 \text{sign} [\sqrt{\mu_1^\pm \mu_2^\pm \mu_3^\pm \mu_4^\pm}] [J_1 - J_2J_3 - J_4(J_2 + J_3 + J_4 - (\mu_0^\pm)^2)]
\end{align*}
\]

\[\tag{16}\]

where \(\gamma = (J_4 + J_5)^2 - 4(J_1 + J_4)(J_2 + J_4)(J_3 + J_4) \geq 0\). The + and − solutions for the coefficients correspond to \(|\psi_{DD}\rangle\) and its dual state. The inversion of the equations for \(I_6\) was done independently in [17]. Note that their definition of \(I_4\) is different from the one in this paper.

Another nice property of the DD is that we can perform an arbitrary measurement on it in space A and stay in the DD form. Since any measurement can be broken into a series of two outcome measurements [18], we can look at the two outcome measurement \(A_1\) and \(A_2\) where \(A_1^\dagger A_1 + A_2^\dagger A_2 = I\). Using the singular value decomposition we can write \(A = U_1D_1V\) where \(V\) does not depend on \(i\) because the two positive hermitian operators \(A_1^\dagger A_1\) and \(A_2^\dagger A_2\) sum to the identity and therefore must be simultaneously diagonalizable. The diagonal matrices, \(D_1\), can be written as

\[
D_1 = \begin{bmatrix} x & 0 \\ 0 & y \end{bmatrix}, \quad D_2 = \begin{bmatrix} \sqrt{1-x^2} & 0 \\ 0 & \sqrt{1-y^2} \end{bmatrix}
\]

\[\tag{17}\]

where \(0 \leq x, y \leq 1\) [18]. Since we are only concerned with what orbit the outcomes are in we may choose the \(U_1\) transformation. Also, matrices of the form

\[
\begin{bmatrix} e^{i\psi_1} & 0 \\ 0 & e^{i\psi_2} \end{bmatrix}
\]

\[\tag{18}\]

where \(\psi_1\) and \(\psi_2\) are real numbers, commute with the \(D_1\) matrices so the most general \(V\) can be written as

\[
\begin{bmatrix} \alpha & \sqrt{1-\alpha^2} e^{-i\theta} \\ -\sqrt{1-\alpha^2} e^{i\theta} & \alpha \end{bmatrix}
\]

\[\tag{19}\]

where \(0 \leq \alpha \leq 1\) and \(\theta\) is real. If we choose

\[
U_1 = \frac{1}{\sqrt{\gamma}} \begin{bmatrix} \frac{y\alpha}{\sqrt{1-\alpha^2}} & x\sqrt{1-\alpha^2} e^{-i\theta} \\ x\sqrt{1-\alpha^2} e^{i\theta} & \frac{-y\alpha}{\sqrt{1-\alpha^2}} \end{bmatrix}
\]

\[
\gamma = y^2\alpha^2 + x^2(1-\alpha^2)
\]

\[\tag{20}\]
and similarly for $U_2$ with $(x, y)$ replaced with $(\sqrt{1-x^2}, \sqrt{1-y^2})$, then in going from $|\psi_{DD}\rangle$ to $A_l|\psi_{DD}\rangle$ the DD coefficients undergo the following transformations

$$
\begin{align*}
\mu_0 &\to \frac{x^2y^2\mu_0}{\gamma} \\
\mu_1 &\to \frac{1}{\gamma} \left| e^{-i\theta}(x^2 - y^2)\alpha\sqrt{\mu_0(1 - \alpha^2)} + e^{i\phi}\gamma\sqrt{\mu_1} \right|^2 \\
\mu_4 &\to \mu_1\gamma \\
\phi &\to \arg \left[ e^{-i\theta}(x^2 - y^2)\alpha\sqrt{\mu_0(1 - \alpha^2)} + e^{i\phi}\gamma\sqrt{\mu_1} \right]
\end{align*}
$$

and again similarly for $A_2|\psi_{DD}\rangle$. Things become more complicated when $\phi$ becomes larger than $\pi$ and we have a dual solution. In this case we need to transform to the dual state which can be quite tedious. It should also be noted that if we want to plug the new form for the DD coefficients into equations (14) then the normalization must be taken into account. The normalization will just be the sum of the new forms for $\mu_0$ through $\mu_4$.

C. The Maximization Decomposition

The Maximization Decomposition [1], MD, has a somewhat different way of decomposing the three qubit states. First we find the states, $|\phi_A\rangle$, $|\phi_B\rangle$, and $|\phi_C\rangle$ each defined up to an overall phase, that maximize

$$
g(|\phi_A\rangle, |\phi_B\rangle, |\phi_C\rangle) = \| \langle \psi | \phi_A \rangle | \phi_B \rangle | \phi_C \rangle \|^2
$$

and apply a unitary such that $|\phi_A\rangle | \phi_B \rangle | \phi_C \rangle$ becomes $|000\rangle$. Defining $|1\rangle$, up to an overall phase, as the vector perpendicular to $|0\rangle$, then the derivative of $g$ along $|1\rangle$ at the point $|000\rangle$, must be zero because $g(|0\rangle, |0\rangle, |0\rangle)$ is a maximum. Since we still have phase freedom in $|0\rangle$ and $|1\rangle$ this implies that $\langle \psi | 000 \rangle = 0$ and similarly for $\langle \psi | 010 \rangle$ and $\langle \psi | 001 \rangle$. Using the remaining phase freedom in the choice of $|0\rangle$ and $|1\rangle$ we can eliminate all but one phase leaving us with

$$
|\psi_{MD}\rangle = ae^{i\phi}|000\rangle + b|011\rangle + c|101\rangle + d|110\rangle + f|111\rangle
$$

where $a^2 + b^2 + c^2 + d^2 + f^2 = 1$, $0 \leq \phi \leq 2\pi$, $0 \leq a$, $b$, $c$, $d$, $f$ and $b$, $c$, $d$, $f \leq a$. Note that $g(|0\rangle, |0\rangle, |0\rangle) = a^2$. Unfortunately, the parameters as they are given above are not in 1 to 1 correspondence with the orbits. While the decomposition is generically unique, there are choices of the parameters within the given ranges that are not the result of the decomposition. For example, states with $a^2 = b^2 = c^2 = d^2 = f^2 = \frac{1}{5} - \frac{4}{5}$ and any choice of $\phi$ have $g \left( \frac{1}{\sqrt{2}}(|0\rangle + |1\rangle), \frac{1}{\sqrt{2}}(|0\rangle + |1\rangle), \frac{1}{\sqrt{2}}(|0\rangle + |1\rangle) \right) \geq a^2$ for $\epsilon \leq 0.014$. Hence, these choices of the parameters are not a result of the decomposition. The true ranges of the parameters that would give a 1 to 1 correspondence with the orbits are as yet unknown.

A nice property of the MD is that it is symmetric in particle exchange. The problem with using $1 - a^2$ as an EM is that one needs to find the global maximum of a 6 dimensional space with many local maxima. One more interesting fact about the MD is that $1 - a^2$ is a non-increasing EM. We know this because in [10] it is shown that a function of the form

$$
E_{k_A,k_B,k_C}(|\psi\rangle) = \max_{\Gamma_A,\Gamma_B,\Gamma_C} \| \Gamma_A \otimes \Gamma_B \otimes \Gamma_C |\psi\rangle \|^2
$$

where $\Gamma_X$ is a $k_X$-dimensional projector on system $X = A, B, C$, is an non-decreasing EM and $E_{1,1,1}(|\psi\rangle) = a^2$. The EM $1 - a^2$ can be shown to be independent of the $\tau$ from equation (12) by looking at the gradient vectors of the $\tau$, $1 - a^2$ and $N = a^2 + b^2 + c^2 + d^2 + f^2$ at, for instance, the point $a = 3$, $b, c, d, f = 1$ and $\phi = \frac{\pi}{2}$. Since the gradient vectors span a 6 dimensional space, $1 - a^2$ cannot be written in terms of the $\tau$ and $N$. The problem with using $1 - a^2$ as an EM is that one needs to find the global maximum of a 6 dimensional space with many local maxima to calculate it. This is a difficult task for most states.
III. FIFTH INDEPENDENT EM

In section II it was shown that all EMs must be invariant under LU and hence are determined by the orbit of the state. For three qubit states this means that EMs are a function of only the polynomial invariants, DD coefficients or MD coefficients. In fact, this determination is unique.

Theorem 1 The set of all EMs for any multipartite pure state, $|\psi\rangle$, uniquely determine the orbit of the state.

Proof. Suppose two states $|\psi\rangle$ and $|\phi\rangle$ in $H_1 \otimes H_2 \ldots \otimes H_n$ have the same values for the EMs but lie in different orbits. We know by using equation (4) that

$$P(|\psi\rangle \rightarrow |\phi\rangle) = P(|\phi\rangle \rightarrow |\psi\rangle) = 1 \tag{28}$$

so $|\psi\rangle$ can be transformed to $|\phi\rangle$ (and vice versa) by n-party LOCC, n-LOCC, with probability 1. Since EMs are non-increasing with any n-LOCC they must remain constant during the entire transformation from $|\psi\rangle$ to $|\phi\rangle$ (and vice versa). Also, we know that any EM between a system $X = A, B, \ldots$ and the rest of the systems thought of as one (e.g. between $B$ and $ACD, \ldots$), I will call these EMs 2-EMs, is also an EM for multipartite states. This is because any n-LOCC on the multipartite state is also a 2-LOCC between $X$ and the rest of the systems, since the 2-EM is non-increasing over 2-LOCC it must also be non-increasing over n-LOCC. In particular the sum of the lowest $k$ eigenvectors of the reduced density matrices,

$$E_k^X(|\psi\rangle) = \sum_{i=1}^k \lambda_i^X(\rho_X(|\psi\rangle)), \tag{29}$$

(i.e. the 2-EMs in equation (2)) must be EMs. So the $E_k^X(|\psi\rangle)$ must remain unchanged and hence the spectrum of $\rho_X$ is unchanged during the transformation from $|\psi\rangle$ to $|\phi\rangle$. In particular a measurement on space $X$, given by $A_1$ and $A_2$, must be such that

$$\rho_X \left( \frac{A_1|\psi\rangle}{\sqrt{N}} \right) = U \rho_X (|\psi\rangle) U^\dagger \tag{30}$$

where $N$ is the normalization. The only way this can be satisfied is if $\frac{A_1}{\sqrt{N}}$ is a unitary matrix. This means that $|\psi\rangle$ and $|\phi\rangle$ are unitarily equivalent which contradicts our original supposition. □

Since we know there are 5 parameters that determine the orbit of a three qubit state then by theorem 1 there must be 5 independent, continuous EMs. To the best of the author’s knowledge the only 4 known independent continuous EMs that don’t require a difficult maximization over a multidimensional space are the four $\tau$ EMs defined in equation (12). Any candidate for the fifth independent EM must depend on $I_4$ since the $\tau$ are invertible functions of $I_1$, $I_2$, $I_3$ and $I_5$ respectively. The following function fulfills that criterion

$$\sigma_{ABC} = 3 - (I_1 + I_2 + I_3)I_4 \tag{31}$$

and numerical results suggest that it is an EM. After generating over 300,000 random states and applying a random operation to each of them the inequality in equation (6) was never violated by $\sigma_{ABC}$. Also, note that $\sigma_{ABC}$ is symmetric in particle permutations as is $\tau_{ABC}$. For the rest of the paper I will assume that $\sigma_{ABC}$ is an EM. Indeed, it may be that there is a set of measure zero or perhaps just a very small measure for which $\sigma_{ABC}$ is not a monotone and my numerical test didn’t explore this space but there must exist some function of the polynomial invariants which is independent of the $\tau$s and is an EM. For it to be useful in improving our upper bound for $P(|\psi\rangle \rightarrow |\phi\rangle)$ there should be pairs of states $|\psi\rangle$ and $|\phi\rangle$ such that

$$\frac{\sigma_{ABC}(|\psi\rangle)}{\sigma_{ABC}(|\phi\rangle)} < \min \tau \left( \frac{|\psi\rangle}{|\phi\rangle} \right) \tag{32}$$

and I have found such states numerically. The largest value of

$$\frac{\sigma_{ABC}(|\psi\rangle)}{\sigma_{ABC}(|\phi\rangle)} - \min \tau \left( \frac{|\psi\rangle}{|\phi\rangle} \right) \tag{33}$$

that I found in my limited number of examples of was 0.01 and I was able to find examples of states for which $\tau (|\psi\rangle)/\tau (|\phi\rangle)$ is greater than one for all $\tau$ and $\sigma_{ABC}(|\psi\rangle)/\sigma_{ABC}(|\phi\rangle)$ is less than one.

IV. OTHER EMS AND THE DISCRETE INVARIANT

The five independent continuous EMs, $\tau_{(AB)C}$, $\tau_{(AC)B}$, $\tau_{(BC)A}$, $\tau_{ABC}$ and $\sigma_{ABC}$, can easily be inverted to find $I_1$ - $I_5$ but to completely determine the orbit of a state we must also have an EM that will give us the value of the discrete invariant $I_6$. This is equivalent to finding an EM that is not the same for a state and it complex conjugate state. Note that $I_1, I_2, I_3$ and hence the $\tau$ and $\sigma_{ABC}$ do not change when a state is conjugated but by looking at any of the sets of LU invariants we can see that generically a state is not LU equivalent to its conjugate. By looking at equation (4) we can see that this implies that there must be EMs that are not the same for the generic state and its conjugate. It is also easy to see that for any operation that takes a state $|\psi\rangle$ to its conjugate $|\bar{\psi}\rangle$ with probability $p$ there is an operation that takes $|\psi\rangle$ to $|\psi\rangle$ with the same probability. So, for a generic state $|\psi\rangle$ there must be an EM that goes down for the operation $|\psi\rangle \rightarrow |\bar{\psi}\rangle$ and a similar one that goes down the same amount for $|\psi\rangle \rightarrow |\psi\rangle$. So, EMs of the following form must exist

$$v^\pm (|\psi\rangle) = \begin{cases} v + v' & \pm I_6 = 1 \\ v & o.w. \end{cases} \tag{34}$$
where \( u \) and \( u' \) are functions of \( \tau_{(AB)C}, \tau_{(AC)B}, \tau_{(BC)A}, \tau_{ABC} \) and \( \sigma_{ABC} \).

Also, from [4] we know that there are two classes of three-part entangled states (i.e., states with \( \tau_{(AB)C}, \tau_{(AC)B}, \tau_{(BC)A} > 0 \)) that can be converted into each other with some non-zero probability within the class and zero probability between the classes. Namely, the GHZ-class which contains
\[
|\text{GHZ}\rangle = \frac{1}{\sqrt{2}} (|000\rangle + |111\rangle)
\]
and has non-zero \( \tau_{ABC} \) and the W-class which contains
\[
|\text{W}\rangle = \frac{1}{\sqrt{3}} (|001\rangle + |010\rangle + |100\rangle).
\]
and has \( \tau_{ABC} = 0 \). Looking again at equation (34) we see that \( \tau_{ABC} \) tells us that \( P(|\psi_{\text{GHZ}}\rangle \rightarrow |\psi_{\text{W}}\rangle) = 0 \) but none of the previously defined EMs tell us that \( P(|\psi_{\text{GHZ}}\rangle \rightarrow |\psi_{\text{W}}\rangle) = 0 \). Since the only way to get \( P(|\psi_{\text{GHZ}}\rangle \rightarrow |\psi_{\text{W}}\rangle) = 0 \) is to have an EM that is finite for GHZ-class states and infinite for W-class states or zero for GHZ-class states and non-zero for W-class states such an EM must exist.

V. FINDING A MINIMAL SET

Since \( \tau_{(AB)C}, \tau_{(AC)B}, \tau_{(BC)A}, \tau_{ABC}, \sigma_{ABC} \) and \( u^\pm \) determine the orbit of the state all other EMs must depend on them. A fairly general way to create new EMs from known EMs is to use what I will call \( f \)-type functions.

**Definition 1** A function \( f : \mathcal{S} \subset \mathbb{R}^n \rightarrow \mathbb{R} \) is an \( f \)-type function if it satisfies the following

1. \( f(0) = 0 \)
2. if \( x_i \geq y_i \) for all \( i = 1, 2, \ldots, n \) then \( f(x) \geq f(y) \) for \( x, y \in \mathcal{S} \)
3. \( f(px + (1 - p)y) \geq pf(x) + (1 - p)f(y) \) for any \( x, y \in \mathcal{S} \) and \( 0 \leq p \leq 1 \).

For a set of EM, \( \{E_i\} \), we have
\[
E_i(|\psi\rangle) = pE_1 \left( \frac{A_1|\psi\rangle}{\sqrt{p}} \right) + (1 - p)E_1 \left( \frac{A_2|\psi\rangle}{\sqrt{1 - p}} \right).
\]
for any measurement \( A_1, A_2 \) and any state \( |\psi\rangle \). So, we have
\[
f[E_i(|\psi\rangle)] \geq pf \left[ pE_1 \left( \frac{A_1|\psi\rangle}{\sqrt{p}} \right) + (1 - p)E_1 \left( \frac{A_2|\psi\rangle}{\sqrt{1 - p}} \right) \right] \\
\geq pf \left[ E_1 \left( \frac{A_1|\psi\rangle}{\sqrt{p}} \right) \right] + (1 - p)\left[ E_1 \left( \frac{A_2|\psi\rangle}{\sqrt{1 - p}} \right) \right] 
\]
where the first inequality comes from property 2 and the second comes from property 3. Hence, \( f(E_1, \ldots, E_n) \) is also an EM. We can show that any EM \( f(E_1, \ldots, E_n) \) that is an \( f \)-type function of monotones \( E_1, \ldots, E_n \) does not modify the upper bound on \( P(|\psi\rangle \rightarrow |\phi\rangle) \) given by
\[
P(|\psi\rangle \rightarrow |\phi\rangle) \leq \min_i \frac{E_i(|\psi\rangle)}{E_i(|\phi\rangle)}.
\]
First for the one dimensional case.

**Lemma 1** If \( f(x) \) is an \( f \)-type function with \( n = 1 \) then
\[
\frac{f(x)}{f(y)} \geq \min \left\{ \frac{x}{y}, 1 \right\}
\]
for any \( x, y \in \mathcal{S} \).

**Proof.**

Case 1 For \( x \geq y \) from property 2 we know \( f(x) \geq f(y) \) and hence
\[
\frac{f(x)}{f(y)} \geq 1.
\]

Case 2 For \( x < y \) if we choose \( p = \frac{x}{y} \in [0, 1) \) then we know from properties 1 and 3 that \( f(py) \geq pf(y) \) and so
\[
\frac{f(x)}{f(y)} \geq \frac{x}{y}, \quad \square
\]
For \( n \) dimensions we have the following theorem (proved with S. Daftuar and D. Whitehouse).

**Theorem 2** If \( f(x) \) is an \( f \)-type function then
\[
\frac{f(x)}{f(y)} \geq \min \left\{ \frac{x_i}{y_i}, 1 \right\} \quad i = 1, 2, \ldots, n
\]
for \( x, y \in \mathcal{S} \).

**Proof.** Let
\[
c = \min \left\{ \frac{x_i}{y_i}, 1 \right\}
\]
then we have

Case 1 If \( c \geq 1 \) then from property 2 \( f(x) \geq f(y) \) and so
\[
\frac{f(x)}{f(y)} \geq 1
\]

Case 2 If \( c < 1 \) then define \( z_i = \frac{x_i}{c}, \quad i = 1, 2, \ldots, n \) and \( g(r) = f(rz) \). Notice that \( g(r) \) is an \( f \)-type function with \( n = 1 \) and hence
\[ g(c) \geq c \]  
\[ g(1) \geq c \]  

or substituting in \( f \) we have

\[ \frac{f(x)}{f(y)} \geq c. \]

Using \( z_i \geq y_i \) and property 2 we have

\[ \frac{f(x)}{f(y)} \geq c. \square \]

For three-qubit states if we take the minimum of \( E(|\psi\rangle)/E(|\phi\rangle) \) over \( \mathcal{E} = \{ \tau_{(AB)C}, \tau_{(AC)B}, \tau_{(BC)A}, \tau_{ABC}, \sigma_{ABC}, \nu^\pm \} \) we are actually taking the minimum over the infinite set of all \( f \)-type functions of \( \mathcal{E} \). Although from theorem 1 we know that all EMs must be a function of \( \mathcal{E} \) it is possible that there exist EMs that are not \( f \)-type functions of \( \mathcal{E} \). These EMs could cause \( P(|\psi\rangle \rightarrow |\phi\rangle) \) to be lower than the minimum of \( E(|\psi\rangle)/E(|\phi\rangle) \) over \( \mathcal{E} \). The EM mentioned at the end of section [14] is an example of such an EM.

\[ f(\vec{x}) \geq c \]

VI. CONCLUSIONS AND FURTHER RESEARCH

Theorem 1 along with theorem 2 implies that there should be a (not necessarily finite) minimal set of EMs, \( M \), for which all EMs for three-qubit states or similarly for any type of multipartite states are \( f \)-type functions of \( M \). I conjecture that such a minimal set should be simple since the \( f \)-type functions seem to be a rather general way of creating EMs that are functions of other EMs. The difficult part seems to be finding the EMs that are minimal and showing that they are minimal. Using numerical results it seems that the \( \tau \) may be minimal. I looked at functions of the \( \tau \) that are almost but not quite \( f \)-type such as \( \tau^{1.01} \) and numerically tested whether they are EMs or not. None of them were EMs. I cannot say the same for \( \sigma_{ABC} \) and definitely not for \( \nu^\pm \) since I do not have an explicit form for the \( \nu \).

There is further research that may help these problems. If one could invert the equations in [26] to write \( a, b, c, d, f \) and \( \phi \) in terms of \( I_1, \ldots, I_6 \) that would allow us to calculate the EM \( 1 - a^2 \) not to mention find the ranges for and calculate the values of \( a, b, c, d, f \) and \( \phi \). The EM \( 1 - a^2 \) could be used to replace \( \sigma_{ABC} \) or perhaps as an addition to \( \mathcal{E} \) and may prove more useful than \( \sigma_{ABC} \). As far as finding the minimal EMs and showing that they are minimal, the arbitrary measurement on the DD at the end of section [11] may be useful since it allows us to look at the value of \( I_1, \ldots, I_6 \) before and after an arbitrary measurement on an arbitrary state with far less parameters than if we didn’t take out the LU freedom. Also, it may be able to tell us the maximal probability of transforming the general complex state \( |\tilde{\psi}\rangle \) and this is a crucial piece of information that is needed to calculate \( \nu \) in equation [34]. Unfortunately, most of these tasks involve trying to solve nontrivial equations or systems of equations with many variables which can be difficult or even impossible.

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