ABSTRACT<br>\section*{Title of dissertation: ENUMERATION OF HARMONIC FRAMES AND FRAME BASED DIMENSION REDUCTION}<br>Matthew J. Hirn, Doctor of Philosophy, 2009<br>\section*{Dissertation directed by: Professor John J. Benedetto<br><br>Professor Kasso A. Okoudjou<br><br>Department of Mathematics}

We investigate two aspects of frame theory, one of a theoretical nature, the other very much on the applied side. In the former, we enumerate all harmonic frames of prime order, and develop partial proofs concerning the structure of the symmetry group for this subset of frames. In the latter, we develop frame theory in the context of kernel eigenmap methods, merging the two theories in a practical manner and applying new algorithms to hyperspectral imagery data for the purposes of material classification. These two problems, while seemingly separate, are united by frame theory and serve to illustrate both the beautiful theoretical nature of frames as well as their practicality in dealing with real world problems.

# ENUMERATION OF HARMONIC FRAMES AND FRAME BASED DIMENSION REDUCTION 

by<br>Matthew J. Hirn

# Dissertation submitted to the Faculty of the Graduate School of the University of Maryland, College Park in partial fulfillment of the requirements for the degree of Doctor of Philosophy <br> 2009 

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## Dedication

For my parents, Darleen and John Hirn.

## Acknowledgments

This document is the culmination of five years of work, almost all of which was collaborative in some way. I hope here to express in some small way, my enduring gratitude to those that made this extraordinary journey possible.

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as someone to help me work out the tough spots and bounce ideas off of. I have always enjoyed just stopping by his office to see if he is there, whether it is to ask a serious math question or simply say hi and see how things are going. And though I have now finished my thesis, I look forward to our continued friendship, and of course working on our next idea.

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## Chapter 1

## Introduction to Frames

### 1.1 Frames

Given a finite dimensional Hilbert space $\mathbb{H}$, a basis is a set of elements that give a unique representation for each element $\mathbb{H}$. Frames, on the other hand, are an overcomplete set of elements that allow for an infinite number of representations of each element in $\mathbb{H}$. While bases are useful in certain situations because the representation is unique, at other times it is better to have the flexibility provided by a frame.

The purpose of this dissertation is two-fold. In chapter two we examine a subclass of finite frames known as harmonic frames. In particular, we will study harmonic frames with a prime number of elements. The main result is to prove a recursive formula for the number of harmonic frames of prime order. A secondary result is to partially determine the symmetry group of all such harmonic frames.

The second main focus is to examine the usefullness of frames, in conjunction with kernel based dimension reducing methods, for the classification of materials in multispectral and hyperspectral imagery data. Results here are theoretically motivated, yet are empirical in nature. We plan to give results that exhibit the promise of this approach. The theoretical motivations can be found in chapter three, while chapter four contains empirical results.

First though, we begin with an introduction to frame theory.

### 1.2 Frame Theory

Let $\mathcal{I}$ be a possibly infinite, but countable, index set. A frame $[24,25,14,22]$ for a separable Hilbert space $\mathbb{H}$ is a collection of vectors

$$
\begin{equation*}
\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\} \subset \mathbb{H} \tag{1.1}
\end{equation*}
$$

for which there exists constants $0<A \leq B<\infty$ such that for each $f \in \mathbb{H}$,

$$
\begin{equation*}
A\|f\|^{2} \leq \sum_{i \in \mathcal{I}}\left|\left\langle\varphi_{i}, f\right\rangle\right|^{2} \leq B\|f\|^{2} \tag{1.2}
\end{equation*}
$$

Constants $A$ and $B$ which satisfy (1.2) are called frame bounds of $\Phi$. Optimally chosen values of $A$ and $B$ are referred to as the optimal frame bounds of the frame. When $A=B$, the frame $\Phi$ is referred to as a tight frame.

As an example of a frame one may choose an orthonormal basis - it is in fact a tight frame with constants $A=B=1$. A union of any two orthonormal bases is a tight frame with constants $A=B=2$, etc. A union of an orthonormal basis with $N$ arbitrary unit norm vectors is a frame with bounds $A=1$ and $B=N+1$. If the Hilbert space is infinite dimensional and $N$ is finite this last example is certainly not a tight frame. Some other examples are given by figures 1.1, 1.2, and 1.3.

Given a frame $\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\}$, a dual frame is a collection of vectors $\widehat{\Phi}=\left\{\hat{\varphi}_{i}: i \in \mathcal{I}\right\} \subset \mathbb{H}$ such that for all $f \in \mathbb{H}$, we have the reconstruction formula

$$
\begin{equation*}
f=\sum_{i \in \mathcal{I}}\left\langle f, \varphi_{i}\right\rangle \hat{\varphi}_{i} . \tag{1.3}
\end{equation*}
$$



Figure 1.1: Frame with six elements in $\mathbb{R}^{3}$


Figure 1.2: Buckyball tight frame


Figure 1.3: The platonic solids form tight frames

It is perhaps not immediately clear that every frame should have a dual frame. In order to obtain a dual frame to a frame $\Phi$, we will define the frame operator. Let $\mathbb{F}=\mathbb{R}$ or $\mathbb{C}$, and define $\ell^{2}(\mathcal{I})$ as the space of all sequences indexed by $\mathcal{I}$ with finite energy, i.e.

$$
\begin{equation*}
\ell^{2}(\mathcal{I}):=\left\{c=\left(c_{i}\right)_{i \in \mathcal{I}}: c_{i} \in \mathbb{F} \forall i \in \mathcal{I} \text { and } \sum_{i \in \mathcal{I}}\left|c_{i}\right|^{2}<\infty\right\} \tag{1.4}
\end{equation*}
$$

Given a frame $\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\}$, the analysis operator $L: \mathbb{H} \rightarrow \ell^{2}(\mathcal{I})$ is defined by

$$
\begin{equation*}
L(f):=\left(\left\langle f, \varphi_{i}\right\rangle\right)_{i \in \mathcal{I}} . \tag{1.5}
\end{equation*}
$$

The adjoint of the analysis operator $L^{\star}$ is called the synthesis operator, and $S=L^{\star} L$ is the frame operator. For each $c \in \ell^{2}(\mathcal{I})$, the synthesis operator is defined by

$$
\begin{equation*}
L^{\star}(c)=\sum_{i \in \mathcal{I}} c_{i} \varphi_{i} \tag{1.6}
\end{equation*}
$$

Given the previous two equations, it is easy to see that the frame operator is defined by

$$
\begin{equation*}
S(f)=\sum_{i \in \mathcal{I}}\left\langle f, \varphi_{i}\right\rangle \varphi_{i} \tag{1.7}
\end{equation*}
$$

where $f \in \mathbb{H}$. The following known theorems characterize the analysis, synthesis, and frame operators.

Theorem 1.2.1 ([16]). Let $\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\} \subset \mathbb{H}$ be a frame for $\mathbb{H}$. Then the following are satisfied:
a. $L$ is a bounded operator from $\mathbb{H}$ into $\ell^{2}(\mathcal{I})$.
b. $L^{\star}$ extends to a bounded operator from $\ell^{2}(\mathcal{I})$ into $\mathbb{H}$.
c. $L$ and $L^{\star}$ are adjoint operators of each other.

Theorem 1.2.2 ([16]). Let $\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\} \subset \mathbb{H}$ be a frame for $\mathbb{H}$. The frame operator $S=L^{\star} L$ maps $\mathbb{H}$ onto $\mathbb{H}$ and is a positive invertible operator satisfying $A \cdot I d \leq S \leq B \cdot I d$ and $B^{-1} \cdot I d \leq S^{-1} \leq A^{-1} \cdot I d$. In particular, $\Phi$ is a tight frame if and only if $S=A \cdot I d$.

Note, in theorem 1.2.2 Id denotes the identity map on $\mathbb{H}$, i.e., $I d(f)=f$ for all $f \in \mathbb{H}$. The sequence of vectors $\left\{S^{-1}\left(\varphi_{i}\right): i \in \mathcal{I}\right\}$ is called the canonical dual frame, and is a dual frame for $\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\}$. That is we have

$$
\begin{equation*}
f=\sum_{i \in \mathcal{I}}\left\langle f, S^{-1}\left(\varphi_{i}\right)\right\rangle \varphi_{i} \tag{1.8}
\end{equation*}
$$

and

$$
\begin{equation*}
f=\sum_{i \in \mathcal{I}}\left\langle f, \varphi_{i}\right\rangle S^{-1}\left(\varphi_{i}\right), \tag{1.9}
\end{equation*}
$$

where both sums converge unconditionally in $\mathbb{H}$.
We note here that dual frames are not in general unique and this underlines the importance of the canonical dual frame.

For a particular given frame, it may not be easy to apply the procedure in the preceding paragraph to obtain a dual frame. One special case in which it is easy is that of Parseval frames. A Parseval frame is a tight frame consisting of unit norm vectors. If $\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\}$ is a Parseval frame, then for every $f \in \mathbb{H}$,

$$
\begin{equation*}
f=\sum_{i \in \mathcal{I}}\left\langle f, \varphi_{i}\right\rangle \varphi_{i} . \tag{1.10}
\end{equation*}
$$

In particular, Parseval frames are dual frames of themselves. For this reason, among others, Parseval frames are the 'best behaved' of frames, and we will present here
some of their additional properties.
Most of the basic, general properties of Parseval frames can be derived from the following.

Theorem 1.2.3 ([16]). A collection of vectors $\Phi=\left\{\varphi_{i}: i \in \mathcal{I}\right\} \subset \mathbb{H}$ is a Parseval frame for $\mathbb{H}$ if and only if there exists a Hilbert space $\mathbb{K}$ containing $\mathbb{H}$ as a closed subspace and an orthonormal basis $\left\{e_{i}: i \in \mathcal{I}\right\}$ of $\mathbb{K}$ such that for all $i \in \mathcal{I}, P e_{i}=\varphi_{i}$, where $P$ is the orthogonal projection onto $\mathbb{H}$.

Equation (1.10) follows immediately from Theorem 1.2.3. Indeed, we have for $f \in \mathbb{H}$,

$$
\begin{aligned}
P^{2} f & =P(P f)=\sum_{i \in \mathcal{I}}\left\langle P f, e_{i}\right\rangle P e_{i} \\
& =\sum_{i \in \mathcal{I}}\left\langle f, P e_{i}\right\rangle \varphi_{i} \\
& =\sum_{i \in \mathcal{I}}\left\langle f, \varphi_{i}\right\rangle \varphi_{i} .
\end{aligned}
$$

### 1.3 Finite Frame Theory

In finite dimensional Hilbert vector spaces, the notion of a frame becomes intutively simple. Let $s, d \in \mathbb{N}$, and suppose $s \geq d ; \Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\}$ is a frame for $\mathbb{F}^{d}$ (recall $\mathbb{F}=\mathbb{R}$ or $\mathbb{C}$ ) if and only if it is a spanning system for $\mathbb{F}^{d}$. In the finite setting it is often convenient to use matrix notation when working with frames. As such, we will consider $\varphi_{j}$ as a vector in $\mathbb{F}^{d}$, and $\Phi$ as a $d \times s$ matrix,
where the $j^{\text {th }}$ column is $\varphi_{j}$. More explicitly:

$$
\begin{array}{r}
\varphi_{j}=\left(\varphi_{j}(i)\right)_{i=1}^{d} \\
\Phi \in \mathcal{M}_{d \times s}(\mathbb{F}) \text { and } \Phi_{i, j}=\varphi_{j}(i) . \tag{1.12}
\end{array}
$$

Recasting section 1.2 in terms of finite frames and matrices, we see that the analysis operator, $L$, now maps $\mathbb{F}^{d}$ into $\mathbb{F}^{s}$. In fact, for each $f \in \mathbb{F}^{d}$, the analysis operator is given by:

$$
\begin{equation*}
L(f)=\Phi^{\star} f=\left(\left\langle f, \varphi_{i}\right\rangle\right)_{i=1}^{s} \tag{1.13}
\end{equation*}
$$

Similarly, the synthesis operator maps $\mathbb{F}^{s}$ onto $\mathbb{F}^{d}$, and for each $c \in \mathbb{F}^{s}$ is given by:

$$
\begin{equation*}
L^{\star}(c)=\Phi c=\sum_{i=1}^{s} c_{i} \varphi_{i} \tag{1.14}
\end{equation*}
$$

Combining equations (1.13) and (1.14), we see that the frame operator maps $\mathbb{F}^{d}$ to $\mathbb{F}^{d}$ and, for each $f \in \mathbb{F}^{d}$, is given by:

$$
\begin{equation*}
S(f)=L^{\star} L(f)=\Phi \Phi^{\star} f=\sum_{i=1}^{s}\left\langle f, \varphi_{i}\right\rangle \varphi_{i} \tag{1.15}
\end{equation*}
$$

A frame that is finite, tight, and unit norm is known as a finite unit norm tight frame, or a FUNTF. If $\Phi$ is a FUNTF with frame constant $A$, then it is known that $A=s / d$ and $S=\frac{s}{d} I$, where $I$ is the $d \times d$ identity matrix.

One way to characterize FUNTFs is the frame potential [8]. Let $\mathbb{S}^{d-1} \subset \mathbb{F}^{d}$ denote the unit sphere in $\mathbb{F}^{d}$. For any unit norm frame $\Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\}$, the frame potential is defined as

$$
\begin{align*}
F P: \underbrace{\mathbb{S}^{d-1} \times \ldots \times \mathbb{S}^{d-1}}_{s \text { times }} & \rightarrow[0, \infty)  \tag{1.16}\\
F P(\Phi) & :=\sum_{i, j=1}^{s}\left|\left\langle\varphi_{i}, \varphi_{j}\right\rangle\right|^{2} . \tag{1.17}
\end{align*}
$$

The following theorem characterizes FUNTFs in terms of the frame potential.

Theorem 1.3.1 ([8]). For a given $s$ and d, the following hold:
a. Every local minimizer of the frame potential is also a global minimizer.
b. If $s \leq d$, the minimum value of the frame potential is $s$, and the minimizers are precisely the orthonormal sequences in $\mathbb{C}^{d}$.
c. If $s \geq d$, the minimum value of the frame potential is $s^{2} / d$, and the minimizers are precisely the FUNTFs for $\mathbb{C}^{d}$.

### 1.4 Finite Subspace Frames ${ }^{1}$

In the finite frame setting, it is not difficult to define the notion of a finite subspace frame. Let $\Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\} \subset \mathbb{F}^{d}$ and let $W$ be a subspace of $\mathbb{F}^{d}$ of dimension $r<d$. We say $\Phi$ is a finite subspace frame for $W$ if $\operatorname{span}(\Phi)=W$. It is clear from this definition that there exist constants $0<A \leq B<\infty$ such that for each $f \in W$,

$$
\begin{equation*}
A\|f\|^{2} \leq \sum_{i=1}^{s}\left|\left\langle f, \varphi_{i}\right\rangle\right|^{2} \leq B\|f\|^{2} \tag{1.18}
\end{equation*}
$$

We note that if we had instead used (1.18) as our definition, then it would not necessarily imply that $\operatorname{span}(\Phi)=W$ but rather that $\operatorname{span}(\Phi) \supseteq W$. The unit norm property as well as the notion of a tight frame remain similar in this setting. More specifically, if we can take $A=B$ in (1.18) then we call $\Phi$ a tight subspace frame.

[^0]Finally, if $\Phi$ is a finite unit norm tight subspace frame, then we say $\Phi$ is a subspace FUNTF.

We define $L, L^{\star}$, and $S$ exactly the same as we did previously, however we note that the properties of these maps change for subspace frames. In particular, we see:
(a) $L: \mathbb{F}^{d} \rightarrow \mathbb{F}^{s}$ is no longer injective, but rather $\operatorname{ker}(L)=\left(\mathbb{F}^{d} \backslash W\right) \cup\{0\}$.
(b) $L^{\star}: \mathbb{F}^{s} \rightarrow \mathbb{F}^{d}$ is no longer surjective, but rather image $\left(L^{\star}\right)=W$.
(c) Based on (a) and (b), we see that $S: \mathbb{F}^{d} \rightarrow \mathbb{F}^{d}$ is no longer invertible.

Because of (c), theorem 1.2.2 nor equations (1.8) and (1.9) hold for subspace frames. The question then becomes: in what sense do subspace frames behave like standard frames? Theorems below show that subspace frames satisfy natural modifications of theorem 1.2.2, equation (1.8), and equation (1.9).

Let $W_{\text {on }}$ be a set of $r$ orthonormal vectors such that $\operatorname{span}\left(W_{o n}\right)=W$. We will also consider $W_{o n}$ as an $d \times r$ matrix where the columns of this matrix are the vectors in the set $W_{o n}$. We define $\Phi_{W}$ to be the $r \times s$ matrix whose columns are the coordinates of $\Phi$ in $W_{o n}$; that is:

$$
\begin{equation*}
\Phi_{W}:=W_{o n}^{\star} \Phi \tag{1.19}
\end{equation*}
$$

The $i^{\text {th }}$ column of $\Phi_{W}$ is the projected $W$-subspace coordinates of $\varphi_{i}$.

Proposition 1.4.1. The columns of $\Phi_{W}$ are a frame for $\mathbb{F}^{r}$.

Proof. Since $\operatorname{span}\left(W_{o n}\right)=W$, we have $\operatorname{ker}\left(W_{o n}^{\star}\right) \cap W=\{0\}$. Therefore, since $\operatorname{span}(\Phi)=W$ as well, we see that $W_{o n}^{\star} \Phi$ has rank $r$.

We denote the analysis, synthesis, and frame operators of $\Phi_{W}$ by $L_{W}, L_{W}^{\star}$, and $S_{W}$, respectively. In terms of the analysis operator, $L$, for $\Phi$, we have for each $g \in \mathbb{F}^{r}$,

$$
\begin{equation*}
L_{W}(g)=L\left(W_{o n} g\right)=\Phi^{\star} W_{o n} g \tag{1.20}
\end{equation*}
$$

Similarly, for each $c \in \mathbb{F}^{s}$, the synthesis operator of $\Phi_{W}$ is defined as

$$
\begin{equation*}
L_{W}^{\star}(c)=W_{o n}^{\star} L^{\star}(c)=W_{o n}^{\star} \Phi c . \tag{1.21}
\end{equation*}
$$

Combining equations (1.20) and (1.21) we see that for each $g \in \mathbb{F}^{r}, S_{W}$ is defined as

$$
\begin{equation*}
S_{W}(g)=L_{W}^{\star} L_{W}(g)=W_{o n}^{\star} L^{\star}\left(L\left(W_{o n} g\right)\right)=W_{o n}^{\star} \Phi \Phi^{\star} W_{o n} g=W_{o n}^{\star} S W_{o n} g . \tag{1.22}
\end{equation*}
$$

By proposition 1.4 .1 we see that $S_{W}$ will satisfy theorem 1.2 .2 as well as equations (1.8) and (1.9).

Theorem 1.4.2. $\Phi$ is a subspace FUNTF for $W$ with frame bound $A$ if and only if $\Phi_{W}$ is a FUNTF for $\mathbb{F}^{r}$ with frame bound $A$.

Proof. We do the forward direction first: let $g \in \mathbb{F}^{r}$, then:

$$
\begin{aligned}
\left\langle S_{W} g, g\right\rangle & =\left\langle L_{W} g, L_{W} g\right\rangle \\
& =\left\langle\Phi^{\star} W_{o n} g, \Phi^{\star} W_{o n} g\right\rangle \\
& =\sum_{j=1}^{s}\left|\left\langle W_{o n} g, \varphi_{j}\right\rangle\right|^{2} \\
& =A\left\|W_{o n} g\right\|^{2} \\
& =A\left\langle W_{o n} g, W_{o n} g\right\rangle
\end{aligned}
$$

Therefore we have:

$$
\begin{aligned}
\left\langle S_{W} g, g\right\rangle-A\left\langle W_{o n} g, W_{o n} g\right\rangle & =0 & \Longrightarrow \\
\left\langle S_{W} g, g\right\rangle-A\left\langle W_{o n}^{\star} W_{o n} g, g\right\rangle & =0 & \Longrightarrow \\
\left\langle g,\left(S_{W}-A I\right) g\right\rangle & =0 & \Longrightarrow \\
S_{W} & =A I . &
\end{aligned}
$$

For the reverse direction, let $f \in W$. There exists $g \in \mathbb{F}^{r}$ such that $W_{o n} g=f$. Therefore,

$$
\begin{aligned}
A\|f\|^{2} & =A\langle f, f\rangle \\
& =A\left\langle W_{o n} g, W_{o n} g\right\rangle \\
& =\langle A g, g\rangle \\
& =\left\langle S_{W} g, g\right\rangle \\
& =\left\langle W_{o n}^{\star} \Phi \Phi^{\star} W_{o n} g, g\right\rangle \\
& =\left\langle\Phi^{\star} W_{o n} g, \Phi^{\star} W_{o n} g\right\rangle \\
& =\left\langle\Phi^{\star} f, \Phi^{\star} f\right\rangle \\
& =\sum_{j=1}^{s}\left|\left\langle f, \varphi_{j}\right\rangle\right|^{2}
\end{aligned}
$$

Corollary 1.4.3. If $\Phi$ is a subspace FUNTF for $W$ with frame bound $A$, then $A=s / r$.

We define the canonical dual frame of $\Phi_{W}$ in the usual way, that is $\widehat{\Phi}_{W}=$
$S_{W}^{-1} \Phi_{W}$. We now define the canonical dual subspace frame of $\Phi$ as follows:

$$
\begin{equation*}
\widehat{\Phi}=W_{o n} \widehat{\Phi}_{W}=W_{o n} S_{W}^{-1} W_{o n}^{\star} \Phi . \tag{1.23}
\end{equation*}
$$

As the name implies, the set $\widehat{\Phi}=\left\{\hat{\varphi}_{i}: i=1, \ldots s\right\}=\left\{W_{o n} S_{W}^{-1} W_{o n}^{\star} \varphi_{i}: i=1, \ldots, s\right\}$ will have the following properties:

Proposition 1.4.4. $\widehat{\Phi}$ is a subspace frame for $W$.

Proof. This follows from proposition 1.4.1.

Theorem 1.4.5. Every $f \in \mathrm{~W}$ can be represented as

$$
\begin{equation*}
f=\sum_{i=1}^{s}\left\langle f, \hat{\varphi}_{i}\right\rangle \varphi_{i}=\sum_{i=1}^{s}\left\langle f, \varphi_{i}\right\rangle \hat{\varphi}_{i} . \tag{1.24}
\end{equation*}
$$

Proof. The first representation formula is $\Phi \widehat{\Phi}^{\star} f=f$ for all $f \in W$. Letting $f=$ $W_{o n} g$ for some $g \in \mathbb{F}^{r}$, we have:

$$
\begin{align*}
\Phi \widehat{\Phi}^{\star} f & =\Phi\left(W_{o n} S_{W}^{-1} W_{o n}^{\star} \Phi\right)^{\star} f \\
& =\Phi \Phi^{\star} W_{o n}\left(S_{W}^{-1}\right)^{\star} W_{o n}^{\star}\left(W_{o n} g\right) \\
& =S W_{o n} S_{W}^{-1} g \\
& =S W_{o n}\left(W_{o n}^{\star} S W_{o n}\right)^{-1} g \tag{1.25}
\end{align*}
$$

Since $W_{o n} W_{o n}^{\star}$ is the identity on $W$,

$$
\begin{aligned}
(1.25) & =W_{o n} W_{o n}^{\star} S W_{o n}\left(W_{o n}^{\star} S W_{o n}\right)^{-1} g \\
& =W_{o n} g \\
& =f
\end{aligned}
$$

The second representation formula is $\widehat{\Phi} \Phi^{\star} f=f$ for all $f \in W$.

$$
\begin{aligned}
\widehat{\Phi} \Phi^{\star} f & =\left(W_{o n} S_{W}^{-1} W_{o n}^{\star} \Phi\right) \Phi^{\star} f \\
& =W_{o n}\left(W_{o n}^{\star} S W_{o n}\right)^{-1} W_{o n}^{\star} S W_{o n} g \\
& =W_{o n} g \\
& =f
\end{aligned}
$$

The following commutative diagram illustrates the above ideas:

Figure 1.4: Subspace frames diagram


We can also extend the frame potential to subspace frames via the following theorem.

Theorem 1.4.6. For a given $s$ and $d$, let $W$ be a subspace of $\mathbb{F}^{d}$ of dimension $r<d$ and consider the resctricted frame potential:

$$
\begin{equation*}
\left.\mathrm{FP}\right|_{W}: W \cap \underbrace{\left.\mathbb{S}^{N-1} \times \cdots \times \mathbb{S}^{N-1}\right)}_{s_{\text {times }}} \rightarrow[0, \infty) \tag{1.26}
\end{equation*}
$$

Then:

1. Every local minimizer of the restricted frame potential is also a global minimizer.
2. If $s \leq r$, the minimum value of the restricted frame potential is $s$, and the minimizers are precisely the orthonormal sequences in $W$.
3. If $s \geq r$, the minimum value of the restricted frame potential is $s^{2} / r$, and the minimizer are precisely the subspace FUNTFs for $W$.

Proof. Let $W_{\text {on }}$ be a set of $r$ orthonormal vectors such that $\operatorname{span}\left(W_{\text {on }}\right)=W$ and consider it as an $d \times r$ matrix. If $\Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\}$ is a finite unit norm set of vectors in $W$, then the coordinates of $\Phi$ in $W_{o n}$ are given by the $r \times s$ matrix $\Phi_{W}=W_{o n}^{\star} \Phi$. In [8] it is shown that $\operatorname{FP}(\Phi)=\operatorname{Tr}\left(S^{2}\right)$, where $S$ is the frame operator of $\Phi$. Using the previous two statements we then have:

$$
\begin{aligned}
\left.\mathrm{FP}\right|_{W}(\Phi) & =\operatorname{Tr}\left(S^{2}\right) \\
& =\operatorname{Tr}\left(\left[\left(W_{o n} \Phi_{W}\right)\left(W_{o n} \Phi_{W}\right)^{\star}\right]^{2}\right) \\
& =\operatorname{Tr}\left(\left[\Phi_{W} \Phi_{W}^{\star}\right]^{2}\right) \\
& =\operatorname{Tr}\left(S_{W}^{2}\right) \\
& =\operatorname{FP}\left(\Phi_{W}\right)
\end{aligned}
$$

Since $\Phi_{W}$ is a finite unit norm set of vectors in $\mathbb{C}^{r}$, we can apply theorem 1.3.1 to get (1) and (2). Combining theorem 1.3.1 along with theorem 1.4.2 gives (3).

## Chapter 2

## Enumeration of Prime Order Harmonic Frames

### 2.1 Introduction

### 2.1.1 Harmonic Frames

Harmonic frames are class of FUNTFs that have their origin in the Discrete Fourier Transform (DFT) matrix. The un-normalized $s \times s$ DFT matrix is defined as

$$
\begin{equation*}
D_{s}:=\left(e^{2 \pi i m n / s}\right)_{m, n=0}^{s-1} . \tag{2.1}
\end{equation*}
$$

Noting that $e^{2 \pi i m n / s}=e^{2 \pi i(m+j s)(n+k s) / s}$ for any $j, k \in \mathbb{Z}$, we introduce the additive group of integers mod $s$,

$$
\begin{equation*}
\mathbb{Z}_{s}=\mathbb{Z} / s \mathbb{Z}:=\{0, \ldots, s-1 \bmod s\} \tag{2.2}
\end{equation*}
$$

Choosing $d \leq s$ distinct columns of $D_{s}$, say $n_{1}, \ldots, n_{d} \in \mathbb{Z}_{s}$, we can form the following $s$ vectors in $\mathbb{C}^{d}$ :

$$
\begin{equation*}
\varphi_{m}=\frac{1}{\sqrt{d}}\left(e^{2 \pi i m n_{j} / s}\right)_{j=1}^{d}, \quad m \in \mathbb{Z}_{s} \tag{2.3}
\end{equation*}
$$

The set $\Phi=\left\{\varphi_{m}: m \in \mathbb{Z}_{s}\right\}$ is in fact a FUNTF for $\mathbb{C}^{d}$, and any frame of this type is called a DFT-FUNTF. As we shall see, the DFT-FUNTFs are a subset of the harmonic frames.

Remark 2.1.1. Since we will be dealing exclusively with finite frames in this chapter, we shall interchangeably consider the frame $\Phi$ as a set or a matrix (whose columns are the vectors $\varphi_{i}$ ), with the appropriate usage being determined by the context. See section 1.3 (Finite frame theory) for details on considering a frame as a matrix.

Let $G$ denote a group. Define $\mathbb{C}^{\times}$as the group of units of $\mathbb{C}$, that is the set $\mathbb{C} \backslash\{0\}$ endowed with multiplication. A character of a group $G$ is a group homomorphism $\xi: G \rightarrow \mathbb{C}^{\times}$that satisfies

$$
\begin{equation*}
\xi\left(g g^{\prime}\right)=\xi(g) \xi\left(g^{\prime}\right), \quad \forall g, g^{\prime} \in G \tag{2.4}
\end{equation*}
$$

Suppose $G$ is a finite group of order $s$, i.e.

$$
\begin{equation*}
G=\left\{g_{i}: i=1, \ldots, s\right\} . \tag{2.5}
\end{equation*}
$$

Then for each $g_{i} \in G, \xi\left(g_{i}\right)$ is a $s$-th root of unity. If $G$ is also abelian, then it has exactly $s$ characters, $\left\{\xi_{i}: i=1, \ldots, s\right\}$. The set of vectors $\left\{\left(\xi_{i}\left(g_{j}\right)\right)_{j=1}^{s}: i=1, \ldots, s\right\}$ form an orthogonal basis for $\mathbb{C}^{s}$. The matrix with these vectors as rows,

$$
\begin{equation*}
\left(\xi_{i}\left(g_{j}\right)\right)_{i, j=1}^{s} \tag{2.6}
\end{equation*}
$$

is the character table of $G$. In particular, when $G \cong \mathbb{Z}_{s}$, the character table of $G$ is $D_{s}$.

Let $\mathcal{U}\left(\mathbb{C}^{d}\right)$ denote the group of unitary transformations on $\mathbb{C}^{d}$, i.e.

$$
\begin{equation*}
\mathcal{U}\left(\mathbb{C}^{d}\right):=\left\{U \in \mathcal{M}_{d \times d}(\mathbb{C}): U^{\star} U=U U^{\star}=I\right\} \tag{2.7}
\end{equation*}
$$

Furthermore, let $\mathcal{I} \subseteq\{1, \ldots, s\}$ with $|\mathcal{I}|=d$, and suppose $G$ is a finite abelian group of order $s$. Then, for any $U \in \mathcal{U}\left(\mathbb{C}^{d}\right)$, the set,

$$
\begin{equation*}
\Phi=\left\{U\left(\xi_{i}\left(g_{j}\right)\right)_{i \in \mathcal{I}}: j=1, \ldots, s\right\} \tag{2.8}
\end{equation*}
$$

is a frame for $\mathbb{C}^{d}$ and is called a harmonic frame. Note that when $G \cong \mathbb{Z}_{s}$ and $U=I$, one obtains a DFT-FUNTF.

Important in the study of harmonic frames is the notion of the symmetry group. The symmetry group of a FUNTF $\Phi$ for $\mathbb{C}^{d}$ is the group:

$$
\begin{equation*}
\operatorname{Sym}(\Phi):=\left\{U \in \mathcal{U}\left(\mathbb{C}^{d}\right):\left\{U \varphi_{i}: i=1, \ldots, s\right\}=\left\{\varphi_{i}: i=1, \ldots, s\right\}\right\} . \tag{2.9}
\end{equation*}
$$

We can recast the definition of symmetry group in terms matrices. Let $S_{k}$ denote the group of permutations on $k$ elements. We say $P \in \mathcal{M}_{d \times d}(\mathbb{C})$ is a permutation matrix if there exists a permutation $\sigma \in S_{d}$ such that

$$
P_{i, j}= \begin{cases}1, & \text { if } j=\sigma(i)  \tag{2.10}\\ 0, & \text { otherwise }\end{cases}
$$

Let $\mathcal{P}_{s}$ denote the set of all $s \times s$ permuation matrices. Then, in terms of matrices, the symmetry group of $\Phi$ is

$$
\begin{equation*}
\operatorname{Sym}(\Phi)=\left\{U \in \mathcal{U}\left(\mathbb{C}^{d}\right): \exists P \in \mathcal{P}_{s} \text { such that } U \Phi=\Phi P\right\} \tag{2.11}
\end{equation*}
$$

While there has been much work on harmonic frames and subjects related to them (see, for example, $[20,27,33,34,36]$ ), we will need only the following result from [33].

Theorem 2.1.2. A FUNTF $\Phi$ of $s$ vectors for $\mathbb{C}^{d}$ is harmonic if and only if it is generated by a finite abelian group $G \subset \operatorname{Sym}(\Phi)$ of order s, i.e., $\Phi=G \varphi$, for all $\varphi \in \Phi$.

### 2.1.2 The Enumeration Problem

The purpose of this chapter is to count all equivalence classes of prime order harmonic frames. The definition of what it means for two harmonic frames to be equivalent will be given in section 2.2. We start with simpler problem concerning the enumeration of DFT-FUNTFs.

Recall the definition of a DFT-FUNTF given by equation (2.3). A basic way of counting the number of DFT-FUNTFs is inspired by the following observation. For any vector $f \in \mathbb{C}^{d}$, the frame $\Phi$ gives the following representation of $f$ :

$$
\begin{equation*}
f \mapsto\left(\left\langle f, \varphi_{m}\right\rangle\right)_{m=0}^{s-1} \in \mathbb{C}^{s} . \tag{2.12}
\end{equation*}
$$

Therefore, even a re-indexing of the frame would change the representation it gives for a fixed $f$. Thus, we could count the number of ordered DFT-FUNTFs. To accomplish this task, we observe that there are $s$ columns in $D_{s}$ and we select $d$ of them. Since each ordered combination of column choices $n_{1}, \ldots, n_{d}$ gives a distinct frame, there are $s(s-1) \cdots(s-d+1)$ ordered DFT-FUNTFs.

There are of course other ways by which we may distinguish frames, and we shall consider two others here. The first is a natural counterpart to the ordered counting scheme, namely, counting the number of DFT-FUNTFs considered as unordered sets of vectors. The techniques developed for this method will then be expanded to our main goal, which is to count all inequivalent harmonic frames of prime order, where two harmonic frames shall be considered equivalent if one is the unitary transformation of the other. As we shall see, this amounts to counting the number of inequivalent DFT-FUNTFs.

There has been some interest in harmonic frames in the literature, see [20, 33]. In particular, [34] presents a computer program for generating all equivalence classes of harmonic frames for a given $s$ and $d$, where there is a limit on the size of either due to computational considerations. From this program, the authors conjecture that there are $\mathcal{O}\left(s^{d-1}\right)$ inequivalent harmonic frames. The content of this chapter is to not only validate this conjecture for the case when $s$ is a prime number, but in fact give an exact formula for the number of harmonic frames in this case. Furthermore, we examine the structure of prime order harmonic frames via their symmetry group.

An outline of the remainder of chapter 2 is as follows: the rest of section 2.1 reviews some algebraic theory and examines the problem of counting unordered DFT-FUNTFs. Section 2.2 presents the main result of this paper. In section 2.3 we define an equivalence relation that is equivalent to (2.23) and then use this to develop a correspondence between inequivalent harmonic frames and the orbits of a particular set. Section 2.4 counts the number of orbits of this particular set, thus giving a formula for the number of inequivalent harmonic frames. The structure of the symmetry group is handled in section 2.5 , and section 2.6 contains a few concluding remarks.

### 2.1.3 Algebra Review ${ }^{1}$

Recall that we denote the additive group of integers $\bmod s$ by $\mathbb{Z}_{s}$, and set

$$
\begin{equation*}
\mathbb{Z}_{s}^{d}:=\underbrace{\mathbb{Z}_{s} \times \cdots \times \mathbb{Z}_{s}}_{d \text { times }} . \tag{2.13}
\end{equation*}
$$

[^1]Furthermore, let $\mathbb{Z}_{s}^{\times}$denote the group of units of $\mathbb{Z}_{s}$, which, when $s$ is prime, is simply the set $\{1, \ldots, s\}$ endowed with multiplication $\bmod s$. Finally, for $k \in \mathbb{N}$, let $S_{k}$ denote the group of permutations of $k$ elements. We will also need the following definitions and proposition:

Definition 2.1.3. A group action of a group $G$ on a set $X$ is a map $\pi$,

$$
\begin{aligned}
\pi: G \times X & \rightarrow X \\
(g, x) & \mapsto g \cdot x
\end{aligned}
$$

satisfying the following properties:

1) $g_{1} \cdot\left(g_{2} \cdot x\right)=\left(g_{1} g_{2}\right) \cdot x \forall g_{1}, g_{2} \in G, x \in X$,
2) $1 \cdot x=x \forall x \in X$.

Definition 2.1.4. Let $X$ be some set and let $G$ be a group. Furthermore, let $\pi: G \times X \rightarrow X$ be a group action. For each $x \in X$ the stabilizer of $x$ in $G$ is the subgroup of $G$ that fixes the element $x$ :

$$
\begin{equation*}
G_{x}:=\{g \in G: g \cdot x=x\} . \tag{2.14}
\end{equation*}
$$

Proposition 2.1.5. Let $G$ be a group acting on the nonempty set $X$. The relation on $X$ defined by:

$$
x_{1} \sim x_{2} \Longleftrightarrow x_{1}=g \cdot x_{2} \text { for some } g \in G
$$

is an equivalence relation. For each $x \in X$, the number of elements in the equivalence class containing $x$ is $\left|G: G_{x}\right|$, the index of the stabilizer of $x$.

Note, when $G$ is a finite group,

$$
\begin{equation*}
\left|G: G_{x}\right|=\frac{|G|}{\left|G_{x}\right|} \tag{2.15}
\end{equation*}
$$

Definition 2.1.6. Let $G$ be a group acting on the nonempty set $X$. The equivalence class $\mathcal{O}_{x}:=\{g \cdot x: g \in G\}$ is called the orbit of $G$ containing $x$.

As such, the orbits of a group action disjointly partition the set $X$. We are now ready to count the number of prime order DFT-FUNTFs, considered as unordered sets. The basic structure of the argument in subsection 2.1 .4 will be used when we count all harmonic frames of prime order, albeit with added complexity.

### 2.1.4 The Number of Unordered DFT-FUNTFs

It is often the case that we would like to consider a frame as a set, where the order of elements does not matter. Given two ordered DFT-FUNTFs $\Phi=$ $\left(\varphi_{0}, \ldots, \varphi_{s-1}\right)$ and $\Psi=\left(\psi_{0}, \ldots, \psi_{s-1}\right)$, we define the following equivalence relation:

$$
\begin{equation*}
\Phi \sim_{1} \Psi \Longleftrightarrow \exists \sigma \in S_{s} \text { s.t. } \varphi_{m}=\psi_{\sigma(m)}, \quad \forall m=0, \ldots, s-1 . \tag{2.16}
\end{equation*}
$$

(2.16) merely formalizes our consideration of frames as sets. An equivalence class of (2.16) will be denoted in the usual way, that is $\Phi=\left\{\varphi_{0}, \ldots, \varphi_{s-1}\right\}$. In this subsection, we count the number of DFT-FUNTFs of prime order under (2.16). First, however, we must change our perspective on the problem.

Remark 2.1.7. For the rest of chapter 2 we will only consider unordered DFTFUNTFs, and as such from now on $\Phi$ will denote $\left\{\varphi_{0}, \ldots, \varphi_{s-1}\right\}$.

### 2.1.4.1 DFT-FUNTFs and Orbits

First notice that every DFT-FUNTF contains the vector $\varphi_{0}=\frac{1}{\sqrt{d}}(1, \ldots, 1) \in$ $\mathbb{C}^{d}$, and so when comparing two such frames we need not consider this vector. Thus we will only compare sets of the form

$$
\begin{equation*}
\Phi^{\prime}=\Phi-\left\{\varphi_{0}\right\} \tag{2.17}
\end{equation*}
$$

Define the set $\tilde{\mathbb{Z}}_{s}^{d}$ as

$$
\begin{equation*}
\tilde{\mathbb{Z}}_{s}^{d}:=\left\{n=\left(n_{1}, \ldots, n_{d}\right) \in \mathbb{Z}_{s}^{d}: n_{i} \neq n_{j}, \quad \forall i \neq j\right\} . \tag{2.18}
\end{equation*}
$$

There is a one-to-one correspondence between the vectors $\varphi_{m}, m \neq 0$, and the elements of $\tilde{\mathbb{Z}}_{s}^{d}$. Considering $\mathbb{Z}_{s}^{\times}$as a group and $\tilde{\mathbb{Z}}_{s}^{d}$ as a set, we define the group action $\pi_{1}$ as:

$$
\begin{align*}
\pi_{1}: \mathbb{Z}_{s}^{\times} \times \tilde{\mathbb{Z}}_{s}^{d} & \rightarrow \tilde{\mathbb{Z}}_{s}^{d}  \tag{2.19}\\
(m, n) & \mapsto m \cdot n:=\left(m n_{1}, \ldots, m n_{d}\right) .
\end{align*}
$$

The orbits of $\pi_{1}$ are then the sets

$$
\begin{equation*}
\mathcal{O}_{n}=\left\{m \cdot n=\left(m n_{1}, \ldots, m n_{d}\right): m \in \mathbb{Z}_{s}^{\times}\right\}, \quad n \in \tilde{\mathbb{Z}}_{s}^{d} . \tag{2.20}
\end{equation*}
$$

Remark 2.1.8. For clarity of exposition we shall sometimes use $\Phi_{n}$ to denote the DFT-FUNTF $\Phi$ and $\varphi_{m, n}$ its corresponding elements, where the subscript $n$ emphasizes the generators $n=\left(n_{1}, \ldots, n_{d}\right)$.

The following proposition relates the equivalence classes of (2.16) and the orbits of $\pi_{1}$.

Proposition 2.1.9. There is a one-to-one correspondence between the equivalence classes of (2.16) and the orbits of $\pi_{1}$, i.e. the sets $\Phi_{n}$ and $\mathcal{O}_{n}$ can be identified. We denote this identification as:

$$
\begin{equation*}
\Phi_{n}=\left\{\varphi_{0}, \ldots, \varphi_{s-1}\right\} \longleftrightarrow \mathcal{O}_{n} \tag{2.21}
\end{equation*}
$$

Proof. As noted above, we have:

$$
\Phi \longleftrightarrow \Phi^{\prime}=\Phi-\left\{\varphi_{0}\right\} .
$$

Define a function $F$ that maps orbits of $\tilde{\mathbb{Z}}_{s}^{d}$ to sets of the form $\Phi^{\prime}$ as follows:

$$
\begin{equation*}
F\left(\mathcal{O}_{n}\right)=\left\{\varphi_{m, n}\right\}_{m=1}^{s} \tag{2.22}
\end{equation*}
$$

We must show that $F$ is both one-to-one and onto, however it is clear that $F$ is surjective. Considering then the former, suppose $F\left(\mathcal{O}_{n}\right)=F\left(\mathcal{O}_{n^{\prime}}\right)$. This would imply that $\left\{\varphi_{m, n}\right\}_{m=1}^{s}=\left\{\varphi_{m^{\prime}, n^{\prime}}\right\}_{m^{\prime}=1}^{s}$. But then for some $m$ and some $m^{\prime}$, we would have $\left(m n_{1}, \ldots, m n_{d}\right)=\left(m^{\prime} n_{1}^{\prime}, \ldots, m^{\prime} n_{d}^{\prime}\right)$, i.e. $\mathcal{O}_{n} \cap \mathcal{O}_{n^{\prime}} \neq \emptyset$, and so in fact $\mathcal{O}_{n}=\mathcal{O}_{n^{\prime}}$.

Remark 2.1.10. Given the content of proposition 2.1.9, we now replace the problem of counting the equivalence classes of (2.16) with the problem of counting the orbits of $\pi_{1}$.

### 2.1.4.2 The Number of Orbits of $\pi_{1}$

By proposition 2.1.5 we see that the orbits of a group action partition the set into disjoint equivalence classes. In particular, the orbits $\mathcal{O}_{n}$ partition the set $\tilde{\mathbb{Z}}_{s}^{d}$.

Furthermore, the size of each $\mathcal{O}_{n}$ is given by $\left|\mathcal{O}_{n}\right|=\left|\mathbb{Z}_{s}^{\times}:\left(\mathbb{Z}_{s}^{\times}\right)_{n}\right|$. Using these facts, we prove the following proposition.

Proposition 2.1.11. Let $s$ be a prime number and $d \leq s$. Then the number of orbits of $\pi_{1}$ is:

1) 2 , if $d=1$ or $d=s=2$.
2) $s(s-2) \cdots(s-d+1), \quad$ if $d \geq 2, s>2$.

Proof. We first consider the case $d=1$. For $n=0$ we have $\left(\mathbb{Z}_{s}^{\times}\right)_{0}=\mathbb{Z}_{s}^{\times}$, and so $\left|\mathcal{O}_{0}\right|=(s-1) /(s-1)=1$. For $n \neq 0$ we see $\left(\mathbb{Z}_{s}^{\times}\right)_{n}=\{1\}$, and thus $\left|\mathcal{O}_{n}\right|=s-1$. Since $\left|\tilde{\mathbb{Z}}_{s}^{1}\right|=s$, there are only two orbits.

Now take $2 \leq d \leq s$. For each $n \in \tilde{\mathbb{Z}}_{s}^{d}$ we have $\left(\mathbb{Z}_{s}^{\times}\right)_{n}=\{1\}$, and thus $\left|\mathcal{O}_{n}\right|=s-1$. Therefore the number of orbits is given by $\gamma$, where

$$
\begin{aligned}
\left|\tilde{\mathbb{Z}}_{s}^{d}\right| & =\gamma\left|\mathcal{O}_{n}\right| \\
s(s-1) \cdots(s-d+1) & =\gamma(s-1)
\end{aligned}
$$

For $s=2$ and $d=2$, we see $\gamma=2$. For $s>2$ we have $\gamma=s(s-2) \cdots(s-d+1)$.

As an addendum to theorem 2.1.11, we note that one of the orbits in the $d=1$ case corresponds to a degenerate DFT-FUNTF. Namely, the orbit $\mathcal{O}_{0}$ corresponds to the DFT-FUNTF consisting of the single element $\{1\}$.

### 2.2 The Number of Harmonic Frames of Prime Order

Using a similar correspondence between harmonic frames and orbits, we count all harmonic frames of prime order up to unitary transformations.

Two harmonic frames $\Phi=\left\{\varphi_{0}, \ldots, \varphi_{s-1}\right\} \subset \mathbb{C}^{d}$ and $\Psi=\left\{\psi_{0}, \ldots, \psi_{s-1}\right\} \subset \mathbb{C}^{d}$ are said to be equivalent if the following equivalence relation holds:

$$
\begin{equation*}
\Phi \sim_{2} \Psi \Longleftrightarrow \exists U \in \mathcal{U}\left(\mathbb{C}^{d}\right) \text { and } P \in \mathcal{P}_{s} \text { s.t. } \Phi=U \Psi P . \tag{2.23}
\end{equation*}
$$

Note that we have used matrix notation for the left hand side of (2.23). In terms of sets, the condition merely states that

$$
\begin{equation*}
\left\{\varphi_{i}: i=1, \ldots s\right\}=\left\{U \psi_{i}: i=1, \ldots, s\right\} . \tag{2.24}
\end{equation*}
$$

(2.23) is a standard form of equivalence in much of the literature when dealing with frames. Recently, [34] conjectured that the number of inequivalent harmonic frames is $O\left(s^{d-1}\right)$. We prove this conjecture for $s$ a prime number as a corollary to theorem 2.2.1, which gives an exact formula for the number of harmonic frames. The proof of theorem 2.2.1 is handled in section 2.4, with much preliminary work accomplished in section 2.3.

For a fixed $s$ and $d$, we backwards recursively define the set

$$
\begin{equation*}
\left\{\alpha_{c} \in \mathbb{N} \cup\{0\}: c \in \mathbb{N}, c \mid s-1, \text { and } c \mid d \text { or } c \mid d-1\right\} \tag{2.25}
\end{equation*}
$$

If $c|s-1, c| d$, and $c>1$, then

$$
\begin{equation*}
\alpha_{c}:=\frac{(s-1-c)(s-1-2 c) \cdots\left(s-1-\left(\frac{d}{c}-1\right) c\right)}{c^{\frac{d}{c}-1}(d / c)!}-\frac{c}{s-1} \sum_{\substack{c<b<s \\ c|b, b| d}}\left(\frac{s-1}{b}\right) \alpha_{b}, \tag{2.2.26d}
\end{equation*}
$$

where we have used the notation $(2.2 .26 \mathrm{~d})$ to emphasize its dependence on the condition $c \mid d$. If $c|s-1, c| d-1$, and $c>1$, then

$$
\begin{equation*}
\alpha_{c}:=\frac{(s-1-c)(s-1-2 c) \cdots\left(s-1-\left(\frac{d-1}{c}-1\right) c\right)}{c^{\frac{d-1}{c}-1}((d-1) / c)!}-\frac{c}{s-1} \sum_{\substack{c<b<s \\ c|b, b| d-1}}\left(\frac{s-1}{b}\right) \alpha_{b} . \tag{2.2.26d-1}
\end{equation*}
$$

Finally, $\alpha_{1}$ is defined as:

$$
\begin{equation*}
\alpha_{1}:=\frac{1}{s-1}\binom{s}{d}-\sum_{\substack{c \mid d \\ c>1}} \frac{\alpha_{c}}{c}-\sum_{\substack{c \mid d-1 \\ c>1}} \frac{\alpha_{c}}{c} . \tag{2.2.27}
\end{equation*}
$$

Theorem 2.2.1. Let $s$ be a prime number and let $1<d<s$. Define the set

$$
\left\{\alpha_{c} \in \mathbb{N} \cup\{0\}: c \in \mathbb{N}, c \mid s-1, \text { and } c \mid d \text { or } c \mid d-1\right\}
$$

as in equations (2.2.26d), (2.2.26 d-1), and (2.2.27). The total number of harmonic frames for $\mathbb{C}^{d}$ with $s$ elements is then given by:

$$
\begin{equation*}
\alpha_{1}+\sum_{\substack{c \mid d \\ c>1}} \alpha_{c}+\sum_{\substack{c \mid d-1 \\ c>1}} \alpha_{c} . \tag{2.2.28}
\end{equation*}
$$

More concisely, we have the following corollary:

Corollary 2.2.2. Let $s$ be any prime number and fix $d$ such that $1<d<s$. Then the number of inequivalent harmonic frames for $\mathbb{C}^{d}$ with $s$ elements is $O\left(s^{d-1}\right)$.

Proof. Using equations $(2.2 .26 d)$ and $(2.2 .26 d-1)$, we see that $\alpha_{c}=O\left(s^{d^{\prime}}\right)$, where $c>1$ and $d^{\prime} \leq \frac{d}{c}-1<d-1$. Therefore, by (2.2.27), we see that $\alpha_{1}=O\left(s^{d-1}\right)$, and the corollary follows.

In the above theorems, the case $d=1$ is omitted, however, it is not hard to see that there are two inequivalent harmonic frames in this case; in fact, there is only one inequivalent harmonic frame for $d=1$ with $s$ distinct vectors.

### 2.3 Harmonic Frames and Orbits

In this section we develop a one-to-one correspondence between inequivalent harmonic frames and the orbits of a particular set, not unlike the ideas presented in
subsection 2.1.4. First, however, we come up with an equivalent condition to (2.23).
We will assume $s$ is prime for the remainder of chapter 2.

### 2.3.1 A New Equivalence Relation

When $s$ is prime, every harmonic frame is of the form $U \Phi$, where $U \in \mathcal{U}\left(\mathbb{C}^{d}\right)$ and $\Phi$ is a DFT-FUNTF (see section 2.1.1). Therefore, finding the number of inequivalent harmonic frames amounts to finding the number of inequivalent DFTFUNTFs. Toward that end, we simplify (2.23) to the following:

Theorem 2.3.1. If $s$ is prime and $\Phi=\left\{\varphi_{0}, \ldots, \varphi_{s-1}\right\}$ and $\Psi=\left\{\psi_{0}, \ldots, \psi_{s-1}\right\}$ are DFT-FUNTFs, then

$$
\begin{array}{r}
\exists \sigma_{1} \in S_{s}, \sigma_{2} \in S_{d} \text { such that } \\
\Phi \sim_{2} \Psi \Longleftrightarrow \quad \varphi_{m}(k)=\psi_{\sigma_{1}(m)}\left(\sigma_{2}(k)\right)  \tag{2.3.29}\\
\forall m=0, \ldots, s-1, \quad k=1, \ldots, d
\end{array}
$$

where $\varphi_{m}(k)$ denotes the $k^{\text {th }}$ element of the vector $\varphi_{m}$.

Proof. It is clear that if the right hand side of (2.3.29) holds, then the left hand side must hold as well. Assume then that $\Phi \sim_{2} \Psi$, i.e., there exists a $U \in \mathcal{U}\left(\mathbb{C}^{d}\right)$ and $P_{\sigma} \in \mathcal{P}_{s}$ such that

$$
\begin{equation*}
\Phi=U \Psi P_{\sigma} \tag{2.3.30}
\end{equation*}
$$

Let $\sigma \in S_{s}$ be the permutation associated with $P_{\sigma}$, and note that (2.3.30) implies that

$$
\begin{equation*}
\varphi_{m}=U \psi_{\sigma(m)}, \quad \forall m=0, \ldots, s-1 \tag{2.3.31}
\end{equation*}
$$

Without loss of generality, we may assume that $\sigma(0)=0$. Indeed, by theorem 2.1.2 there exists a $U_{0} \in \operatorname{Sym}(\Psi)$ such that $U_{0} \psi_{0}=\psi_{\sigma(0)}$. By definition, $U_{0}$ is a $d \times d$ matrix that permutes the columns of $\Psi$ by acting on the left. Therefore, there exists an $s \times s$ permutation matrix $P_{U_{0}}$ that permutes the columns of $\Psi$ in the exact same manner, yet acts on the right. In particular, $U_{0} \Psi=\Psi P_{U_{0}}$, and thus

$$
\begin{equation*}
\Phi=U U_{0} \Psi P_{U_{0}}^{-1} P_{\sigma} \tag{2.3.32}
\end{equation*}
$$

Set $V:=U U_{0}$ and $P:=P_{U_{0}}^{-1} P_{\sigma}$. It is clear that $V$ is a unitary transformation and that $P$ is its associated permutation matrix. Furthermore, $\varphi_{0}=V \psi_{0}$, and so we can assume from the start that $\varphi_{0}=U \psi_{0}$, i.e., that $\sigma(0)=0$.

Now let $n_{1}, \ldots, n_{d}$ denote the column choices of $\Phi$, and consider the following:

$$
\begin{equation*}
\left\langle\varphi_{m}, \varphi_{0}\right\rangle=\sum_{k=1}^{d} e^{2 \pi i m n_{k} / s} \tag{2.3.33}
\end{equation*}
$$

Letting $l_{1}, \ldots, l_{d}$ denote the column choices of $\Psi$, we also have:

$$
\begin{equation*}
\left\langle\varphi_{m}, \varphi_{0}\right\rangle=\left\langle U \psi_{\sigma(m)}, U \psi_{0}\right\rangle=\left\langle\psi_{\sigma(m)}, \psi_{0}\right\rangle=\sum_{k=1}^{d} e^{2 \pi i \sigma(m) l_{k} / s} \tag{2.3.34}
\end{equation*}
$$

Define $p_{\varphi}, p_{\psi} \in \mathbb{Z}[z] /\left\langle z^{s}\right\rangle$ as follows:

$$
\begin{equation*}
p_{\varphi}(z):=\sum_{k=1}^{d} z^{m n_{k}} \quad \text { and } \quad p_{\psi}(z):=\sum_{k=1}^{d} z^{\sigma(m) l_{k}} \tag{2.3.35}
\end{equation*}
$$

By equations (2.3.33) and (2.3.34), we see that $p_{\varphi}(z)=p_{\psi}(z)$ when $z=e^{2 \pi i / s}$. In other words, $z=e^{2 \pi i / s}$ is a root of the polynomial $p(z):=p_{\varphi}(z)-p_{\psi}(z)$. However, since $p \in \mathbb{Z}[z] /\left\langle z^{s}\right\rangle$, and the minimum polynomial of $z=e^{2 \pi i / s}$ is $q(z):=\sum_{k=0}^{s-1} z^{k}$, $p$ must either be an integer multiple of $q$ or the zero polynomial. It is clear, though, that only the latter option is feasible, thus giving

$$
\begin{equation*}
p_{\varphi}(z)=p_{\psi}(z) \tag{2.3.36}
\end{equation*}
$$

Combining equations (2.3.35) and (2.3.36), we see there exists a $\sigma_{2} \in S_{d}$ such that

$$
\begin{equation*}
m n_{k}=\sigma(m) l_{\sigma_{2}(k)}, \quad \forall k=1, \ldots, d \tag{2.3.37}
\end{equation*}
$$

Note that $\sigma_{2}$ is dependent on the choice of $m$. Taking $m=1$ in (2.3.37), one has $n_{k}=\sigma(1) l_{\sigma_{2}(k)}$. Letting $\sigma_{1}(m):=\sigma(1) m$, we have:

$$
\begin{equation*}
\varphi_{m}=\left(e^{2 \pi i m n_{k} / s}\right)_{k=1}^{d}=\left(e^{2 \pi i \sigma_{1}(m) l_{\sigma_{2}}(k)}\right)_{k=1}^{d}=\psi_{\sigma_{1}(m)}\left(\sigma_{2}(k)\right) . \tag{2.3.38}
\end{equation*}
$$

### 2.3.2 Inequivalent DFT-FUNTFs and Orbits

Similar to section 2.1.4.1, we now develop a one-to-one correspondence between inequivalent DFT-FUNTFs and the orbits of a particular set. As a matter of notation, we shall denote equivalence classes of (2.23) by $[\Phi]$, where $\Phi=\left\{\varphi_{0}, \ldots, \varphi_{s-1}\right\}$ is a DFT-FUNTF representative. By theorem 2.3.1, the equivalence classes of (2.23) are identical to the equivalence classes of the right hand side of (2.3.29). We now turn our attention to the set with which we will identify the equivalence classes $[\Phi]$. Consider the following equivalence relation on the set $\tilde{\mathbb{Z}}_{s}^{d}$,

$$
\begin{equation*}
\left(n_{1}, \ldots, n_{d}\right) \sim\left(n_{1}^{\prime}, \ldots, n_{d}^{\prime}\right) \Longleftrightarrow \exists \sigma \in S_{d} \text { s.t. }\left(n_{1}, \ldots, n_{d}\right)=\left(n_{\sigma(1)}^{\prime}, \ldots, n_{\sigma(d)}^{\prime}\right) \tag{2.3.39}
\end{equation*}
$$

Denote an equivalence class of (2.3.39) by the representative $[n]=\left[n_{1}, \ldots, n_{d}\right]$, and define $\mathbb{A}_{s}^{d}$ as the set of all equivalence classes, i.e.

$$
\begin{equation*}
\mathbb{A}_{s}^{d}:=\tilde{\mathbb{Z}}_{s}^{d} / \sim \tag{2.3.40}
\end{equation*}
$$

It is easy to see $\left|\mathbb{A}_{s}^{d}\right|=\binom{s}{d}$. Considering $\mathbb{Z}_{s}^{\times}$as a group and $\mathbb{A}_{s}^{d}$ as a set, we define the group action $\pi_{2}$,

$$
\begin{align*}
\pi_{2}: \mathbb{Z}_{s}^{\times} \times \mathbb{A}_{s}^{d} & \rightarrow \mathbb{A}_{s}^{d}  \tag{2.3.41}\\
(m,[n]) & \mapsto m \cdot[n]:=\left[m n_{1}, \ldots, m n_{d}\right]
\end{align*}
$$

The orbits of $\pi_{2}$ are the sets $\mathcal{O}_{[n]}=\left\{m \cdot[n]=\left[m n_{1}, \ldots, m n_{d}\right]: m \in \mathbb{Z}_{s}^{\times}\right\}$. The following proposition relates the equivalence classes of (2.23) and the orbits of $\pi_{2}$.

Proposition 2.3.2. There is a one-to-one correspondence between the equivalences classes of (2.23) and the orbits of $\pi_{2}$, i.e.

$$
\begin{equation*}
\left[\Phi_{n}\right] \longleftrightarrow \mathcal{O}_{[n]} \tag{2.3.42}
\end{equation*}
$$

Proof. Define the function $F$ as follows:

$$
\begin{equation*}
F\left(\left[\Phi_{n}\right]\right)=\mathcal{O}_{[n]}=\left\{\left[m n_{1}, \ldots, m n_{d}\right]: m \in \mathbb{Z}_{s}^{\times}\right\} \tag{2.3.43}
\end{equation*}
$$

We must show that $F$ is well defined, one-to-one, and onto. Surjectivity is clear, so we focus on the first two. To show $F$ is well defined, suppose that $\left[\Phi_{n}\right]=\left[\Psi_{n^{\prime}}\right]$. We want to show $F\left(\left[\Phi_{n}\right]\right)=F\left(\left[\Psi_{n^{\prime}}\right]\right)$, i.e. $\mathcal{O}_{[n]}=\mathcal{O}_{\left[n^{\prime}\right]}$. We have:

$$
\begin{aligned}
{\left[\Phi_{n}\right]=\left[\Psi_{n^{\prime}}\right] } & \Longleftrightarrow \varphi_{m}(k)=\psi_{\sigma_{1}(m)}\left(\sigma_{2}(k)\right) \forall k=1, \ldots, d, \forall m=0, \ldots, s-1 \\
& \Longleftrightarrow\left\{\varphi_{0}(k)_{k=1}^{d}, \ldots, \varphi_{s-1}(k)_{k=1}^{d}\right\}=\left\{\psi_{0}\left(\sigma_{2}(k)\right)_{k=1}^{d}, \ldots, \psi_{s-1}\left(\sigma_{2}(k)\right)_{k=1}^{d}\right\} \\
& \Longleftrightarrow\left\{\varphi_{1}(k)_{k=1}^{d}, \ldots, \varphi_{s-1}(k)_{k=1}^{d}\right\}=\left\{\psi_{1}\left(\sigma_{2}(k)\right)_{k=1}^{d}, \ldots, \psi_{s-1}\left(\sigma_{2}(k)\right)_{k=1}^{d}\right\} \\
& \Longleftrightarrow\left\{\left(m n_{1}, \ldots, m n_{d}\right): m \in \mathbb{Z}_{s}^{\times}\right\}=\left\{\left(m n_{\sigma_{2}(1)}^{\prime}, \ldots, m n_{\sigma_{2}(d)}^{\prime}\right): m \in \mathbb{Z}_{s}^{\times}\right\} \\
& \Longleftrightarrow\left\{\left[m n_{1}, \ldots, m n_{d}\right]: m \in \mathbb{Z}_{s}^{\times}\right\}=\left\{\left[m n_{1}^{\prime}, \ldots, m n_{d}^{\prime}\right]: m \in \mathbb{Z}_{s}^{\times}\right\} \\
& \Longleftrightarrow \mathcal{O}_{[n]}=\mathcal{O}_{\left[n^{\prime}\right]},
\end{aligned}
$$

where the first equivalence is due to theorem 2.3.1, and the third equivalence is because $\varphi_{0}=\psi_{0}=\frac{1}{\sqrt{d}}(1, \ldots, 1)$.

To prove injectivity, we assume $\mathcal{O}_{[n]}=\mathcal{O}_{\left[n^{\prime}\right]}$. According to this assumption, there must exist an $m_{0}^{\prime} \in \mathbb{Z}_{s}^{\times}$such that $\left[n_{1}, \ldots, n_{d}\right]=\left[m_{0}^{\prime} n_{1}^{\prime}, \ldots, m_{0}^{\prime} n_{d}^{\prime}\right]$. Therefore we have:

$$
\begin{aligned}
\mathcal{O}_{[n]}=\mathcal{O}_{\left[n^{\prime}\right]} & \Longleftrightarrow\left[n_{1}, \ldots, n_{d}\right]=\left[m_{0}^{\prime} n_{1}^{\prime}, \ldots, m_{0}^{\prime} n_{d}^{\prime}\right] \\
& \Longleftrightarrow\left(n_{1}, \ldots, n_{d}\right)=\left(m_{0}^{\prime} n_{\sigma_{2}(1)}^{\prime}, \ldots, m_{0}^{\prime} n_{\sigma_{2}(d)}^{\prime}\right) \\
& \Longleftrightarrow\left(m n_{1}, \ldots, m n_{d}\right)=\left(m m_{0}^{\prime} n_{\sigma_{2}(1)}^{\prime}, \ldots, m m_{0}^{\prime} n_{\sigma_{2}(d)}^{\prime}\right), \forall m \in \mathbb{Z}_{s}^{\times} \\
& \Longleftrightarrow\left\{\left(m n_{1}, \ldots, m n_{d}\right): m \in \mathbb{Z}_{s}^{\times}\right\}=\left\{\left(m n_{\sigma_{2}(1)}^{\prime}, \ldots, m n_{\sigma_{2}(d)}^{\prime}\right): m \in \mathbb{Z}_{s}^{\times}\right\} \\
& \Longleftrightarrow\left\{\varphi_{1}(k)_{k=1}^{d}, \ldots, \varphi_{s-1}(k)_{k=1}^{d}\right\}=\left\{\psi_{1}\left(\sigma_{2}(k)\right)_{k=1}^{d}, \ldots, \psi_{s-1}\left(\sigma_{2}(k)\right)_{k=1}^{d}\right\} \\
& \Longleftrightarrow\left\{\varphi_{0}(k)_{k=1}^{d}, \ldots, \varphi_{s-1}(k)_{k=1}^{d}\right\}=\left\{\psi_{0}\left(\sigma_{2}(k)\right)_{k=1}^{d}, \ldots, \psi_{s-1}\left(\sigma_{2}(k)\right)_{k=1}^{d}\right\} \\
& \Longleftrightarrow \varphi_{m}(k)=\psi_{\sigma_{1}(m)}\left(\sigma_{2}(k)\right), \forall k=1, \ldots, d, m=0, \ldots, s-1 \\
& \Longleftrightarrow\left[\Phi_{n}\right]=\left[\Psi_{n^{\prime}}\right],
\end{aligned}
$$

where the fourth equivalence uses the fact that $\left\{m m_{0}^{\prime}: m \in \mathbb{Z}_{s}^{\times}\right\}=\{m: m \in$ $\left.\mathbb{Z}_{s}^{\times}\right\}$.

To conclude this section, we note that when $d=s$, we see $\left|\mathbb{A}_{s}^{d}\right|=1$, and so there can be only one orbit. Thus there is only one harmonic frame in this case.

### 2.4 The Number of Orbits of $\mathbb{A}_{s}^{d}$

We begin by counting the number of orbits of $\mathbb{A}_{s}^{d}$ under the group action $\pi_{2}$ for the cases $d=2$ and $d=3$. We then generalize these results for all $1<d<s$.

### 2.4.1 Some Examples: $\mathrm{d}=2$ and $\mathrm{d}=3$

Proposition 2.4.1. Let $s$ be an odd prime number and let $d=2$. Then there are $(s+1) / 2$ orbits of $\mathbb{A}_{s}^{2}$. Therefore, there are $(s+1) / 2$ inequivalent harmonic frames for $\mathbb{C}^{2}$.

Proof. Let $[n] \in \mathbb{A}_{s}^{2}$. If $\left(\mathbb{Z}_{s}^{\times}\right)_{[n]}=\{1\}$, then $\left|\mathcal{O}_{[n]}\right|=s-1$. Therefore, if we can find all $[n] \in \mathbb{A}_{s}^{2}$ with non-trivial stabilizer and their corresponding orbits, we will be able to solve for the total number of orbits. Assume that $m \cdot\left[n_{1}, n_{2}\right]=\left[m n_{1}, m n_{2}\right]=\left[n_{1}, n_{2}\right]$ for some $m \neq 1$. This implies that

$$
\begin{aligned}
m n_{1} & \equiv n_{2} \bmod s, \\
m n_{2} & \equiv n_{1} \bmod s .
\end{aligned}
$$

Combining the above equations yields

$$
\begin{aligned}
m^{2} n_{1} & \equiv n_{1} \bmod s \\
\Rightarrow \quad m & \equiv \pm 1 \bmod s
\end{aligned}
$$

Thus we see that we can take $m \equiv-1 \bmod s$, which implies $n_{2} \equiv-n_{1} \bmod s$. Therefore all $[n] \in \mathbb{A}_{s}^{2}$ of the form $[n]=\left[n_{1},-n_{1}\right], n_{1} \neq 0$, have stabilizer $\{1,-1\}$. Furthermore, since

$$
\begin{equation*}
\mathcal{O}_{[1,-1]}=\left\{m \cdot[1,-1]=[m,-m]: m \in \mathbb{Z}_{s}^{\times}\right\} \tag{2.4.44}
\end{equation*}
$$

we see that all such $[n]$ lie in the orbit $\mathcal{O}_{[1,-1]}$. Finally, these are the only elements of $\mathbb{A}_{s}^{2}$ with nontrivial stabilizer, and thus the number of orbits of $\mathbb{A}_{s}^{2}$ is $\gamma_{1}+1$, where
$\gamma_{1}$ is the number of orbits of size $s-1$. Therefore,

$$
\begin{aligned}
\left|\mathbb{A}_{s}^{2}\right| & =\gamma_{1}(s-1)+\left|\mathcal{O}_{(1,-1)}\right| \\
\binom{s}{2} & =\gamma_{1}(s-1)+(s-1) / 2 \\
s(s-1) / 2 & =\gamma_{1}(s-1)+(s-1) / 2
\end{aligned}
$$

Solving for $\gamma_{1}$ we get $\gamma_{1}=(s-1) / 2$ and so $\mathbb{A}_{s}^{2}$ has $\gamma_{1}+1=(s-1) / 2+1=$ $(s+1) / 2$ orbits.

Proposition 2.4.2. Let $s$ be a prime number, $s>3$, and let $d=3$ :

1. If $s \equiv 1 \bmod 3$, then there are $\left(s^{2}-2 s+7\right) / 6$ orbits of $\mathbb{A}_{s}^{3}$.
2. If $s \equiv 2 \bmod 3$, then there are $\left(s^{2}-2 s+3\right) / 6$ orbits of $\mathbb{A}_{s}^{3}$.

Therefore, if $s \equiv 1 \bmod 3$, there are $\left(s^{2}-2 s+7\right) / 6$ inequivalent harmonic frames for $\mathbb{C}^{3}$, and if $s \equiv 2 \bmod 3$, there are $\left(s^{2}-2 s+3\right) / 6$ inequivalent harmonic frames for $\mathbb{C}^{3}$.

Proof. As in the proof of proposition 2.4.1, we are looking for all $[n] \in \mathbb{A}_{s}^{3}$ with non-trivial stabilizer and their corresponding orbits. So again we suppose

$$
\begin{equation*}
m \cdot\left[n_{1}, n_{2}, n_{3}\right]=\left[m n_{1}, m n_{2}, m n_{3}\right]=\left[n_{1}, n_{2}, n_{3}\right] \tag{2.4.45}
\end{equation*}
$$

for some $m \neq 1$. We now consider two cases:
I: Suppose $n_{1}=0$. Then we want $m \cdot\left[0, n_{2}, n_{3}\right]=\left[0, m n_{2}, m n_{3}\right]=\left[0, n_{2}, n_{3}\right]$. But this is just the same situation as the $d=2$ case, and so the elements of $\mathbb{A}_{s}^{3}$ of this form with non-trivial stabilizer all lie in the following orbit:

$$
\begin{align*}
\mathcal{O}_{[0,1,-1]} & =\left\{m \cdot[0,1,-1]=[0, m,-m]: m \in \mathbb{Z}_{s}^{\times}\right\}  \tag{2.4.46}\\
\left|\mathcal{O}_{[0,1,-1]}\right| & =(s-1) / 2
\end{align*}
$$

II: Suppose $n_{k} \neq 0$ for all $k=1,2,3$. According to (2.4.45), we have three options for the value of $m n_{1}$ :

$$
m n_{1} \equiv\left\{\begin{array}{c}
n_{1} \bmod s  \tag{2.4.47}\\
n_{2} \bmod s \\
n_{3} \bmod s
\end{array}\right.
$$

If $m n_{1} \equiv n_{1} \bmod s$, then $m=1$ is the only solution, which is trivial and so we disregard this case. Since the order of elements does not matter in $\mathbb{A}_{s}^{3}$, there is no difference between $m n_{1} \equiv n_{2} \bmod s$ and $m n_{1} \equiv n_{3} \bmod s$, and so we choose the former. Moving on to the value of $m n_{2}$, we once again have the same three options. However, $m n_{2} \equiv n_{1} \bmod s$, combined with $m n_{1} \equiv n_{2} \bmod s$ would imply that $m n_{3} \equiv n_{3} \bmod s$, thus resulting in $m=1 . m n_{2} \equiv n_{2} \bmod s$ not only would imply $m=1$, but since $m n_{1} \equiv n_{2} \bmod s$, would also lead to a contradiction. Therefore $m n_{2} \equiv n_{3} \bmod s$ must hold, which in turn forces $m n_{3} \equiv n_{1} \bmod s$. Summarizing, we have

$$
\begin{align*}
m n_{1} & \equiv n_{2} \bmod s \\
m n_{2} & \equiv n_{3} \bmod s,  \tag{2.4.48}\\
m n_{3} & \equiv n_{1} \bmod s
\end{align*}
$$

Proceeding in a similar fashion to the proof of proposition 2.4.1, we see that (2.4.48) implies

$$
\begin{equation*}
m^{3} n_{1} \equiv n_{1} \bmod s \tag{2.4.49}
\end{equation*}
$$

We now find all $m \in \mathbb{Z}_{s}^{\times}$that satisfy (2.4.49). Naturally $m=1$ works; for the remaining solutions, let $g$ be any primitive root $\bmod s$, i.e. $\langle g\rangle=\mathbb{Z}_{s}^{\times}$. Then all
nontrivial solutions to (2.4.49) are of the form

$$
\begin{equation*}
m \equiv g^{(s-1) / 3} \bmod s \quad \text { or } \quad m \equiv g^{2(s-1) / 3} \bmod s \tag{2.4.50}
\end{equation*}
$$

We have two cases:
II.a: If 3 does not divide $s-1$, i.e. $s \equiv 2 \bmod 3$, then the only solution to (2.4.49) is $m=1$.
II.b: If 3 does divide $s-1$, i.e. $s \equiv 1 \bmod 3$, then the solution set to (2.4.49) is:

$$
\begin{equation*}
\left\{1, g^{(s-1) / 3}, g^{2(s-1) / 3}: g \text { is a primitive root } \bmod s\right\} \tag{2.4.51}
\end{equation*}
$$

Therefore all elements in $\mathbb{A}_{s}^{3}$ of the form $\left[n_{1}, g^{(s-1) / 3} n_{1}, g^{2(s-1) / 3} n_{1}\right], n_{1} \neq 0$, have stabilizer $\left\{1, g^{(s-1) / 3}, g^{2(s-1) / 3}\right\}$. Furthermore, all elements of this form lie in the following orbit:

$$
\begin{equation*}
\mathcal{O}_{\left[1, g^{(s-1) / 3}, g^{2(s-1) / 3}\right]}=\left\{\left[m, m g^{(s-1) / 3}, m g^{2(s-1) / 3}\right]: m \in \mathbb{Z}_{s}^{\times}\right\} \tag{2.4.52}
\end{equation*}
$$

where

$$
\begin{equation*}
\left|\mathcal{O}_{\left[1, g^{(s-1) / 3}, g^{2(s-1) / 3}\right]}\right|=(s-1) / 3 \tag{2.4.53}
\end{equation*}
$$

Indeed, since we have assumed that $n_{1} \neq 0$, there are $s-1$ choices for $n_{1}$. However, since the order of elements in the 3 -tuple does not matter, choosing $n_{1}$ is the same as choosing $g^{(s-1) / 3} n_{1}$ or $g^{2(s-1) / 3} n_{1}$. Therefore there are $(s-1) / 3$ elements of this form, and they must all lie in the orbit $\mathcal{O}_{\left[1, g^{(s-1) / 3}, g^{2(s-1) / 3]}\right.}$. Using the same techniques as in proposition 2.4.1, we may now count the number of orbits (recall that $\gamma_{1}$ is the number of orbits of size $s-1$ ):

1. If $s \equiv 1 \bmod 3$, then there are $\gamma_{1}+2$ orbits:

$$
\left|\mathbb{A}_{s}^{3}\right|=\gamma_{1}(s-1)+(s-1) / 2+(s-1) / 3 .
$$

Solving for $\gamma_{1}$ we get $\gamma_{1}+2=\left(s^{2}-2 s+7\right) / 6$.
2. If $s \equiv 2 \bmod 3$, then there are $\gamma_{1}+1$ orbits:

$$
\left|\mathbb{A}_{s}^{3}\right|=\gamma_{1}(s-1)+(s-1) / 2 .
$$

Solving for $\gamma_{1}$ we get $\gamma_{1}+1=\left(s^{2}-2 s+3\right) / 6$.

### 2.4.2 The Structure of the Orbits of $\mathbb{A}_{s}^{d}$

We now turn our attention to the more general setting, beginning with the following theorem which addresses the order of the orbits of $\mathbb{A}_{s}^{d}$ and the form of the elements in the orbits.

Theorem 2.4.3. Let $s$ be a prime number and let $1<d<s$. If $\mathcal{O}$ is an orbit of $\mathbb{A}_{s}^{d}$ under the group action $\pi_{2}$, then there exists $c \in \mathbb{N}$ such that $c \mid d$ or $c \mid d-1$, and

$$
\begin{equation*}
|\mathcal{O}|=(s-1) / c . \tag{2.4.54}
\end{equation*}
$$

Furthermore, let $g$ be a primitive root mod $s$ and set

$$
\begin{equation*}
n_{k}^{c}:=\left[n_{k}, g^{(s-1) / c} n_{k}, \ldots, g^{(c-1)(s-1) / c} n_{k}\right], \quad n_{k} \neq 0 \tag{2.4.55}
\end{equation*}
$$

If $[n] \in \mathcal{O}$, then $[n]$ can be written in the form

$$
[n]= \begin{cases}{\left[n_{1}^{c}, n_{2}^{c}, \ldots, n_{d / c}^{c}\right]} & \text { if } c \mid d  \tag{2.4.56c}\\ {\left[0, n_{1}^{c}, n_{2}^{c}, \ldots, n_{(d-1) / c}^{c}\right]} & \text { if } c \mid d-1\end{cases}
$$

Proof. Let $m \in \mathbb{Z}_{s}^{\times}$; we determine which elements of $\mathbb{A}_{s}^{d}$ are stabilized by $m$ based on the order of $m$. In particular, we will break the argument into two cases: $|m|=c>d$ and $|m|=c \leq d$. We begin with the former.
I. Assume $|m|=c>d$.

We show that no element in $\mathbb{A}_{s}^{d}$ can be stabilized by $m$. Let $[n]=\left[n_{1}, \ldots, n_{d}\right] \in$ $\mathbb{A}_{s}^{d}, n_{j} \neq 0$ for all $j=1, \ldots, d$, and suppose

$$
\begin{aligned}
m \cdot[n] & =[n], \\
\Longrightarrow m \cdot\left[n_{1}, \ldots, n_{d}\right] & =\left[n_{1}, \ldots, n_{d}\right], \\
\Longrightarrow\left[m n_{1}, \ldots, m n_{d}\right] & =\left[n_{1}, \ldots, n_{d}\right] .
\end{aligned}
$$

Therefore, $m n_{1} \equiv n_{j} \bmod s$ for some $j \in\{1, \ldots, d\}$, and because the order of $n_{1}, \ldots, n_{d}$ does not matter, without loss of generality we have two choices:

$$
m n_{1} \equiv\left\{\begin{array}{c}
n_{1} \bmod s  \tag{2.4.57}\\
n_{2} \bmod s
\end{array}\right.
$$

If $m n_{1} \equiv n_{1} \bmod s$, then $m=1$ and we have a contradiction to the assumption $|m|=c>d$. Therefore, $m n_{1} \equiv n_{2} \bmod s$ must hold. Continuing, we see that $m n_{2} \equiv n_{j} \bmod s$ for some $j \in\{1, \ldots, d\}$. Without loss of generality, we now have three choices:

$$
m n_{2} \equiv\left\{\begin{array}{c}
n_{1} \bmod s  \tag{2.4.58}\\
n_{2} \bmod s \\
n_{3} \bmod s
\end{array}\right.
$$

If $m n_{2} \equiv n_{1} \bmod s$, then, combining this with the fact that $m n_{1} \equiv n_{2} \bmod s$, we see that $m^{2}=1$. However, this contradicts our initial assumption, and so is eliminated
from consideration. Similarly, $m n_{2} \equiv n_{2} \bmod s$ implies $m=1$ and again leads to a contradiction. Therefore, $m n_{2} \equiv n_{3} \bmod s$ must hold. Continuing in the same manner, we see:

$$
\begin{aligned}
& m n_{1} \equiv m n_{1} \\
& \equiv n_{2} \bmod s \\
& m n_{2} \equiv m^{2} n_{1} \equiv n_{3} \bmod s \\
& m n_{3} \equiv m^{3} n_{1} \equiv n_{4} \bmod s \\
& \vdots \\
& m n_{d-1} \equiv m^{d-1} n_{1} \equiv n_{d} \bmod s
\end{aligned}
$$

Therefore, we must have $m n_{d} \equiv m^{d} n_{1} \equiv n_{1} \bmod s$, which implies $m^{d}=1$. Since this contradicts our initial assumption, we see that no element $m \in \mathbb{Z}_{s}^{\times}$with $|m|=c>d$ can stabilize an element of $\mathbb{A}_{s}^{d}$ of the form $\left[n_{1}, \ldots, n_{d}\right], n_{j} \neq 0$ for all $j=1, \ldots, d$. The argument for elements of the form $\left[0, n_{1}, \ldots, n_{d-1}\right], n_{j} \neq 0$ for all $j=1, \ldots, d-1$, follows similarly.
II. Assume $|m|=c \leq d$.

We show an element of $\mathbb{A}_{s}^{d}$ is stabilized by $m$ if and only if $c \mid d$ or $c \mid d-1$. First, suppose $c \nmid d$ and $c \nmid d-1$. Therefore, there exists $q, r \in \mathbb{Z}$ such that

$$
\begin{equation*}
d=q c+r, \quad q \geq 0, \quad 1<r<c . \tag{2.4.59}
\end{equation*}
$$

Let $[n]=\left[n_{1}, \ldots, n_{d}\right] \in \mathbb{A}_{s}^{d}, n_{j} \neq 0$ for all $j=1, \ldots, d$, and suppose $m \cdot[n]=[n]$.

Following the same argument as in part I of this proof, we see:

$$
\begin{aligned}
& m n_{1} \equiv m n_{1} \equiv n_{2} \bmod s \\
& m n_{2} \equiv m^{2} n_{1} \equiv n_{3} \bmod s \\
& \vdots \\
& m n_{c-1} \equiv m^{c-1} n_{1} \equiv n_{c} \bmod s, \\
& m n_{c} \equiv m^{c} n_{1} \equiv n_{1} \bmod s
\end{aligned}
$$

where the last line results from the fact that $|m|=c \leq d$. Continuing, we see there are two possibilities for $m n_{c+1}$ :

$$
m n_{c+1} \equiv\left\{\begin{array}{l}
n_{j} \bmod s \text { for some } j \in\{1, \ldots, c\}  \tag{2.4.60}\\
n_{c+2} \bmod s
\end{array}\right.
$$

If $m n_{c+1} \equiv n_{j} \bmod s$ for some $j \in\{1, \ldots, c\}$, then $m n_{c+1} \equiv m n_{j-1} \bmod s$, where $n_{0}:=n_{c} \bmod s$. However, this would imply that $n_{c+1} \equiv n_{j-1} \bmod s$, a contradiction. Therefore, $m n_{c+1} \equiv m n_{c+2} \bmod s$ must hold, and we can continue with the previous line of reasoning to obtain:

$$
\begin{aligned}
& m n_{c+1} \equiv m n_{c+1} \equiv n_{c+2} \bmod s \\
& m n_{c+2} \equiv m^{2} n_{c+1} \equiv n_{c+3} \bmod s \\
& \vdots \\
& m n_{2 c-1} \equiv m^{c-1} n_{c+1} \equiv n_{2 c} \bmod s \\
& m n_{2 c} \equiv m^{c} n_{c+1} \equiv n_{c+1} \bmod s .
\end{aligned}
$$

Continuing with the pattern that has now been established, we arrive at:

$$
\begin{gathered}
m n_{q c+1} \equiv m n_{q c+1} \equiv n_{q c+2} \bmod s, \\
m n_{q c+2} \equiv m^{2} n_{q c+1} \equiv n_{q c+3} \bmod s, \\
\vdots \\
m n_{q c+r-1} \equiv m^{r-1} n_{q c+1} \equiv n_{q c+r} \bmod s .
\end{gathered}
$$

We must then have:

$$
\begin{equation*}
m n_{q c+r} \equiv m^{r} n_{q c+1} \equiv n_{q c+1} \bmod s, \tag{2.4.61}
\end{equation*}
$$

which in turn implies $m^{r}=1$, a contradiction. Therefore, no element of $\mathbb{A}_{s}^{d}$ of the form $\left[n_{1}, \ldots, n_{d}\right], n_{j} \neq 0$ for all $j=1, \ldots, d$, can be stabilized by an $m \in \mathbb{Z}_{s}^{\times}$with $|m|=c \leq d$ such that $c \nmid d$ and $c \nmid d-1$. The argument for elements of $\mathbb{A}_{s}^{d}$ of the form $\left[0, n_{1}, \ldots, n_{d-1}\right], n_{j} \neq 0$ for all $j=1, \ldots, d-1$, follows similarly.

We now shift our attention to $m \in \mathbb{Z}_{s}^{\times}$such that $c \mid d$ or $c \mid d-1$. In either case there exists a $q \in \mathbb{Z}$ such that,

$$
\begin{equation*}
d=q c, \quad q \geq 0, \quad \text { or } \quad d-1=q c, \quad q \geq 0 . \tag{2.4.62}
\end{equation*}
$$

Using the same argument that we just completed, we see that if $c \mid d$ then $m$ stabilizes certain elements of the form $\left[n_{1}, \ldots, n_{d}\right], n_{j} \neq 0$ for all $j=1, \ldots, d$, whereas if $c \mid d-1$ then $m$ stabilizes certain elements of the form $\left[0, n_{1}, \ldots, n_{d-1}\right]$, $n_{j} \neq 0$ for all $j=1, \ldots, d-1$. The only difference in reasoning comes at the end, where in this case we do not run into a contradiction. Furthermore, looking back at the above reasoning, we see all elements $\left[n_{1}, \ldots, n_{d}\right] \in \mathbb{A}_{s}^{d}$ stabilized by $m$ must satisfy:

$$
\begin{equation*}
m n_{j c+k} \equiv m^{k} n_{j c+1} \equiv n_{j c+k+1}, \quad \forall j=0, \ldots, q-1, \quad k=1, \ldots, c-1 \tag{2.4.63}
\end{equation*}
$$

where $d=q c$ or $d-1=q c$, depending on the type of element of $\mathbb{A}_{s}^{d}$. By equation (2.4.63), any element in $\mathbb{A}_{s}^{d}$ stabilized by $m$ can be written in one of two general forms:

$$
[n]=\left\{\begin{array}{l}
{\left[n_{1}, m n_{1}, \ldots, m^{c-1} n_{1}, \ldots, n_{\frac{d}{c}}, m n_{\frac{d}{c}}, \ldots, m^{c-1} n_{\frac{d}{c}}\right]}  \tag{2.4.64}\\
{\left[0, n_{1}, m n_{1}, \ldots, m^{c-1} n_{1}, \ldots, n_{\frac{d-1}{c}}, m n_{\frac{d-1}{c}}, \ldots, m^{c-1} n_{\frac{d-1}{c}}\right]}
\end{array}\right.
$$

where $n_{j} \neq 0$ and $n_{j} \neq n_{k}$ for all $j, k=1, \ldots, d / c$ or $j, k=1, \ldots,(d-1) / c$, depending on the form of $[n]$. Also, since $|m|=c$, there must exist a primitive root $\bmod s, g$, such that

$$
\begin{equation*}
m=g^{(s-1) / c} \tag{2.4.65}
\end{equation*}
$$

noting that $c \mid s-1$ since the order of any group element must divide the order of the group. Combining equations (2.4.64) and (2.4.65) gives (2.4.56 $c$ ).

In order to prove (2.4.54), we exploit the fact that

$$
\begin{equation*}
\left|\mathcal{O}_{[n]}\right|=\frac{s-1}{\left|\left(\mathbb{Z}_{s}^{\times}\right)_{[n]}\right|} \tag{2.4.66}
\end{equation*}
$$

By (2.4.66), we need only compute the stabilizer of $[n]$ in $\mathbb{Z}_{s}^{\times}$, that is $\left(\mathbb{Z}_{s}^{\times}\right)_{[n]}$. But (2.4.64) and (2.4.65) easily give

$$
\begin{equation*}
\left(\mathbb{Z}_{s}^{\times}\right)_{[n]}=\left\{g^{l(s-1) / c}: l=0, \ldots, c-1\right\} . \tag{2.4.67}
\end{equation*}
$$

Clearly $\left|\left(\mathbb{Z}_{s}^{\times}\right)_{[n]}\right|=c$, thus proving (2.4.54).

Before counting the number of orbits $\mathbb{A}_{s}^{d}$, we prove two lemmas that simplify this task. The first shows that the choice of $g$ in (2.4.55) does not matter.

Lemma 2.4.4. If $g_{1}$ and $g_{2}$ are two primitive roots $\bmod s$, and $n_{1} \in \mathbb{Z}_{s}, n_{1} \neq 0$, then

$$
\begin{equation*}
\left[n_{1}, g_{1}^{(s-1) / c} n_{1}, \ldots, g_{1}^{(c-1)(s-1) / c} n_{1}\right]=\left[n_{1}, g_{2}^{(s-1) / c} n_{1}, \ldots, g_{2}^{(c-1)(s-1) / c} n_{1}\right] \tag{2.4.68}
\end{equation*}
$$

Proof. Since $g_{1}$ and $g_{2}$ are both primitive roots $\bmod s$, the sets $\left\{1, g_{1}^{(s-1) / c}, \ldots, g_{1}^{(c-1)(s-1) / c}\right\}$ and $\left\{1, g_{2}^{(s-1) / c}, \ldots, g_{2}^{(c-1)(s-1) / c}\right\}$ are both complete solution sets to $x^{c} \equiv 1 \bmod s$. Therefore $\left(n_{1}, g_{2}^{(s-1) / c} n_{1}, \ldots, g_{2}^{(c-1)(s-1) / c} n_{1}\right)$ is a rearrangement of $\left(n_{1}, g_{1}^{(s-1) / c} n_{1}, \ldots, g_{1}^{(c-1)(s-1) / c} n_{1}\right)$, and the lemma follows.

The second lemma shows that the representation given by (2.4.56 c) is not unique and gives the instances where confusion can occur.

Lemma 2.4.5. Let $[n] \in \mathbb{A}_{s}^{d}$ such that $[n]$ can be written in the form (2.4.35 b). If $c \mid b$, then $[n]$ can be written in the form (2.4.56c) as well.

Proof. We assume $[n]=\left[\tilde{n}_{1}^{b}, \tilde{n}_{2}^{b}, \ldots, \tilde{n}_{d / b}^{b}\right]$ and show that we can rewrite this as $[n]=\left[n_{1}^{c}, n_{2}^{c}, \ldots, n_{d / c}^{c}\right]$. If $[n]=\left[0, \tilde{n}_{1}^{b}, \tilde{n}_{2}^{b}, \ldots, \tilde{n}_{(d-1) / b}^{b}\right]$ then a similar proof shows how to rewrite this as $[n]=\left[0, n_{1}^{c}, n_{2}^{c}, \ldots, n_{(d-1) / c}^{c}\right]$. Recall

$$
\tilde{n}_{k}^{b}=\left[\tilde{n}_{k}, g^{(s-1) / b} \tilde{n}_{k}, \ldots, g^{(b-1)(s-1) / b} \tilde{n}_{k}\right] .
$$

Let $a=b / c$ and set $n_{1}=\tilde{n}_{1}$; we want to construct $n_{1}^{c}$ out of elements of $\tilde{n}_{1}^{b}$, where:

$$
\tilde{n}_{1}^{b}=\left[\tilde{n}_{1}, g^{(s-1) / b} \tilde{n}_{1}, \ldots, g^{(b-1)(s-1) / b} \tilde{n}_{1}^{b}\right] .
$$

Since the order of elements does not matter, we may pick them however we like and rearrange them as we wish. We have that $n_{1}^{c}$ is formed out of the following elements
of $\tilde{n}_{1}^{b}$ :

$$
\begin{aligned}
n_{1}^{c} & =\left[\tilde{n}_{1}, g^{a(s-1) / b} \tilde{n}_{1}, \ldots, g^{(c-1) a(s-1) / b} \tilde{n}_{1}\right] \\
& =\left[n_{1}, g^{a(s-1) / b} n_{1}, \ldots, g^{(c-1) a(s-1) / b} n_{1}\right] \\
& =\left[n_{1}, g^{a(s-1) / c a} n_{1}, \ldots, g^{(c-1) a(s-1) / c a} n_{1}\right] \\
& =\left[n_{1}, g^{(s-1) / c} n_{1}, \ldots, g^{(c-1)(s-1) / c} n_{1}\right] .
\end{aligned}
$$

Likewise, set $n_{k}=\tilde{n}_{k}$ for $k=2, \ldots, d / b$, and construct $n_{k}^{c}$ in a similar manner. For the next $c$-tuple, set $n_{\frac{d}{b}+1}=g^{(s-1) / b} \tilde{n}_{1}$. We then have:

$$
\begin{aligned}
n_{\frac{d}{b}+1}^{c} & =\left[g^{(s-1) / b} \tilde{n}_{1}, g^{(a+1)(s-1) / b} \tilde{n}_{1}, \ldots, g^{((c-1) a+1)(s-1) / b} \tilde{n}_{1}\right] \\
& =\left[n_{\frac{d}{b}+1}, g^{a(s-1) / b} n_{\frac{d}{b}+1}, \ldots, g^{(c-1) a(s-1) / b} n_{\frac{d}{b}+1}\right] \\
& =\left[n_{\frac{d}{b}+1}, g^{(s-1) / c} n_{\frac{d}{b}+1}, \ldots, g^{(c-1)(s-1) / c} n_{\frac{d}{b}+1}\right] .
\end{aligned}
$$

In general,

$$
\begin{equation*}
n_{\frac{j d}{b}+k}=g^{j(s-1) / b} \tilde{n}_{k}, \quad \forall j=0, \ldots, a-1, \quad k=1, \ldots, d / b \tag{2.4.69}
\end{equation*}
$$

and the resulting $n_{\frac{j d}{b}+k}^{c}$ follows similarly.

### 2.4.3 Proof of Theorem 2.2.1

Using theorem 2.4.3 as well as lemmas 2.4.4 and 2.4.5, we now count the number of orbits of $\mathbb{A}_{s}^{d}$. By proposition 2.3.2 this is the same as counting the number of inequivalent harmonic frames, and so will complete the proof of theorem 2.2.1. Let $\gamma_{c}$ denote the total number of orbits of $\mathbb{A}_{s}^{d}$ with $(s-1) / c$ elements. Then,
by theorem 2.4.3, the total number of orbits of $\mathbb{A}_{s}^{d}$ is given by

$$
\begin{equation*}
\gamma_{1}+\sum_{\substack{c \mid d \\ c>1}} \gamma_{c}+\sum_{\substack{c \mid d-1 \\ c>1}} \gamma_{c} \tag{2.4.70}
\end{equation*}
$$

Notice the similarity between equations (2.4.70) and (2.2.28). In fact, we shall prove that

$$
\begin{equation*}
\gamma_{c}=\alpha_{c}, \quad \forall c \in \mathbb{N} \text { such that } c \mid s-1 \text { and } c \mid d \text { or } c \mid d-1 \tag{2.4.71}
\end{equation*}
$$

Theorem 2.4.6. Let $s$ be a prime number, $1<d<s, c \mid s-1, c>1$, and let $\beta_{c}$ denote the cumulative order of all orbits of size $(s-1) / c$. Furthermore, let $\gamma_{c}$ denote the number of orbits of $\mathbb{A}_{s}^{d}$ of size $(s-1) / c$, so that

$$
\gamma_{c}=\frac{c \beta_{c}}{s-1}
$$

If $c \mid d$, then $\beta_{c}$ is given by the following backwards recursive formula:

$$
\beta_{c}=\frac{(s-1)(s-1-c) \cdots\left(s-1-\left(\frac{d}{c}-1\right) c\right)}{c^{\frac{d}{c}}(d / c)!}-\sum_{\substack{c<b<s \\ c|b, b| d}}\left(\frac{s-1}{b}\right) \gamma_{b}
$$

If $c \mid d-1$, then $\beta_{c}$ is given by the following backwards recursive formula:

$$
\beta_{c}=\frac{(s-1)(s-1-c) \cdots\left(s-1-\left(\frac{d-1}{c}-1\right) c\right)}{c^{\frac{d-1}{c}}((d-1) / c)!}-\sum_{\substack{c<b<s \\ c|b, b| d-1}}\left(\frac{s-1}{b}\right) \gamma_{b} .
$$

The number of orbits of $\mathbb{A}_{s}^{d}$ of size $s-1$, denoted $\gamma_{1}$, is given by:

$$
\gamma_{1}=\frac{1}{s-1}\binom{s}{d}-\sum_{\substack{c \mid d \\ c>1}} \frac{\gamma_{c}}{c}-\sum_{\substack{c \mid d-1 \\ c>1}} \frac{\gamma_{c}}{c} .
$$

Proof. We prove the formula for $\beta_{c}$ when $c \mid d$, noting that the proof is identical for the case when $c \mid d-1$. In order to accomplish this task, we will build up the formula
using combinatorial arguments. By theorem 2.4.3, the elements we are counting are of the form $\left[n_{1}^{c}, n_{2}^{c}, \ldots, n_{d / c}^{c}\right]$, where

$$
n_{k}^{c}=\left[n_{k}, g^{(s-1) / c} n_{k}, \ldots, g^{(c-1)(s-1) / c} n_{k}\right], \quad n_{k} \neq 0
$$

It is clear then, that we have $s-1$ choices for $n_{1}, s-1-c$ choices for $n_{2}, s-1-2 c$ choices for $n_{3}$, and so on. Continuing to the end, we see there are $s-1-(d / c-1) c$ choices for $n_{d / c}$. Furthermore, by lemma 2.4.4, the choice of $g$ does not matter, and so does not add any new elements to count. Therefore, at the moment, we have

$$
\begin{equation*}
(s-1)(s-1-c) \cdots(s-1-(d / c-1) c) \tag{2.4.72}
\end{equation*}
$$

elements. Fixing the choice of $n_{1}$ temporarily, it is clear that if we chose any of $g^{(s-1) / c} n_{1}, \ldots, g^{(c-1)(s-1) / c} n_{1}$ instead of $n_{1}$, then we would have a rearranged version of $n_{1}^{c}$. However, the order of elements does not matter in $\mathbb{A}_{s}^{d}$, and so these choices are in fact the same as choosing $n_{1}$. Since there are $c$ such elements (including $n_{1}$ ), the number of distinct choices for $n_{1}$ is in fact $(s-1) / c$. Similarly, we must divide the number of choices for each $n_{k}$ by a factor of $c$, thus giving

$$
\begin{equation*}
\frac{(s-1)(s-1-c) \cdots(s-1-(d / c-1) c)}{c^{d / c}} \tag{2.4.73}
\end{equation*}
$$

elements. Furthermore, again recalling that the order of elements does not matter, we see that the order in which we choose $n_{1}, \ldots, n_{d / c}$ does not matter either. Consequently, we are now down to

$$
\begin{equation*}
\frac{(s-1)(s-1-c) \cdots(s-1-(d / c-1) c)}{c^{d / c}(d / c)!} \tag{2.4.74}
\end{equation*}
$$

elements. We note that equation (2.4.74) gives the number of elements of the form $(2.4 .56 c)$. However, we are not counting all elements of the form (2.4.56 $c$ ), but only
those that are in an orbit of size $(s-1) / c$. In fact, by lemma 2.4.5 any element in an orbit of size $(s-1) / b$, where $c \mid b$ and $b \mid d$, can be rewritten as $\left[n_{1}^{c}, n_{2}^{c}, \ldots, n_{d / c}^{c}\right]$. Therefore, we must subtract all elements in orbits of size $(s-1) / b$, where $c \mid b$ and $b \mid d$, thus giving:

$$
\begin{equation*}
\beta_{c}=\frac{(s-1)(s-1-c) \cdots\left(s-1-\left(\frac{d}{c}-1\right) c\right)}{c^{\frac{d}{c}}(d / c)!}-\sum_{\substack{c<b<s \\ c|b, b| d}}\left(\frac{s-1}{b}\right) \gamma_{b} . \tag{2.4.75}
\end{equation*}
$$

The equation for $\gamma_{1}$ follows from

$$
\begin{equation*}
\left|\mathbb{A}_{s}^{d}\right|=\gamma_{1}(s-1)+\sum_{\substack{c \mid d \\ c>1}}\left(\frac{s-1}{c}\right) \gamma_{c}+\sum_{\substack{c \mid d-1 \\ c>1}}\left(\frac{s-1}{c}\right) \gamma_{c}, \tag{2.4.76}
\end{equation*}
$$

and the fact that $\left|\mathbb{A}_{s}^{d}\right|=\binom{s}{d}$.

Example 2.4.7. We apply theorem 2.4.6 for the case when $d=3$ and $s \equiv 1 \bmod 3$. In this case, $c=3$ divides $d$ as well as $s-1$, while $c=2$ divides $d-1$ as well as $s-1$. Therefore we compute:

$$
\beta_{3}=\frac{s-1}{3} \quad \text { and } \quad \beta_{2}=\frac{s-1}{2}
$$

which in turn gives:

$$
\gamma_{3}=1 \quad \text { and } \quad \gamma_{2}=1
$$

Thus,

$$
\begin{aligned}
\gamma_{1} & =\frac{1}{s-1}\binom{s}{3}-\frac{1}{3}-\frac{1}{2} \\
& =\frac{s^{2}-2 s}{6}-\frac{2}{6}-\frac{3}{6} \\
& =\frac{s^{2}-2 s-5}{6}
\end{aligned}
$$

and so the total number of orbits is:

$$
\begin{aligned}
\gamma_{1}+\gamma_{2}+\gamma_{3} & =\frac{s^{2}-2 s-5}{6}+1+1 \\
& =\frac{s^{2}-2 s+7}{6}
\end{aligned}
$$

Notice this is the same result as proposition 2.4.2.

### 2.5 The Symmetry Group

We now turn our attention to the symmetry group of prime order harmonic frames. The following theorem proves the existence of a particular subgroup of $\operatorname{Sym}\left(\Phi_{n}\right)$ that is dependent on the generators $n_{1}, \ldots, n_{d}$ as well as the order of $\mathcal{O}_{[n]}$.

Theorem 2.5.1. Let $\mathcal{O}_{[n]}$ be an orbit of $\mathbb{A}_{s}^{d}$ such that $\left|\mathcal{O}_{[n]}\right|=(s-1) / c$, and let $\Phi_{n}$ be the harmonic frame that corresponds to $\mathcal{O}_{[n]}$ under the one-to-one correspondence described by proposition 2.3.2. Then

$$
\left\langle\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right), Q\right\rangle \subseteq \operatorname{Sym}\left(\Phi_{n}\right)
$$

where $\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right)$ denotes a $d \times d$ matrix with $\omega^{n_{1}}, \ldots, \omega^{n_{d}}$ on the diagonal and zeros elsewhere, $\omega=e^{2 \pi i / s}, Q$ is a $d \times d$ permutation matrix dependent on $\Phi_{n}$, and $|\langle Q\rangle|=c$.

Proof. Similar to the proof of theorem 2.3.1, we shall almost exclusively consider $\Phi$ as a $d \times s$ matrix. We recall that $U \in \operatorname{Sym}\left(\Phi_{n}\right)$ if and only if there exists an $s \times s$ permutation matrix $P$ such that

$$
\begin{equation*}
U \Phi=\Phi P \tag{2.5.77}
\end{equation*}
$$

First using the left hand side of (2.5.77), we have

$$
\begin{equation*}
(U \Phi)^{\star}(U \Phi)=\Phi^{\star} U^{\star} U \Phi=\Phi^{\star} \Phi \tag{2.5.78}
\end{equation*}
$$

and then equivalently for the right hand side of (2.5.77),

$$
\begin{equation*}
(\Phi P)^{\star}(\Phi P)=P^{\star} \Phi^{\star} \Phi P \tag{2.5.79}
\end{equation*}
$$

Combining (2.5.78) and (2.5.79) we obtain the following necessary condition for (2.5.77),

$$
\Phi^{\star} \Phi=P^{\star} \Phi^{\star} \Phi P
$$

or equivalently,

$$
\begin{equation*}
P \Phi^{\star} \Phi P^{\star}=\Phi^{\star} \Phi . \tag{2.5.80}
\end{equation*}
$$

The matrix $\Phi^{\star} \Phi$ is called the Gram matrix and has the following form:

$$
\begin{equation*}
\left(\Phi^{\star} \Phi\right)_{j, k}=\left\langle\varphi_{k}, \varphi_{j}\right\rangle=\sum_{l=1}^{d} e^{2 \pi i n_{l}(k-j) / s}, \quad \forall j, k=0, \ldots, s-1 . \tag{2.5.81}
\end{equation*}
$$

Two elements $\left\langle\varphi_{k}, \varphi_{j}\right\rangle$ and $\left\langle\varphi_{k^{\prime}}, \varphi_{j^{\prime}}\right\rangle$ of $\Phi^{\star} \Phi$ are equal if and only if

$$
\begin{equation*}
\sum_{l=1}^{d} e^{2 \pi i n_{l}(k-j) / s}=\sum_{l=1}^{d} e^{2 \pi i n_{l}\left(k^{\prime}-j^{\prime}\right) / s} . \tag{2.5.82}
\end{equation*}
$$

Using the same minimum polynomial argument as the one found in the proof of theorem 2.3.1, we see that (2.5.82) holds for off diagonal elements of $\Phi^{\star} \Phi$ if and only if there exists a permutation $\mu \in S_{d}$ such that

$$
\begin{equation*}
n_{l}(k-j) \equiv n_{\mu(l)}\left(k^{\prime}-j^{\prime}\right) \bmod s, \quad \forall l=1, \ldots, d, \quad k \neq j, \quad k^{\prime} \neq j^{\prime} \tag{2.5.83}
\end{equation*}
$$

(2.5.83) is in fact the same condition as (2.3.39), and so we may define the following equivalence relation between the off diagonal entries of $\Phi^{\star} \Phi$ and the elements of $\mathbb{A}_{s}^{d}$ :

$$
\begin{equation*}
\left\langle\varphi_{k}, \varphi_{j}\right\rangle \sim(k-j \bmod s) \cdot[n], \quad k \neq j \tag{2.5.84}
\end{equation*}
$$

For the diagonal entries of $\Phi^{\star} \Phi$, we define the representative [0] as

$$
\begin{equation*}
[0]:=[\underbrace{0, \ldots, 0}_{d}], \tag{2.5.85}
\end{equation*}
$$

and extend our equivalence relation to diagonal elements:

$$
\begin{equation*}
\left\langle\varphi_{j}, \varphi_{j}\right\rangle \sim[0] \tag{2.5.86}
\end{equation*}
$$

In order to ease notation, we set $0 \cdot[n]:=[0]$, and thus can write $k \cdot[n]$ for all $k \in \mathbb{Z}_{s}$. Combining (2.5.84) and (2.5.86), we see $\sim$ induces an equivalence relation between the set of inner products, $\left\{\left\langle\varphi_{j}, \varphi_{k}\right\rangle: j, k=0, \ldots, s-1\right\}$, and the set $\mathbb{A}_{s}^{d} \cup\{[0]\}$. Defining the matrix $G$ as

$$
\begin{equation*}
G_{j, k}:=(k-j) \cdot[n], \quad \forall j, k \in \mathbb{Z}_{s}, \tag{2.5.87}
\end{equation*}
$$

we then have an equivalence relation between $\Phi^{\star} \Phi$ and $G$ :

$$
\begin{equation*}
\Phi^{\star} \Phi \sim G . \tag{2.5.88}
\end{equation*}
$$

Combining (2.5.80) with (2.5.88) gives the following necessary condition for (2.5.77) to hold:

$$
\begin{equation*}
P G P^{\star}=G \tag{2.5.89}
\end{equation*}
$$

Returning to (2.5.87), we see $G$ has the form:

$$
G=\left(\begin{array}{cccccc}
a_{0} & a_{s-1} & a_{s-2} & \cdots & a_{2} & a_{1}  \tag{2.5.90}\\
a_{1} & a_{0} & a_{s-1} & a_{s-2} & \cdots & a_{2} \\
a_{2} & a_{1} & a_{0} & \ddots & \ddots & \vdots \\
\vdots & \ddots & \ddots & \ddots & a_{s-1} & a_{s-2} \\
a_{s-2} & \cdots & a_{2} & a_{1} & a_{0} & a_{s-1} \\
a_{s-1} & a_{s-2} & \cdots & a_{2} & a_{1} & a_{0}
\end{array}\right)
$$

where $a_{k}=k \cdot[n]$ for all $k \in \mathbb{Z}_{s}$. Therefore $G$ is a circulant matrix, and is completely determined by its first column vector. The permutation matrix

$$
T:=\left(\begin{array}{cccccc}
0 & 1 & 0 & 0 & \cdots & 0  \tag{2.5.91}\\
0 & 0 & 1 & 0 & \cdots & 0 \\
0 & 0 & 0 & 1 & \cdots & 0 \\
\vdots & \vdots & \vdots & \vdots & \ddots & \vdots \\
0 & 0 & 0 & 0 & \cdots & 1 \\
1 & 0 & 0 & 0 & \cdots & 0
\end{array}\right)
$$

is called the basic circulant permutation matrix. A matrix $A$ can be written in the form

$$
\begin{equation*}
A=\sum_{k=0}^{s-1} a_{k} T^{k} \tag{2.5.92}
\end{equation*}
$$

if and only if $A$ is circulant. Therefore, $G$ can be written in the form (2.5.92), and as such, it is clear that

$$
\begin{equation*}
T^{k} G\left(T^{k}\right)^{\star}=G, \quad \forall k=0, \ldots, s-1 \tag{2.5.93}
\end{equation*}
$$

A simple computation shows that when $U=\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right)$, one has

$$
\begin{equation*}
U^{k} \Phi=\Phi T^{k}, \quad \forall k=0, \ldots, s-1 . \tag{2.5.94}
\end{equation*}
$$

Thus, regardless of the size $\mathcal{O}_{[n]}$,

$$
\begin{equation*}
\operatorname{diag}\left(\omega^{k n_{1}}, \ldots, \omega^{k n_{d}}\right) \in \operatorname{Sym}\left(\Phi_{n}\right), \quad \forall k=0, \ldots, s-1 \tag{2.5.95}
\end{equation*}
$$

Note this proves the theorem for the case $\left|\mathcal{O}_{[n]}\right|=s-1$.
To prove the existence of the matrix $Q \in \operatorname{Sym}\left(\Phi_{n}\right)$ with $|\langle Q\rangle|=c$, suppose that $\Phi_{n}$ corresponds to $\mathcal{O}_{[n]}$ such that $\left|\mathcal{O}_{[n]}\right|=(s-1) / c$, where $c>1$. Note that by theorem 2.4.3 we have

$$
\begin{equation*}
g^{k(s-1) / c} m \cdot[n]=m \cdot[n], \quad \forall m \in \mathbb{Z}_{s}^{\times}, \quad k=1, \ldots, c, \tag{2.5.96}
\end{equation*}
$$

and in particular

$$
\begin{equation*}
g^{k(s-1) / c} \cdot[n]=[n], \quad \forall k=1, \ldots, c \tag{2.5.97}
\end{equation*}
$$

Therefore, the action of $g^{(s-1) / c}$ on $n$ defines a permutation $\rho \in S_{d}$ such that

$$
\begin{equation*}
\left(n_{\rho^{k}(1)}, \ldots, n_{\rho^{k}(d)}\right)=g^{k(s-1) / c} \cdot\left(n_{1}, \ldots, n_{d}\right), \quad \forall k=1, \ldots, c . \tag{2.5.98}
\end{equation*}
$$

Since a permutation of the generators $n_{1}, \ldots, n_{d}$ is equivalent to a permutation of the rows of $\Phi,(2.5 .98)$ implies the existence of a $d \times d$ permutation matrix $Q$, where $Q$ is the matrix equivalent of $\rho$, as well as an $s \times s$ permutation matrix $P_{0}$, such that

$$
\begin{equation*}
Q^{k} \Phi=\Phi P_{0}^{k}, \quad \forall k=1, \ldots, c \tag{2.5.99}
\end{equation*}
$$

In other words, $Q \in \operatorname{Sym}\left(\Phi_{n}\right)$, and since $\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right) \in \operatorname{Sym}\left(\Phi_{n}\right)$ as well, we must have

$$
\begin{equation*}
\left\langle\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right), Q\right\rangle \subseteq \operatorname{Sym}\left(\Phi_{n}\right) \tag{2.5.100}
\end{equation*}
$$

Corollary 2.5.2. Let $\mathcal{O}_{[n]}$ be an orbit of $\mathbb{A}_{s}^{d}$ such that $\left|\mathcal{O}_{[n]}\right|=s-1$, and let $\Phi_{n}$ be the harmonic frame that corresponds to $\mathcal{O}_{[n]}$ under the one-to-one correspondence described by proposition 2.3.2. Then

$$
\operatorname{Sym}\left(\Phi_{n}\right)=\left\langle\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right)\right\rangle
$$

where $\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right)$ denotes a $d \times d$ matrix with $\omega^{n_{1}}, \ldots, \omega^{n_{d}}$ on the diagonal and zeros elsewhere, and $\omega=e^{2 \pi i / s}$.

Proof. Recall the matrices $G$ and $T$ from the proof of theorem 2.5.1, as given by equations (2.5.87) and (2.5.91), respectively. We will show that $P=T^{k}, k=$
$0, \ldots, s-1$, are the only matrices satisfying the necessary condition given by equation (2.5.89). Combining the fact that $\mathcal{O}_{[n]}=\left\{m \cdot[n]: m \in \mathbb{Z}_{s}^{\times}\right\}$with the assumption that $\left|\mathcal{O}_{[n]}\right|=s-1$, we have

$$
\begin{equation*}
k \cdot[n]=k^{\prime} \cdot[n] \Longleftrightarrow k \equiv k^{\prime} \bmod s \tag{2.5.101}
\end{equation*}
$$

Furthermore, let $\sigma \in S_{s}$ be the permutation corresponding to the permutation matrix $P$. Equation (2.5.89) can be rewritten as

$$
\begin{equation*}
(\sigma(j)-\sigma(k)) \cdot[n]=(j-k) \cdot[n], \quad \forall j, k \in \mathbb{Z}_{s} . \tag{2.5.102}
\end{equation*}
$$

Combining equations (2.5.101) and (2.5.102), one obtains

$$
\begin{equation*}
\sigma(j)-\sigma(k)=j-k, \quad \forall j, k \in \mathbb{Z}_{s} . \tag{2.5.103}
\end{equation*}
$$

One can think of (2.5.103) as a system of $s^{2}$ linear equations in the $s$ variables $\sigma(0), \ldots, \sigma(s-1)$, with the two added constraints:

1. $\sigma(k) \in \mathbb{Z}_{s}$ for all $k \in \mathbb{Z}_{s}$,
2. $\sigma(j)=\sigma(k)$ if and only if $j=k$.

Clearly (2.5.103) is an overdetermined system. However, (2.5.103) has $s-1$ independent equations, given by:

$$
\begin{aligned}
\sigma(1)-\sigma(0) & \equiv 1 \bmod s \\
\sigma(2)-\sigma(0) & \equiv 2 \bmod s \\
& \vdots \\
\sigma(s-1)-\sigma(0) & \equiv s-1 \bmod s .
\end{aligned}
$$

Thus $\sigma(0)$ is a free variable, and can be assigned any value from $\mathbb{Z}_{s}$. The remaining values of $\sigma$ are then given by:

$$
\begin{equation*}
\sigma(j) \equiv j+\sigma(0) \bmod s, \quad \forall j=1, \ldots, s-1 \tag{2.5.104}
\end{equation*}
$$

In conclusion, there are $s$ possible permutations, each corresponding to a different value of $\sigma(0)$. In particular, we have the following correspondence:

$$
\begin{equation*}
\sigma(0)=k \Longleftrightarrow P=T^{k} . \tag{2.5.105}
\end{equation*}
$$

The following conjecture asserts that the subgroup described in theorem 2.5.1 in fact is the symmetry group for all prime order harmonic frames, not just those corresponding to orbits of size $s-1$.

Conjecture 2.5.3. Let $\mathcal{O}_{[n]}$ be an orbit of $\mathbb{A}_{s}^{d}$ such that $\left|\mathcal{O}_{[n]}\right|=(s-1) / c$, and let $\Phi_{n}$ be the harmonic frame that corresponds to $\mathcal{O}_{[n]}$ under the one-to-one correspondence described by proposition 2.3.2. Then

$$
\operatorname{Sym}\left(\Phi_{n}\right)=\left\langle\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right), Q\right\rangle,
$$

where $\operatorname{diag}\left(\omega^{n_{1}}, \ldots, \omega^{n_{d}}\right)$ denotes a $d \times d$ matrix with $\omega^{n_{1}}, \ldots, \omega^{n_{d}}$ on the diagonal and zeros elsewhere, $\omega=e^{2 \pi i / s}, Q$ is a $d \times d$ permutation matrix dependent on $\Phi_{n}$, and $|\langle Q\rangle|=c$.

### 2.6 Closing remarks

We have enumerated all harmonic frames for $\mathbb{C}^{d}$ with $s$ elements, where $s$ is a prime number. A natural question is how to extend these results to all $s$. Certain
problems arise, however, with the techniques used in this chapter, since in several instances the fact that $s$ is prime is a key element. In particular, for a general $s$, distinct harmonic frames will arise from groups other than $\mathbb{Z}_{s}$. Also, even for those harmonic frames that do come from $\mathbb{Z}_{s}$, new representations must be developed since in general $\mathbb{Z}_{s}^{\times} \subseteq\{1, \ldots, s\}$.

## Chapter 3

## Frame Based Kernel Methods

### 3.1 Introduction to Multispectral and Hyperspectral Imagery Data

When a camera takes a picture, reflected light from the subject is passed through three filters: red, green, and blue. The resulting bands are then combined to form a color image; see figure 3.1. Multispectral and hyperspectral cameras,

Figure 3.1: Color image decomposition

(a) Color image

(b) Red band

(c) Green band

(d) Blue band
on the other hand, are in a sense a generalization of a regular camera. Rather
than filter the reflected light through red, green, and blue filters, these cameras are able to measure reflectance at a multitude of different wavelengths; see figure 3.2. Observing the sample bands in figure 3.2, one notices that the reflectance is measured at wavelengths far beyond the visible spectrum (recall the visible spectrum is about $380-750 \mathrm{~nm}$ ). Unlike standard cameras, the purpose of multispectral and hyperspectral cameras is not to create a color image, but rather to collect as much information about a particular scene as possible. This additional information can then be used for many different tasks, including (see [30] for more details):

- target detection
- material mapping
- material identification
- mapping details of surface properties

What figure 3.2 does not illustrate, though, is the central difference between multispectral imagery (MSI) data sets and hyperspectral imagery (HSI) data sets. While the bands are spread across the spectrum, there are large gaps where no measurements are shown. HSI data sets are in fact characterized by the narrowness and contiguous nature of their measurements, leaving few if any large spectral gaps. One can imagine stacking each of the bands and forming a cube, as illustrated by figure 3.3. Given this abundance of spectral information, HSI data sets are generally spectrally overdetermined, and are thus able to distinguish between spectrally similar materials. In order to achieve the contiguous nature of the measurements,

Figure 3.2: Selected bands of a hyperspectral data set

(a) 412 nm

(d) 650 nm

(g) 1237 nm

(j) 1992 nm

(b) 468 nm

(e) 808 nm

(h) 1451 nm

(k) 2144 nm

(c) 543 nm

(f) 1014 nm

(i) 1648 nm

(l) 2284 nm

Figure 3.3: Hyperspectral data cube

usually there are hundreds of spectral bands. MSI data sets on the other hand have anywhere from 4 spectral bands up to one hundred. In practical applications, the number of pixels is on the order of hundreds of thousands, sometimes even millions.

Mathematically speaking, we can model a MSI/HSI data set in the following way. Let $X$ denote a MSI/HSI data cube with dimensions $N_{1} \times N_{2} \times D$, where $N_{1}$ and $N_{2}$ are the spatial dimensions (length and width), and $D$ is the spectral dimension. Thus we have $N=N_{1} N_{2}$ pixels, each measured at $D$ different wavelengths. Most of the time it will be easier to think of $X$ as a list, and so we set $X=\left\{x_{i}: i=\right.$ $1, \ldots, N\} \subset \mathbb{R}^{D}$, where each $x_{i}$ corresponds to a pixel.

We will use MSI/HSI data for the purpose of material classification: given a list of potential classes within a data set, we aim to correctly classify each pixel as a certain class. Since HSI data sets are spectrally overdetermined, there are usually
less classes than the number of wavelengths measured. The MSI data sets that we examine will also have this property, although in general this is not true. A traditional method for classification in HSI data is through the use of endmember extraction algorithms. Endmembers are defined as a collection of the scene's constituent spectra. If $E=\left\{e_{i}: i=1, \ldots, s\right\}$ are endmembers for the HSI data set $X$, then the linear mixture model is

$$
\begin{equation*}
x_{i}=\sum_{j=1}^{s} \alpha_{i, j} e_{j}+N_{x_{i}}, \quad \forall x_{i} \in X \tag{3.1.1}
\end{equation*}
$$

where $N_{x_{i}}$ is a noise vector. The set $\left\{\alpha_{i, j}: i=1, \ldots, N, j=1, \ldots, s\right\}$ are the coefficients, and it is usually assumed that they satisfy the following two conditions:

$$
\begin{gather*}
\alpha_{i, j} \geq 0, \quad \forall i=1, \ldots, N, \quad j=1, \ldots, s  \tag{3.1.2}\\
\sum_{j=1}^{s} \alpha_{i, j}=1, \quad \forall i=1, \ldots, N \tag{3.1.3}
\end{gather*}
$$

Let $\alpha_{i,}=\left(\alpha_{i, 1}, \ldots, \alpha_{i, s}\right)$ and let $\tilde{\alpha}=\left(\tilde{\alpha}_{1}, \ldots, \tilde{\alpha}_{s}\right) \in \mathbb{R}^{s}$. Two common endmember coefficient sets are given by:

$$
\begin{gather*}
\alpha_{i, .}=\arg \min _{\tilde{\alpha}}\left\|x_{i}-\sum_{j=1}^{s} \tilde{\alpha}_{j} e_{j}\right\|_{\ell^{2}} \quad \text { subject to } \quad \text { (3.1.2), (3.1.3), }  \tag{3.1.4}\\
\alpha_{i,,}=\arg \min _{\tilde{\alpha}}\left\|x_{i}-\sum_{j=1}^{s} \tilde{\alpha}_{j} e_{j}\right\|_{\ell^{2}}+\tau_{i}\|\tilde{\alpha}\|_{\ell^{1}} \quad \text { subject to } \quad \text { (3.1.2), (3.1.3), } \tag{3.1.5}
\end{gather*}
$$

where $\tau_{i}$ is a positive real number. Most endmember extraction algorithms determine $E$ as a subset of $X$, i.e. it is assumed that the endmembers lie within the given data set. There are several endmember extraction algorithms, including N-FINDR [35], ORASIS [11], Pixel Purity Index [10], and Support Vector Data Description (SVDD) $[6,31]$; see also $[15,21]$.

### 3.2 Overview of New Algorithm

The main objective of this part of the thesis is to introduce a new algorithm for the purposes of material classification in MSI and HSI data sets. This algorithm is based on the theory of frames and dimension reduction, in particular kernel eigenmap methods.

As stated above, traditional endmember algorithms determine a subset $E \subset$ $X$ by which to represent the elements of $X$. Another way to view this is that they are determining a low dimensional subspace of interest, in this case span $(E)$. The algorithm presented here uses techniques from dimension reduction to give an alternate method for determining a low dimensional space of interest. We shall use kernel eigenmap methods to map the high dimensional space $X$ to a low dimensional space $Y$. Unlike endmember algorithms, $Y$ will not be determined as a subspace $X$, but rather through a nonlinear mapping.

We will then construct frame by which to represent the space $Y$. Akin to endmember extraction algorithms, this frame $\Phi$ can be a subset of $Y$. We will also present methodologies by which to construct a data dependent frame from scratch that is not a subset of $Y$. Regardless of how they are constructed, unlike endmember sets, these frames will provide overcomplete representations, a fact we shall exploit.

There are many techniques for dimension reduction, e.g., Principal Component Analysis (PCA) [23], Locally Linear Embedding (LLE) [28], Isomap [32], genetic algorithms, and neural networks. We are interested in a subfamily of these techniques known as kernel eigenmap methods. These include Kernel PCA [29], LLE, Hessian

LLE (HLLE) [18], and Laplacian eigenmaps [7]. Kernel eigenmap methods require two steps.

1. Construction of a symmetric, positive semi-definite kernel (a matrix), $K$, for given data and a specific type of dimension reduction problem to solve.
2. Diagonalization of $K$ to obtain the eigenmaps (eigenvectors).

We shall interpret the data and kernel dependent Hilbert space $\mathbb{K}$, mentioned in section 3.4 for general kernel eigenmap methods, in terms of the theory of frames. Frames provide non-orthogonal overcomplete signal decompositions. In dealing with dimension reduction, our experiments to compare spectral signatures illustrate that different classes are almost never orthogonal, whereas eigenmap methods provide processed orthogonal decompositions. On the contrary, frame elements are not necessarily orthogonal. As such, for given data, they can be constructed to reflect empirical non-orthogonal angular relations between classes. Driven by this mathematical modeling in terms of frames, and inspired solely by given data, we describe an innovative methodology to achieve class separability and object identification.

### 3.3 Kernel Eigenmap Methods

Given a high dimensional data set $X=\left\{x_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{D}$, we assume that the data points $x_{i}$ in fact lie on a low dimensional manifold $M^{d}$, where $d$ is the dimension of the manifold, and $d<D$. As an example, see figure 3.4 , where we have a collection of points in $\mathbb{R}^{2}$ that lie on a one dimensional manifold. Dimension reduction methods construct a mapping from $\mathbb{R}^{D}$ to $\mathbb{R}^{d}$, and in particular map $X$

Figure 3.4: Points in $\mathbb{R}^{2}$ on a one dimensional manifold

to low dimensional coordinates $Y=\left\{y_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{d}$, where $x_{i} \mapsto y_{i}$. The main goal of these methods is to have the new coordinates $Y$ preserve the underlying geometric structure of the manifold $M^{d}$.

Kernel eigenmap methods are a subset of dimension of reduction methods. The key component of these methods is the construction of a data dependent, $N \times N$, symmetric, positive semi-definite kernel $K$ :

$$
\begin{equation*}
K_{i, j}=K\left(x_{i}, x_{j}\right), \quad \forall i, j=1, \ldots, N \tag{3.3.6}
\end{equation*}
$$

The kernel $K$ is then diagonalized, and the $d$ significant eigenvectors of $K$ are retained. Let $v_{1}, \ldots, v_{d} \in \mathbb{R}^{N}$ denote these eigenvectors; the new low dimensional coordinates $Y$ are then given by:

$$
\begin{equation*}
y_{i}=\left(v_{1}(i), \ldots, v_{d}(i)\right), \quad \forall i=1, \ldots, N . \tag{3.3.7}
\end{equation*}
$$

### 3.3.1 Spectral Clustering

We now give a more in depth overview of the theory behind kernel eigenmap methods. Recall

$$
\begin{equation*}
X=\left\{x_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{D} \tag{3.3.8}
\end{equation*}
$$

We then define a distance $\rho: \mathbb{R}^{D} \times \mathbb{R}^{D} \longrightarrow \mathbb{R}^{+}$such that

$$
\begin{align*}
\rho\left(x_{i}, x_{j}\right) & =\rho\left(x_{j}, x_{i}\right), \quad \forall i, j=1, \ldots, N,  \tag{3.3.9}\\
\rho\left(x_{i}, x_{i}\right) & =0, \quad \forall i=1, \ldots, N .
\end{align*}
$$

Since $\mathbb{R}^{D}$ is a vector space, we can define $\rho$ in terms of a norm $\|\cdot\|$, where $\rho\left(x_{i}, x_{j}\right)=$ $\left\|x_{i}-x_{j}\right\|,\|\cdot\|: \mathbb{R}^{D} \longrightarrow \mathbb{R}^{+}$, and $\|0\|=0$. We then compute

$$
\begin{align*}
\widetilde{A} & =\left(\widetilde{A}_{i, j}\right)_{i, j=1}^{N},  \tag{3.3.10}\\
\widetilde{A}_{i, j} & =\rho\left(x_{i}, x_{j}\right) .
\end{align*}
$$

$\widetilde{A}$ has many nonzero entries and is therefore computationally intensive to diagonalize. We think of $\widetilde{A}$ as global information, since it gives the 'distance' between any two points in $X$.

For each $x_{i} \in X$, let

$$
\begin{equation*}
\mathcal{N}_{k}\left(x_{i}\right)=\left\{k \text { nearest neighbors of } x_{i} \text { with respect to } \rho\right\} . \tag{3.3.11}
\end{equation*}
$$

In order to find $\mathcal{N}_{k}\left(x_{i}\right)$ for a fixed $i$, we order the elements of $\left\{\widetilde{A}_{i, j}\right\}_{j=1}^{N}$ :

$$
\begin{equation*}
0=\widetilde{A}_{i, i} \leq \widetilde{A}_{i, \sigma(1)} \leq \widetilde{A}_{i, \sigma(2)} \leq \ldots \leq \widetilde{A}_{i, \sigma(k)} \leq \ldots \leq \widetilde{A}_{i, \sigma(N-1)} \tag{3.3.12}
\end{equation*}
$$

where $\sigma \in S_{N}$. We then set

$$
\begin{equation*}
\mathcal{N}_{k}\left(x_{i}\right)=\left\{x_{\sigma(j)}: j=1, \ldots, k\right\} \tag{3.3.13}
\end{equation*}
$$

Our adjacency matrix $A=\left(A_{i, j}\right)$ is then given by

$$
A_{i, j}= \begin{cases}1, & x_{j} \in \mathcal{N}_{k}\left(x_{i}\right)  \tag{3.3.14}\\ 0, & \text { otherwise }\end{cases}
$$

$A$ has zeros down its diagonal and is not necessarily symmetric. If we want $A$ to be symmetric we could use $\varepsilon$-balls instead of the $k$ nearest neighbors, in which case we would replace $\mathcal{N}_{k}\left(x_{i}\right)$ with $B_{\varepsilon}\left(x_{i}\right)=\left\{x_{j}: \rho\left(x_{i}, x_{j}\right)<\varepsilon\right\}$.

We now define our directed graph $G=\{X, E\}$ where

$$
\begin{equation*}
E \subset X \times X \text { and }\left(x_{i}, x_{j}\right) \in U \Longleftrightarrow A_{i, j}=1 \tag{3.3.15}
\end{equation*}
$$

The weight matrix $\widetilde{W}=\left(\widetilde{W}_{i, j}\right)$ is then given by

$$
\begin{equation*}
\widetilde{W}_{i, j}=h\left(\widetilde{A}_{i, j}^{2}\right) \cdot A_{i, j} \tag{3.3.16}
\end{equation*}
$$

where $h$ has exponential decay at $\infty$, e.g. $h(x)=e^{-x}$. Finally, we define a normalizing matrix $D=\left(D_{i, j}\right)$ where

$$
D_{i, j}=\left\{\begin{array}{cc}
\sum_{l} \widetilde{W}_{i, l}, & i=j  \tag{3.3.17}\\
0, & i \neq j
\end{array}\right.
$$

We then set

$$
\begin{equation*}
W=D^{-1} \widetilde{W} \tag{3.3.18}
\end{equation*}
$$

$W$ contains local information on the relative distances between points; we can think of $W_{i, j}$ as the probability of walking from $x_{i}$ to $x_{j}$. We now examine the following diagram, which shows two points, $x_{i}$ and $x_{j}$, that are two edges apart.


We see that the probability of walking from $x_{i}$ to $x_{j}$ is given by $W_{i, p} W_{p, j}+W_{i, q} W_{q, j}$. But this is just an example of the following identity

$$
\begin{equation*}
W_{i, j}^{2}=\sum_{\substack{x_{p} \in \mathcal{N}_{k}\left(x_{i}\right) \\ x_{j} \in \mathcal{N}_{k}\left(x_{p}\right)}} W_{i, p} W_{p, j} . \tag{3.3.19}
\end{equation*}
$$

Thus $W_{i, j}^{2}$ is the probability of walking from $x_{i}$ to $x_{j}$ in exactly two steps. More generally we have

$$
\begin{equation*}
W_{i, j}^{l}=\text { the probability of walking from } x_{i} \text { to } x_{j} \text { in exactly } l \text { steps. } \tag{3.3.20}
\end{equation*}
$$

We now look at the following example. Consider the following graph, depicted in figure 3.5, where we assume the probability of walking to any given neighbor is equal to that of some other neighbor (the arrows illustrate this point). Furthermore,

Figure 3.5: Diffusion distance

assume the graph is embedded in $\mathbb{R}^{2}$ and take $\rho$ as the Euclidean distance. Clearly $\widetilde{A}_{1,2}<\widetilde{A}_{1,3}$. However, $W_{1,2}^{5}=W_{1,3}^{5}$. This is called diffusion similarity, and thus we call $W$ the diffusion matrix.

Recall we have a directed graph $G=\{X, E\}$, where $X=\left\{x_{i}: i=1, \ldots, N\right\}$. Let $f: X \rightarrow \mathbb{R}$. Since $X$ has $N$ elements, we can think of $f$ as a vector in $\mathbb{R}^{N}$. However, each $x_{i} \in \mathbb{R}^{D}$, so $f$ takes $\mathbb{R}^{D}$ into $\mathbb{R}$, i.e. $f: \mathbb{R}^{D} \rightarrow \mathbb{R}$. We want to define
$\nabla f \in \mathbb{R}^{D}$.

Think of the directed edges of $G$ as vectors. Define

$$
\begin{equation*}
u_{j}=\lambda x_{j}+(1-\lambda) x_{i}, \quad 0 \leq \lambda \leq 1 . \tag{3.3.21}
\end{equation*}
$$

Assuming $f$ is a linear function on $\mathbb{R}^{D} \rightarrow \mathbb{R}$ we have $f\left(u_{j}\right)=\lambda f\left(x_{j}\right)+(1-\lambda) f\left(x_{i}\right)$.

Thus we set

$$
\begin{equation*}
\nabla_{u_{j}} f=f\left(x_{j}\right)-f\left(x_{i}\right) . \tag{3.3.22}
\end{equation*}
$$

We now want to reparameterize to take into account the weights

$$
\begin{align*}
& 0 \leq \lambda \leq \frac{1}{\sqrt{W_{i, j}}}, \\
& u_{j}=\sqrt{W_{i, j}} \lambda x_{j}+\sqrt{W_{i, j}}(1-\lambda) x_{i},  \tag{3.3.23}\\
& \nabla_{u_{j}} f=\sqrt{W_{i, j}}\left(f\left(x_{j}\right)-f\left(x_{i}\right)\right) .
\end{align*}
$$

We now define

$$
\begin{align*}
\operatorname{div} f & =\sum_{j=1}^{N} \nabla_{u_{j}} f  \tag{3.3.24}\\
\Delta & =-\operatorname{div} \cdot \nabla \tag{3.3.25}
\end{align*}
$$

Therefore

$$
\begin{align*}
\Delta f\left(x_{i}\right) & =\sum_{x_{j} \in \mathcal{N}_{k}\left(x_{i}\right)} W_{i, j}\left(f\left(x_{j}\right)-f\left(x_{i}\right)\right) \\
& =-f\left(x_{i}\right)+\sum_{x_{j} \in \mathcal{N}_{k}\left(x_{i}\right)} W_{i, j} f\left(x_{j}\right)  \tag{3.3.26}\\
& =(-I+W) f . \tag{3.3.27}
\end{align*}
$$

(3.3.26) shows that our definition of $\Delta$ is in fact a good one, i.e., $\Delta f\left(x_{i}\right)=0$ if and only if $f$ satisfies the mean value property on $\mathcal{N}_{k}\left(x_{i}\right)$. (3.3.27) shows that $W-I$ is the discrete Laplacian operator.

We now describe the two types of kernels used in this research.

### 3.3.2 Locally Linear Embedding

For each $x_{i} \in X$ compute $k$-nearest neighbors of $x_{i}, \mathcal{N}_{k}\left(x_{i}\right)$, and construct the directed graph described in section 3.3.1. Furthermore, we assume that the graph $G$ is connected. The weights are computed by solving the following minimization problem:

$$
\begin{equation*}
W=\arg \min _{\widetilde{W}} \sum_{i=1}^{N}\left|x_{i}-\sum_{x_{j} \in \mathcal{N}\left(x_{i}\right)} \widetilde{W}_{i, j} x_{j}\right|^{2}, \tag{3.3.28}
\end{equation*}
$$

subject to the constraint:

$$
\begin{equation*}
\sum_{j=1}^{N} \widetilde{W}_{i, j}=1, \quad \forall i=1, \ldots, N \tag{3.3.29}
\end{equation*}
$$

Notice that equation (3.3.28) can be rewritten line by line:

$$
\begin{equation*}
W_{i, .}=\arg \min _{\widetilde{W}_{i}, \cdot}\left|x_{i}-\sum_{x_{j} \in \mathcal{N}\left(x_{i}\right)} \widetilde{W}_{i, j} x_{j}\right|^{2} . \tag{3.3.30}
\end{equation*}
$$

The LLE kernel is then defined as:

$$
\begin{equation*}
K=(I-W)^{\star}(I-W) \tag{3.3.31}
\end{equation*}
$$

which, when compared to (3.3.27), one sees that

$$
\begin{equation*}
K=\Delta^{2} . \tag{3.3.32}
\end{equation*}
$$

The eigenvectors of $K$ will have nonnegative eigenvalues; there will be one with an eigenvalue of zero. The $d$ significant eigenvectors are given by those eigenvectors that correspond the $d$ smallest nonzero eigenvalues.

### 3.3.3 Laplacian Eigenmaps

Like LLE, we use $k$-nearest neighbors to determine the neighborhoods of each $x_{i} \in X$ and we construct the graph $G$. The weights are defined as:

$$
\widetilde{W}_{i, j}= \begin{cases}\exp \left\{-\left\|x_{i}-x_{j}\right\|_{\ell^{2}}^{2} / \sigma\right\}, & \text { if } x_{j} \in \mathcal{N}\left(x_{i}\right) \text { or } x_{i} \in \mathcal{N}\left(x_{j}\right)  \tag{3.3.33}\\ 0, & \text { otherwise }\end{cases}
$$

where $\sigma$ is a positive real number. Define the $N \times N$, diagonal matrix $D$ that same as in equation (3.3.17):

$$
D_{i, j}= \begin{cases}\sum_{l=1}^{N} \widetilde{W}_{i, l}, & i=j  \tag{3.3.34}\\ 0, & i \neq j\end{cases}
$$

We then set the Laplacian eigenmap kernel to be:

$$
\begin{equation*}
K=D-\widetilde{W} \tag{3.3.35}
\end{equation*}
$$

Notice that

$$
\begin{equation*}
K=-D \Delta \tag{3.3.36}
\end{equation*}
$$

For Laplacian eigenmaps, the eigenmaps are obtained by solving the following generalized eigenvector problem:

$$
\begin{equation*}
K f=\lambda D f \tag{3.3.37}
\end{equation*}
$$

Like LLE, the we select the $d$ eigenvectors corresponding to the $d$ smallest non-zero eigenvalues.

### 3.4 Theoretical Foundations of the Algorithm

This algorithm, which appears mathematical and is mathematically sound, is the foundation for our computational work. It is not, however, a direct transcription
of the actual computations, but rather the inspiration for them. The differences between the theoretical work here and the actual computations are detailed in section

## 3.5 .

Given a data set $X=\left\{x_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{D}$, we create the kernel, $K$, using existing kernel methods such as locally linear embedding (LLE), Laplacian eigenmaps, and Hessian LLE. $K \in \mathcal{M}_{N, N}(\mathbb{R})$ is a square matrix of size $N$ and rank $r$. Furthermore, we construct $K$ to be positive semi-definite, i.e., for all vectors $f \in \mathbb{R}^{N}$, we have $f^{\star} K f \geq 0$, where $f^{\star}$ is the transpose of the complex conjugate of $f$.

As with any square matrix, we can diagonalize $K$. In particular, there exists $V \in \mathcal{M}_{N, N}(\mathbb{R})$ and $\left\{\lambda_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{+}$such that

$$
\begin{equation*}
V V^{\star}=V^{\star} V=I \tag{3.4.38}
\end{equation*}
$$

and

$$
\begin{equation*}
K=V \operatorname{diag}\left(\lambda_{i}\right) V^{\star} \tag{3.4.39}
\end{equation*}
$$

Furthermore, by Mercer's theorem [26], there exists a reproducing kernel Hilbert space $\mathbb{K}$ and a set of vectors $\left\{\psi_{i}: i=1, \ldots, N\right\} \subset \mathbb{K}$, such that for all $i, j=1, \ldots, N$,

$$
\begin{equation*}
\left\langle\psi_{i}, \psi_{j}\right\rangle_{\mathbb{K}}=K_{i, j} \tag{3.4.40}
\end{equation*}
$$

We note that $\operatorname{dim} \mathbb{K}=\operatorname{rank}(K)=r$. Furthermore, let $\mathcal{B}=\left\{b_{i}: i=1, \ldots, r\right\}$ be any orthonormal basis for $\mathbb{K}$. Then there exists $\widetilde{Y} \in \mathcal{M}_{N, r}(\mathbb{R})$ such that for each $i=1, \ldots, N$ we have

$$
\begin{equation*}
\psi_{i}=\sum_{j=1}^{r} \widetilde{Y}_{i, j} b_{j} \tag{3.4.41}
\end{equation*}
$$

where

$$
\begin{equation*}
\widetilde{Y}_{i, j}=\left\langle\psi_{i}, b_{j}\right\rangle_{\mathbb{K}} . \tag{3.4.42}
\end{equation*}
$$

It is clear then that

$$
\begin{equation*}
\tilde{Y} \tilde{Y}^{\star}=K \tag{3.4.43}
\end{equation*}
$$

and that $\operatorname{rank}(\widetilde{Y})=\operatorname{dim} \mathbb{K}=\operatorname{rank}(K)=r$. We note that this $N \times r$ matrix $\widetilde{Y}$, while not the same as the set $Y=\left\{y_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{d}$ first introduced in section 3.3, plays a similar role.

Lemma 3.4.1. We can choose $\mathcal{B}$ so that

$$
\begin{equation*}
\widetilde{Y}=V_{N, r}^{\prime} \operatorname{diag}\left(\sqrt{\lambda_{i}}\right)_{r, r} \tag{3.4.44}
\end{equation*}
$$

where $V_{N, r}^{\prime}$ is the matrix of columns of $V$ (i.e. the eigenvectors of $K$ ) that correspond to non-zero $\lambda_{i}$.

Proof. Consider the singular value decomposition of $\tilde{Y}$ : there exists $U_{1} \in \mathcal{M}_{N, N}(\mathbb{R})$, $U_{2} \in \mathcal{M}_{r, r}(\mathbb{R})$, and $\left\{\omega_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}$ such that

$$
\begin{align*}
& U_{1} U_{1}^{\star}=U_{1}^{\star} U_{1}=I,  \tag{3.4.45}\\
& U_{2} U_{2}^{\star}=U_{2}^{\star} U_{2}=I \tag{3.4.46}
\end{align*}
$$

and

$$
\begin{equation*}
\tilde{Y}=U_{1} \Omega U_{2}^{\star} \tag{3.4.47}
\end{equation*}
$$

where $\Omega=\left[\frac{\operatorname{diag}_{r, r}\left(\omega_{i}\right)}{\mathbf{0}}\right]_{N, r}$. Note that $\mathbf{0}$ denotes a block of zeros. Hence,

$$
\begin{equation*}
K=\widetilde{Y} \widetilde{Y}^{\star}=U_{1} \Omega U_{2}^{\star} U_{2} \Omega^{\star} U_{1}^{\star}=U_{1} \Omega \Omega^{\star} U_{1}^{\star} \tag{3.4.48}
\end{equation*}
$$

where

$$
\begin{equation*}
\Omega \Omega^{\star}=\left[\frac{\operatorname{diag}\left(\left|\omega_{i}\right|^{2}\right) \mid \mathbf{0}}{\mathbf{0} \mid \mathbf{0}}\right] \tag{3.4.49}
\end{equation*}
$$

Therefore, the $\omega_{i}$ are uniquely determined (up to a phase factor) by the eigendecomposition of $K$, and so are the columns of $U_{1}$ that correspond to non-zero $\omega_{i}$.

We pick $s \in \mathbb{Z}$, with $r \leq s \leq N$. Let $\Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\}$ be a FUNTF for $\mathbb{K}$. There exists a coefficient matrix $C \in \mathcal{M}_{N, s}(\mathbb{R})$ such that for each $i=1, \ldots, N$ we have

$$
\begin{equation*}
\psi_{i}=\sum_{j=1}^{s} C_{i, j} \varphi_{j} \tag{3.4.50}
\end{equation*}
$$

We can choose

$$
\begin{equation*}
C_{i, j}=\left\langle\psi_{i}, \varphi_{j}\right\rangle_{\mathbb{K}} \tag{3.4.51}
\end{equation*}
$$

but these are not the unique coefficients for which equation (3.4.51) is valid. We can also represent $\Phi$ in terms of the basis $\mathcal{B}$, i.e., there exists $Z \in \mathcal{M}_{s, r}(\mathbb{R})$ such that for each $i=1, \ldots, s$, we have

$$
\begin{equation*}
\varphi_{i}=\sum_{j=1}^{r} Z_{i, j} b_{j} \tag{3.4.52}
\end{equation*}
$$

We note that $\Phi$ is a FUNTF if and only if $Z^{\star} Z=\frac{s}{r} I$, or alternatively, if and only if $\left\|Z^{\star} Z\right\|_{F R O}=\frac{s^{2}}{r}$. Combining equations (3.4.42) and (3.4.52), we have

$$
\begin{equation*}
\left\langle\psi_{i}, \varphi_{j}\right\rangle_{\mathbb{K}}=\sum_{l=1}^{r}\left\langle\psi_{i}, b_{l}\right\rangle_{\mathbb{K}}\left\langle b_{l}, \varphi_{j}\right\rangle_{\mathbb{K}}=\sum_{l=1}^{r} \widetilde{Y}_{i, l} Z_{j, l} \tag{3.4.53}
\end{equation*}
$$

Thus, one possible coefficient matrix is given by:

$$
\begin{equation*}
C=\tilde{Y} Z^{\star} \tag{3.4.54}
\end{equation*}
$$

Inspired by the above calculations, we define how to find the FUNTF $\Psi$ and set up a more general method for computing $C$. Let $c_{1}$ and $c_{2}$ denote cost functions on
the space of matrices $\mathcal{M}_{N, s}(\mathbb{R})$. We want to find a basis $\mathcal{B}=\left\{b_{i}: i=1, \ldots, r\right\}$ for $\mathbb{K}$ and a FUNTF $\Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\}$ for $\mathbb{K}$ that satisfy the following minimization problem:

$$
\begin{equation*}
\min _{\widetilde{\mathcal{B}}, \widetilde{\Phi}} c_{1}\left(\widetilde{Y} Z^{\star}\right), \quad \text { subject to } \quad \text { (3.4.44) } \tag{3.4.55}
\end{equation*}
$$

where $\widetilde{\mathcal{B}}$ is any basis for $\mathbb{K}$ and $\widetilde{\Phi}$ is any FUNTF for $\mathbb{K}$. Recall that by equation (3.4.41) $\widetilde{Y}$ is completely determined by $\widetilde{\mathcal{B}}$ and that by equation (3.4.52) $Z$ is completely determined by $\widetilde{\mathcal{B}}$ and $\widetilde{\Phi}$.

Given a FUNTF $\Phi$, we then want to find a coefficient matrix $C$ that satisfies (3.4.50). Using our second cost function $c_{2}$, we find $C$ by solving the following minimization problem:

$$
\begin{equation*}
C=\arg \min _{\widetilde{C}} c_{2}(\widetilde{C}), \quad \text { subject to } \quad(3.4 .50), \tag{3.4.56}
\end{equation*}
$$

where $\widetilde{C}$ is any possible coefficient set.

### 3.5 The Algorithm in Practice

In practice the algorithm consists of the following five steps:

1. Landmarking
2. Kernel eigenmap method
3. Out of sample extension
4. Frame construction
5. Frame coefficients

The main differences between the actual algorithm and the theoretical ideas are the following. Steps one and three, landmarking and out of sample extension, are employed for certain kernels on large data sets so that the algorithm can run in a reasonable amount of time. Also, the frame construction and handling of kernels are viewed from a more practical point of view, and are implemented correspondingly. We give detailed explanations below.

### 3.5.1 Landmarking

For certain kernels to be used on large scale data sets, landmarking must be employed. For our purposes, we use landmarking only when applying the LLE kernel. Laplacian eigenmaps, with its simple kernel construction, we have found feasible on the data sets of interest. In the case of Laplacian eigenmaps, one my think of the algorithm as skipping steps one and three.

The idea of landmarking is to determine a subset of $X$ that will be used to compute the kernel $K$, as opposed to using all of $X$. We denote this subset as

$$
\begin{equation*}
X_{\text {sam }}:=\left\{x_{i_{j}}: j=1, \ldots, n\right\} \subset X \tag{3.5.57}
\end{equation*}
$$

where $n$ is the number of samples and we assume that $n \ll N$. To obtain $X_{\text {sam }}$, we sample $X$ uniformly at random without replacement.

### 3.5.2 Kernel Eigenmap Methods

We apply the LLE and Laplacian eigenmap kernel eigenmap methods. LLE is used for HSI terrain data, while Laplacian eigenmaps is applied to MSI biological
data. Unlike in section 3.4 where the work is done in the kernel space $\mathbb{K}$, we are forced to diagonalize $K$ in practice and use the traditional reduced coordinates $Y=\left\{y_{i}=\left(v_{1}(i), \ldots, v_{d}(i)\right): i=1, \ldots, N\right\}$, which were described in section 3.3. As a matter of notation, we shall denote the reduced dimensional coordinates of $X_{\text {sam }}$ as $y_{i_{j}} \in \mathbb{R}^{d}$. The reason for returning to the traditional methodology is the following: both LLE and Laplacian eigenmaps were designed to be run on a connected graph $G$, and as such, the rank of these kernels is $r=N-1$. Thus $\operatorname{dim} \mathbb{K}=N-1$, where in practice $N$ is on the order of $10^{6}$. To work in such a space is computationally infeasible, and so a subspace must be determined. The natural first subspace to try is the one given by the eigendecomposition. It should be noted that in the future the development of a true frame based kernel, with low rank, would at least theoretically be optimal.

### 3.5.3 Out of Sample Extension

Given the $n$ low dimensional coordinates $\left\{y_{i_{j}}: j=1, \ldots, n\right\}$ corresponding to the sampled set $X_{\text {sam }}=\left\{x_{i_{j}}: j=1, \ldots, n\right\} \subset X$, we wish to extend these new coordinates to all of $X$ via an out of sample extension [9]. To do so we extend the definition of $k$-nearest neighbors to include reference points $x_{i} \in X$ that are out of sample, i.e., for all $x_{i} \in X \backslash X_{\text {sam }}$, we define:
$\mathcal{N}_{k}^{\prime}\left(x_{i}\right):=\left\{x_{i_{j}} \in X_{\text {sam }}: x_{i_{j}}\right.$ is one of the $k$ nearest neighbors of $x_{i}$ with respect to $\left.\rho\right\}$.

Notice that while the reference point may now come from $X$, the neighbors are still selected only from the sampled subset $X_{\text {sam }}$. In the case of LLE, we must similarly define weights for $x_{i} \in X \backslash X_{\text {sam }}$. Let $W^{\prime}\left(x_{i}, \cdot\right)=\left(W^{\prime}\left(x_{i}, x_{i_{1}}\right), \ldots, W^{\prime}\left(x_{i}, x_{i_{n}}\right)\right)$, and define it as:

$$
\begin{equation*}
W^{\prime}\left(x_{i}, \cdot\right)=\arg \min _{\widetilde{W}^{\prime}\left(x_{i}, \cdot\right)}\left|x_{i}-\sum_{x_{i_{j}} \in \mathcal{N}^{\prime}\left(x_{i}\right)} \widetilde{W}^{\prime}\left(x_{i}, x_{i_{j}}\right) x_{i_{j}}\right|^{2}, \tag{3.5.59}
\end{equation*}
$$

subject to the constraint:

$$
\begin{equation*}
\sum_{j=1}^{n} \widetilde{W}^{\prime}\left(x_{i}, x_{i_{j}}\right)=1, \quad \forall x_{i} \in X \backslash X_{\text {sam }} \tag{3.5.60}
\end{equation*}
$$

The low dimensional coordinates for $x_{i} \in X \backslash X_{\text {sam }}$ are then given by:

$$
\begin{equation*}
y_{i}=\sum_{j=1}^{n} y_{i_{j}} W^{\prime}\left(x_{i}, x_{i_{j}}\right) \tag{3.5.61}
\end{equation*}
$$

thus giving a complete set of low dimensional coordinates $Y=\left\{y_{i}: i=1, \ldots, N\right\}$ that correspond to $x_{i} \mapsto y_{i}$.

### 3.5.4 Frame Construction

Given that we have already departed from the theoretical methodology outlined in section 3.4, it is only natural that our frame construction algorithms deviate as well. Again we forgo working in the kernel space $\mathbb{K}$ for the simpler reduced coordinates given by $Y=\left\{y_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{d}$. Given the new coordinates $Y$, we construct a frame $\Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\} \subset \mathbb{R}^{d}$, where $s \geq d$, for the space $\operatorname{span}(Y)$. We then represent the coordinates $Y$ in terms of the frame $\Phi$. In this research we have used two methods to construct frames, detailed below.

### 3.5.4.1 Endmember Frames

When dealing with HSI terrain data, which is initially processed using the LLE kernel eigenmap method, we have applied existing endmember algorithms to the low dimensional coordinates $Y$. In particular, we have extensively tested the support vector data description (SVDD) endmember algorithm $[6,31]$ within this framework. The benefit of using an existing endmember algorithm such as SVDD is that it is fast, immediately available, and gives a means by which to compare our new framework with an existing endmember algorithm.

### 3.5.4.2 Maximum Separation Frames

We have also developed frame construction algorithms based on modified versions of the frame potential. Through the use of penalty terms, we are able to guide the frame to separate out various features within the data. More specifically, for a FUNTF $\Phi=\left\{\varphi_{i}: i=1, \ldots, s\right\}$ and coordinates $Y=\left\{y_{i}: i=1, \ldots, N\right\}$, define the following penalty function, $p$, as follows:

$$
\begin{equation*}
p\left(\varphi_{i}\right):=\sum_{j=1}^{N}\left|\left\langle y_{j}, \varphi_{i}\right\rangle\right| . \tag{3.5.62}
\end{equation*}
$$

We also set:

$$
\begin{equation*}
p(\Phi):=\sum_{i=1}^{s} p\left(\varphi_{i}\right) \tag{3.5.63}
\end{equation*}
$$

For a given $t, 0 \leq t \leq s$, and $\varepsilon, 0 \leq \varepsilon \leq 1$, we then compute a 'separated' FUNTF $\Phi$ by solving the following modified frame potential:

$$
\Phi=\arg \min _{\widetilde{\Phi} \in \mathbb{S}^{d-1} \times \ldots \times \mathbb{S}^{d-1}} F P(\widetilde{\Phi})
$$

$$
\begin{gathered}
\text { subject to } \\
\sum_{i=t+1}^{s} \frac{p\left(\tilde{\varphi}_{i}\right)}{p(\Phi)}=\sum_{i=t+1}^{s} \sum_{j=1}^{N} \frac{\left|\left\langle y_{j}, \tilde{\varphi}_{i}\right\rangle\right|}{p(\Phi)}<\varepsilon
\end{gathered}
$$

The above frame construction has been applied to biological data, which is first processed using Laplacian eigenmaps.

### 3.5.5 Frame Coefficients

Given reduced coordinates $Y=\left\{y_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{d}$ and a frame $\Phi=$ $\left\{\varphi_{i}: i=1, \ldots, s\right\} \subset \mathbb{R}^{d}$ for $\operatorname{span}(Y)$, we are left to compute coefficients $C=\left\{C_{i, j}:\right.$ $i=1, \ldots, N ; j=1, \ldots, s\}$ to represent $Y$ in terms of $\Phi$. We use two types of coefficients: canonical and sparse.

### 3.5.5.1 Canonical Coefficients

Given a frame $\Phi$, recall the definition of the frame operator $S$ :

$$
\begin{equation*}
S(f)=\sum_{i=1}^{s}\left\langle f, \varphi_{i}\right\rangle \varphi_{i} . \tag{3.5.65}
\end{equation*}
$$

The canonical coefficients of $\Psi$ are then given by:

$$
\begin{equation*}
C_{i, j}=\left\langle y_{i}, S^{-1}\left(\varphi_{j}\right)\right\rangle, \quad \forall i=1, \ldots, N, \quad j=1, \ldots, s \tag{3.5.66}
\end{equation*}
$$

It is well known that the canonical coefficients satisfy the following reconstruction formula:

$$
\begin{equation*}
y_{i}=\sum_{j=1}^{s} C_{i, j} \varphi_{j}, \quad \forall i=1, \ldots, N . \tag{3.5.67}
\end{equation*}
$$

Furthermore, the canonical coefficients are easy and fast to compute, especially so when $\Phi$ is a FUNTF and $S=\frac{s}{d} I$.

### 3.5.5.2 Sparse Coefficients

To compute sparse coefficients for a frame $\Phi$ we solve an $\ell^{1}$ minimization problem for each $y_{i} \in Y$. Let $C_{i, .}=\left(C_{i, 1}, \ldots, C_{i, s}\right)$; we then compute:

$$
\begin{equation*}
C_{i, \cdot}=\arg \min _{\widetilde{C}_{i,},}\left\|\widetilde{C}_{i, \cdot}\right\|_{\ell^{1}}, \quad \text { subject to } \quad \sum_{j=1}^{s} \widetilde{C}_{i, j} \varphi_{j}=y_{i} \tag{3.5.68}
\end{equation*}
$$

We use $\ell^{1}$ minimization as a substitute for the following $\ell^{0}$ minimization problem:

$$
\begin{equation*}
\arg \min _{\widetilde{C}_{i,}}\left\|\widetilde{C}_{i,},\right\|_{\ell^{0}}, \quad \text { subject to } \quad \sum_{j=1}^{s} \widetilde{C}_{i, j} \varphi_{j}=y_{i} \tag{3.5.69}
\end{equation*}
$$

where $\|f\|=\# \operatorname{supp}(f)$. Solving (3.5.69) is NP hard and requires and exhaustive combinatorial search, thus making it intractable. Via the theory of compressed sensing $[12,13,17]$, it has been shown that (3.5.68) can, in certain situations, be used as a direct substitute to (3.5.69), or at the very least, a good approximation. Furthermore, (3.5.68) is a convex optimization problem, and can be solved (reasonably quickly) using linear programming techniques. While the sparse coefficients are more computationally intensive than the canonical coefficients, they can provide enhanced separation of classes, especially when considered from a visual perspective.

## Chapter 4

## Empirical Results

In this chapter we present empirical results derived from running our algorithm on real world hyperspectral and multispectral data sets. These results are broken into two main categories:

1. Hyperspectral terrain data
(a) Urban
(b) Smith Island
2. Multispectral eye data

We give more details on each category, as well as the results, in the sections below.

### 4.1 Hyperspectral Terrain Data

We have two hyperspectral terrain data sets, Urban and Smith Island [3, 2, $4,1,5]$. For these data sets we use the LLE kernel for the dimension reduction, and the SVDD endmember algorithm to construct a frame; we use both canonical and sparse coefficients. For both Urban and Smith Island we have a small subset of ground truth training data, which gives sample pixels of each class contained within the data. Since we have some ground truth, numerically based classification
and comparison is possible; we also present frame coefficient images as well as class maps.

### 4.1.1 Classification Methodology

Once again denote our HSI data set as $X=\left\{x_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{D}$. Let $T \subset X$ denote our ground truth, and suppose there are $q$ classes within $T$. Let $T_{i} \subset T, i=1, \ldots, q$, denote the set of pixels corresponding to class $i$, so that:

$$
\begin{align*}
& \bigcup_{i=1}^{q} T_{i}=T  \tag{4.1.1}\\
& T_{i} \cap T_{j}=\emptyset, \quad \forall i \neq j .
\end{align*}
$$

In order to perform classification on the set $X$, we first construct average representative vectors for each class $T_{i}$. Denote the elements of $T_{i}$ (and therefore $T$ as well) as:

$$
\begin{equation*}
T_{i}=\left\{t_{i, j}: j=1, \ldots, q_{i}\right\} \tag{4.1.2}
\end{equation*}
$$

where $q_{i}$ denotes the number of pixels in the class $T_{i}$; note that, by definition,

$$
\begin{equation*}
\sum_{i=1}^{q} q_{i}=\# T \tag{4.1.3}
\end{equation*}
$$

The average representative vector for the class $T_{i}$, denoted $\tilde{t}_{i}$, is given by:

$$
\begin{equation*}
\tilde{t}_{i}=\frac{1}{q_{i}} \sum_{j=1}^{q_{i}} t_{i, j} \tag{4.1.4}
\end{equation*}
$$

Let $\widetilde{T}$ denote the set of average representative vectors, i.e.,

$$
\begin{equation*}
\widetilde{T}=\left\{\tilde{t}_{i}: i=1, \ldots, q\right\} \tag{4.1.5}
\end{equation*}
$$

We then classify each vector $x_{i} \in X$ by comparing $x_{i}$ with the elements of $\widetilde{T}$. We use the spectral angle between $x_{i}$ and each $\tilde{t}_{j}$ as the determining factor, where the
angle between two vectors is given by:

$$
\begin{equation*}
\theta_{x_{i}, \tilde{t}_{j}}=\cos ^{-1}\left(\frac{\left\langle x_{i}, \tilde{t}_{j}\right\rangle}{\left\|x_{i}\right\|\left\|\tilde{t}_{j}\right\|}\right) . \tag{4.1.6}
\end{equation*}
$$

If the angle between $x_{i}$ and $\tilde{t}_{j_{0}}$ is smaller than the angle between $x_{i}$ and all other $\tilde{t}_{j}$, then we place $x_{i}$ in class $j_{0}$. Mathematically speaking, if

$$
\begin{equation*}
j_{0}=\arg \min _{j=1, \ldots, q} \theta_{x_{i}, \tilde{t}_{j}}, \tag{4.1.7}
\end{equation*}
$$

then $x_{i}$ is placed in class $j_{0}$. To obtain numerical statistics, we use the ground truth data set $T$, and its subsets corresponding to classes, $T_{i}, i=1, \ldots, q$. For each $t_{i, j} \in T_{i}$, we see if the spectral angle classifier indeed places $t_{i, j}$ in class $i$. This allows us to determine a percentage correct for each class, as well as for the ground truth data set as a whole.

Note that we can extend this classification method to reduced coordinates $Y$ and coefficient coordinates $C$ by simply computing everything in terms of these coordinates. In these situations, the indexes of $T$ would remain the same, but we would now assume that $T \subset Y$ or $T \subset\left\{C_{i,}: i=1, \ldots, N\right\}$, respectively. $\widetilde{T}$ would thus be computed again in terms of these new coordinates as well. In the sections to follow we present spectral angle classification statistics based on ground truth for the original data set $X=\left\{x_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{D}$, the LLE low dimensional coordinates $Y=\left\{y_{i}: i=1, \ldots, N\right\} \subset \mathbb{R}^{d}$, SVDD endmember coefficients $\left\{\alpha_{i,}: i=1, \ldots, N\right\}$ computed according to (3.1.4) and (3.1.5), as well as our frame coefficients $\left\{C_{i,}: i=1, \ldots, N\right\}$ computed according to (3.5.66) and (3.5.68).

### 4.1.2 Overview of the Trials

We have run our algorithm on the Urban and Smith Island data sets, which entails the following steps. First process the data set through the LLE kernel eigenmap method to obtain reduced coordinates $Y$. We then compute a frame for $\operatorname{span}(Y)$ using the SVDD algorithm. A frame coefficient cube $C$ is then produced, and we run the spectral angle classification method on these new frame coefficient vectors to obtain class maps and statistical data.

There are two trials for the Urban data set and one for the Smith data set. For every trial there were three variable parameters: the reduced dimension of the LLE coordinates, $d$, the number of frame elements, $s$, and type of frame coefficients - canonical or sparse. We have run each trial through a variety of choices for $d$ and $s$ and computed the canonical or minimum $\ell^{2}$ error coefficients for each iteration, depending on whether $s \geq d$ (canonical) or $s<d$ (minimum $\ell^{2}$ error). We then ran spectral angle classification on each coefficient cube. For each trial, we highlight a particular $d$ and $s$ that had one of the highest overall percentages correct. For this particular $d$ and $s$ we then also compute the sparse coefficients, and present the corresponding statistics and maps.

There are also three different competing results, each serving as means of comparison to our algorithm. First among these is that we classify the raw, unprocessed, Urban and Smith Island data sets. Secondly, we use the LLE reduced coordinates alone, varying the reduced dimension $d$ and selecting the best one according to percentage. Note that the range for $d$ is the same regardless of whether it is for our
algorithm or the LLE coordinates alone. We also run solely SVDD on the Urban data set, this time varying the number of endmembers, $s$. For each $s$ we compute the minimum $\ell^{2}$ error coefficients, and highlight the particular $s$ value with the highest overall percentage. For this particular $s$ we then also compute the mixed $\ell^{2}-\ell^{1}$ coefficients, and present these results as well. Again note, we use the same parameters when running SVDD alone as in the context of our algorithm, although in this case that does not necessarily mean that the $s$ values are the same since there is no way to directly control the number of endmembers/frame elements that SVDD returns.

### 4.1.3 Urban

### 4.1.3.1 Description of the Urban Data Set

The Urban data set is a hyperspectral imagery data set that is freely available at:
http : //www.agc.army.mil/Hypercube/index.html

The dimensions of Urban are $307 \times 307 \times 161$ : that is $307 * 307=94249$ pixels and 161 spectral bands. A pseudocolor image of the Urban data set is given in figure 4.1. There are 932 ground truth pixels, broken into 22 distinct classes; these classes are:

1. AsphaltDrk
2. AsphaltLgt
3. Concrete01
4. VegPasture
5. VegGrass
6. VegTrees01
7. Soil01
8. Soil02
9. Soil03Drk
10. Roof01Wal
11. Roof02A
12. Roof02BGvl
13. Roof03LgtGray
14. Roof04DrkBrn
15. Roof05AChurch
16. Roof06School
17. Roof07Bright
18. Roof08BlueGrn
19. TennisCrt
20. PoolWater
21. ShadedVeg
22. ShadedPav


Figure 4.1: Pseudocolor image of Urban

### 4.1.3.2 Urban Trial 1

The results of Urban trial 1 were obtained with the following settings:

- Data set: $X=$ Urban
- Kernel: LLE
- Number of neighbors: $k=20$
- Number of samples: $\# X_{\text {sam }}=20000$ pixels
- Frame construction: SVDD

The classification results for varying $s$ and $d$ and the canonical coefficients are displayed in figure 4.2. Note that not every combination of $d$ and $s$ have a result due to the nature of the SVDD algorithm, it is only possible to 'guide' the number of endmembers by tweaking certain parameters, there is no direct way to select $s$. Also, the black line represents the line where $d=s$. We highlight the following


Figure 4.2: Urban trial 1 canonical coefficients classification results for varying $s$ and $d$
particular cases.
Urban Trial 1 A

- Number of reduced dimensions: $d=25$
- Number of frame elements: $s=57$
- Type of coefficients: canonical

Statistical results for Urban trial 1 A can be found in table 4.1. Figure 4.4 shows the class map for this trial, while figures 4.8 and 4.9 show the individual class maps. Figures $4.16,4.17,4.18$, and 4.19 show the coefficient maps for each of the frame elements.

## Urban Trial 1 B

- Number of reduced dimensions: $d=25$
- Number of frame elements: $s=57$
- Type of coefficients: sparse

Statistical results for Urban trial 1 B can be found in table 4.2. Figure 4.5 shows the class map for this trial, while figures 4.10 and 4.11 show the individual class maps. Figures $4.20,4.21,4.22$, and 4.23 show the coefficient maps for each of the frame elements.

### 4.1.3.3 Urban Trial 2

For Urban trial 2 we increased the number of neighbors over trial 1, but otherwise kept the settings the same:

- Data set: $X=$ Urban
- Kernel: LLE
- Number of neighbors: $k=40$
- Number of samples: $\# X_{\text {sam }}=20000$ pixels

Table 4.1: Urban trial 1 A ground truth results

|  | \# | \# correct | \% correct | \# false positives | \# false negatives |
| :---: | :---: | :---: | :---: | :---: | :---: |
| AsphaltDrk | 45 | 45 | 100\% | 8 | 0 |
| AsphaltLgt | 26 | 21 | 81\% | 9 | 5 |
| Concrete01 | 64 | 54 | 84\% | 0 | 10 |
| VegPasture | 116 | 116 | 100\% | 3 | 0 |
| VegGrass | 65 | 63 | 97\% | 12 | 2 |
| VegTrees01 | 123 | 85 | 69\% | 8 | 38 |
| Soil01 | 52 | 51 | 98\% | 0 | 1 |
| Soil02 | 24 | 20 | 83\% | 6 | 4 |
| Soil03Drk | 27 | 27 | 100\% | 0 | 0 |
| Roof01Wal | 57 | 57 | 100\% | 1 | 0 |
| Roof02A | 44 | 43 | 98\% | 3 | 1 |
| Roof02BGvl | 17 | 15 | 88\% | 5 | 2 |
| Roof03LgtGray | 12 | 10 | 83\% | 0 | 2 |
| Roof04DrkBrn | 39 | 39 | 100\% | 5 | 0 |
| Roof05AChurch | 38 | 34 | 89\% | 0 | 4 |
| Roof06School | 28 | 28 | 100\% | 0 | 0 |
| Roof07Bright | 35 | 35 | 100\% | 0 | 0 |
| Roof08BlueGrn | 21 | 15 | 71\% | 0 | 6 |
| TennisCrt | 47 | 42 | 89\% | 4 | 5 |
| PoolWater | 5 | 3 | 60\% | 0 | 2 |
| ShadedVeg | 17 | 9 | 53\% | 31 | 8 |
| ShadedPav | 30 | 24 | 80\% | 1 | 6 |
| Total | 932 | 836 | 90\% | 96 | 96 |

Table 4.2: Urban trial 1 B ground truth results

|  | \# | \# correct | \% correct | \# false positives | \# false negatives |
| :---: | :---: | :---: | :---: | :---: | :---: |
| AsphaltDrk | 45 | 45 | 100\% | 8 | 0 |
| AsphaltLgt | 26 | 21 | 81\% | 5 | 5 |
| Concrete01 | 64 | 60 | 94\% | 3 | 4 |
| VegPasture | 116 | 116 | 100\% | 2 | 0 |
| VegGrass | 65 | 63 | 97\% | 10 | 2 |
| VegTrees01 | 123 | 80 | 65\% | 7 | 43 |
| Soil01 | 52 | 44 | 85\% | 1 | 8 |
| Soil02 | 24 | 19 | 79\% | 3 | 5 |
| Soil03Drk | 27 | 26 | 96\% | 0 | 1 |
| Roof01Wal | 57 | 57 | 100\% | 1 | 0 |
| Roof02A | 44 | 43 | 98\% | 3 | 1 |
| Roof02BGvl | 17 | 15 | 88\% | 11 | 2 |
| Roof03LgtGray | 12 | 11 | 92\% | 3 | 1 |
| Roof04DrkBrn | 39 | 31 | 79\% | 4 | 8 |
| Roof05AChurch | 38 | 35 | 92\% | 0 | 3 |
| Roof06School | 28 | 28 | 100\% | 0 | 0 |
| Roof07Bright | 35 | 35 | 100\% | 0 | 0 |
| Roof08BlueGrn | 21 | 15 | 71\% | 0 | 6 |
| TennisCrt | 47 | 40 | 85\% | 3 | 7 |
| PoolWater | 5 | 3 | 60\% | 2 | 2 |
| ShadedVeg | 17 | 11 | 64\% | 38 | 6 |
| ShadedPav | 30 | 23 | 77\% | 7 | 7 |
| Total | 932 | 821 | 88\% | 111 | 111 |

- Frame construction: SVDD

The classification results for varying $s$ and $d$ and the canonical coefficients are displayed in figure 4.3. We highlight the following particular cases.


Figure 4.3: Urban trial 2 canonical coefficients classification results for varying $s$ and $d$

## Urban Trial 2 A

- Number of reduced dimensions: $d=44$
- Number of frame elements: $s=86$
- Type of coefficients: canonical

Statistical results for Urban trial 2 A can be found in table 4.3. Figure 4.6 shows the class map for this trial, while figures 4.12 and 4.13 show the individual class maps. Figures $4.24,4.25,4.26,4.27$, and 4.28 show the coefficient maps for each of

Table 4.3: Urban trial 2 A ground truth results

|  | $\#$ | \# correct | $\%$ correct | \# false positives | \# false negatives |
| :--- | :---: | :---: | :---: | :---: | :---: |
| AsphaltDrk | 45 | 45 | $100 \%$ | 0 | 0 |
| AsphaltLgt | 26 | 20 | $77 \%$ | 1 | 6 |
| Concrete01 | 64 | 64 | $100 \%$ | 1 | 0 |
| VegPasture | 116 | 116 | $100 \%$ | 2 | 0 |
| VegGrass | 65 | 64 | $98 \%$ | 9 | 1 |
| VegTrees01 | 123 | 88 | $72 \%$ | 4 | 35 |
| Soil01 | 52 | 52 | $100 \%$ | 0 | 0 |
| Soil02 | 24 | 22 | $92 \%$ | 1 | 2 |
| Soil03Drk | 27 | 27 | $100 \%$ | 0 | 0 |
| Roof01Wal | 57 | 57 | $100 \%$ | 3 | 0 |
| Roof02A | 44 | 44 | $100 \%$ | 0 | 0 |
| Roof02BGvl | 17 | 17 | $100 \%$ | 1 | 0 |
| Roof03LgtGray | 12 | 11 | $92 \%$ | 1 | 1 |
| Roof04DrkBrn | 39 | 39 | $100 \%$ | 3 | 0 |
| Roof05AChurch | 38 | 37 | $97 \%$ | 0 | 0 |
| Roof06School | 28 | 28 | $100 \%$ | 0 | 0 |
| Roof07Bright | 35 | 35 | $100 \%$ | 0 | 0 |
| Roof08BlueGrn | 21 | 21 | $100 \%$ | 0 | 0 |
| TennisCrt | 47 | 43 | $91 \%$ | 1 | 0 |
| PoolWater | 5 | 3 | $60 \%$ | 0 | 0 |
| ShadedVeg | 17 | 13 | $76 \%$ | 34 | 0 |
| ShadedPav | 30 | 25 | $83 \%$ | 0 | 0 |
| Total | 932 | 871 | $93 \%$ |  | 0 |

the frame elements.

## Urban Trial 2 B

- Number of reduced dimensions: $d=44$
- Number of frame elements: $s=86$
- Type of coefficients: sparse

Statistical results for Urban trial 2 B can be found in table 4.4. Figure 4.7 shows the class map for this trial, while figures 4.14 and 4.15 show the individual class maps. Figures $4.29,4.30,4.31,4.32$, and 4.33 show the coefficient maps for each of the frame elements.

Table 4.4: Urban trial 2 B ground truth results

|  | $\#$ | \# correct | $\%$ correct | \# false positives | \# false negatives |
| :--- | :---: | :---: | :---: | :---: | :---: |
| AsphaltDrk | 45 | 45 | $100 \%$ | 1 | 0 |
| AsphaltLgt | 26 | 20 | $77 \%$ | 0 | 6 |
| Concrete01 | 64 | 61 | $95 \%$ | 1 | 3 |
| VegPasture | 116 | 116 | $100 \%$ | 2 | 0 |
| VegGrass | 65 | 64 | $98 \%$ | 10 | 1 |
| VegTrees01 | 123 | 92 | $75 \%$ | 5 | 31 |
| Soil01 | 52 | 52 | $100 \%$ | 0 | 0 |
| Soil02 | 24 | 24 | $100 \%$ | 1 | 0 |
| Soil03Drk | 27 | 27 | $100 \%$ | 0 | 0 |
| Roof01Wal | 57 | 56 | $98 \%$ | 1 | 1 |
| Roof02A | 44 | 44 | $100 \%$ | 3 | 0 |
| Roof02BGvl | 17 | 17 | $100 \%$ | 0 | 0 |
| Roof03LgtGray | 12 | 12 | $100 \%$ | 4 | 0 |
| Roof04DrkBrn | 39 | 39 | $100 \%$ | 2 | 0 |
| Roof05AChurch | 38 | 38 | $100 \%$ | 0 | 0 |
| Roof06School | 28 | 28 | $100 \%$ | 0 | 0 |
| Roof07Bright | 35 | 35 | $100 \%$ | 0 | 0 |
| Roof08BlueGrn | 21 | 20 | $95 \%$ | 0 | 0 |
| TennisCrt | 47 | 41 | $87 \%$ | 1 | 0 |
| PoolWater | 5 | 3 | $60 \%$ | 0 | 0 |
| ShadedVeg | 17 | 12 | $71 \%$ | 30 | 0 |
| ShadedPav | 30 | 25 | $83 \%$ | 0 | 0 |
| Total | 932 | 871 | $93 \%$ |  | 0 |

### 4.1.3.4 Urban Competing Results

Table 4.5 contains the overall results of the competing Urban results. We note that the LLE and SVDD results were obtained at the following points:

- LLE only (trial 1 ): $d=45$
- LLE only (trial 2): $d=27$
- SVDD (both coefficient cubes): $s=8$

Table 4.5: Urban competing overall results

|  | $\#$ | \# correct | \% correct | \# false pos/neg |
| :--- | :---: | :---: | :---: | :---: |
| Raw data | 932 | 785 | $84 \%$ | 147 |
| LLE only (trial 1) | 932 | 835 | $90 \%$ | 97 |
| LLE only (trial 2) | 932 | 873 | $94 \%$ | 59 |
| SVDD only (min $\ell^{2}$ error coeffs) | 932 | 861 | $92 \%$ | 71 |
| SVDD only (mixed $\ell^{2}-\ell^{1}$ coeffs) | 932 | 334 | $36 \%$ | 598 |

### 4.1.3.5 Urban Class Maps



Figure 4.4: Urban trial 1 A class map


Figure 4.5: Urban trial 1 B class map


Figure 4.6: Urban trial 2 A class map


Figure 4.7: Urban trial 2 B class map

### 4.1.3.6 Urban Individual Class Maps



Figure 4.8: Urban trial 1 A individual class maps 1-9


Figure 4.9: Urban trial 1 A individual class maps 10-22

(a) AsphaltDrk

(d) VegPasture

(g) Soil01

(j) Roof01Wal

(m) Roof03LgtGray

(b) AsphaltLgt

(e) VegGrass

(h) Soil02

(k) Roof02A

(n) Roof04DrkBrn

(c) Concrete01

(f) VegTrees01

(i) Soil03Drk

(l) Roof02BGvl

(o) Roof05AChurch

Figure 4.10: Urban trial 1 B individual class maps 1-15


Figure 4.11: Urban trial 1 B individual class maps 16-22

(a) AsphaltDrk

(d) VegPasture

(g) Soil01

(j) Roof01Wal

(m) Roof03LgtGray

(b) AsphaltLgt

(e) VegGrass

(h) Soil02

(k) Roof02A

(n) Roof04DrkBrn

(c) Concrete01

(f) VegTrees01

(i) Soil03Drk

(l) Roof02BGvl

(o) Roof05AChurch

Figure 4.12: Urban trial 2 A individual class maps 1-15


Figure 4.13: Urban trial 2 A individual class maps 16-22

(a) AsphaltDrk

(d) VegPasture

(g) Soil01

(j) Roof01Wal

(m) Roof03LgtGray

(b) AsphaltLgt

(e) VegGrass

(h) Soil02

(k) Roof02A

(n) Roof04DrkBrn

(c) Concrete01

(f) VegTrees01

(i) Soil03Drk

(l) Roof02BGvl

(o) Roof05AChurch

Figure 4.14: Urban trial 2 B individual class maps 1-15


Figure 4.15: Urban trial 2 B individual class maps 16-22

### 4.1.3.7 Urban Coefficient Maps



Figure 4.16: Urban trial 1 A canonical coefficients 1-12


Figure 4.17: Urban trial 1 A canonical coefficients 13-30


Figure 4.18: Urban trial 1 A canonical coefficients 31-48


Figure 4.19: Urban trial 1 A canonical coefficients 49-57


Figure 4.20: Urban trial 1 B sparse coefficients 1-18


Figure 4.21: Urban trial 1 B sparse coefficients 19-36


Figure 4.22: Urban trial 1 B sparse coefficients 37-54


Figure 4.23: Urban trial 1 B sparse coefficients 55-57


Figure 4.24: Urban trial 2 A canonical coefficients 1-18


Figure 4.25: Urban trial 2 A canonical coefficients 19-36


Figure 4.26: Urban trial 2 A canonical coefficients 37-54


Figure 4.27: Urban trial 2 A canonical coefficients 55-72


Figure 4.28: Urban trial 2 A canonical coefficients 73-86


Figure 4.29: Urban trial 2 B sparse coefficients 1-18


Figure 4.30: Urban trial 2 B sparse coefficients 19-36


Figure 4.31: Urban trial 2 B sparse coefficients 37-54


Figure 4.32: Urban trial 2 B sparse coefficients 55-72


Figure 4.33: Urban trial 2 B sparse coefficients 73-86

### 4.1.4 Smith Island

### 4.1.4.1 Description of the Smith Island Data Set

The dimensions of the Smith Island data set are $679 \times 944 \times 110$ : that is $679 * 944=640976$ pixels and 110 spectral bands. A pseudocolor image of the Smith Island data set is given in figure 4.34. There are 2743 ground truth pixels, broken into 22 distinct classes; these classes are:

1. phrag
2. scirpus
3. juncus
4. patens
5. distichlis
6. andropogon
7. ammophila
8. mud
9. alterniflora
10. borrichia
11. salicornia
12. iva

Figure 4.34: Pseudocolor image of Smith Island

13. pine
14. hardwood
15. pond_water
16. sand
17. wrack
18. myrica
19. seaoats
20. typha
21. water_nshore
22. submerged_nets

### 4.1.4.2 Smith Island Trial 1

The results of Smith Island trial 1 were obtained with the following settings:

- Data set: $X=$ Smith Island
- Kernel: LLE
- Number of neighbors: $k=50$
- Number of samples: $\# X_{\text {sam }}=40000$
- Frame construction: SVDD

The classification results for varying $s$ and $d$ and the canonical coefficients are displayed in figure 4.35. We highlight the following particular cases.

## Smith Island Trial 1 A

- Number of reduced dimensions: $d=21$
- Number of frame elements: $s=69$
- Type of coefficients: canonical

Statistical results for Smith Island trial 1 A can be found in table 4.6. Figure 4.36 shows the class map for this trial, while figures 4.38 and 4.39 show the individual class maps. Figures $4.42,4.43,4.44,4.45$, and 4.46 show the coefficient maps for each of the frame elements.

## Smith Island Trial 1 B

- Number of reduced dimensions: $d=21$


Figure 4.35: Smith Island trial 1 canonical coefficients classification results for varying $s$ and $d$

Table 4.6: Smith trial 1 A ground truth results

|  | \# | \# correct | \% correct | \# false positives | \# false negatives |
| :---: | :---: | :---: | :---: | :---: | :---: |
| phrag | 196 | 138 | 70\% | 68 | 58 |
| scirpus | 246 | 155 | 63\% | 55 | 91 |
| juncus | 184 | 116 | 63\% | 33 | 68 |
| patens | 66 | 57 | 86\% | 33 | 9 |
| distichlis | 97 | 90 | 93\% | 18 | 7 |
| andropogon | 57 | 38 | 67\% | 9 | 19 |
| ammophila | 32 | 25 | 78\% | 29 | 7 |
| mud | 70 | 63 | 90\% | 25 | 7 |
| alterniflora | 200 | 182 | 91\% | 60 | 18 |
| borrichia | 90 | 84 | 93\% | 24 | 6 |
| salicornia | 76 | 58 | 76\% | 3 | 18 |
| iva | 58 | 49 | 84\% | 51 | 9 |
| pine | 166 | 134 | 81\% | 59 | 32 |
| hardwood | 328 | 193 | 59\% | 41 | 135 |
| pond_water | 105 | 69 | 66\% | 3 | 36 |
| sand | 159 | 157 | 99\% | 0 | 2 |
| wrack | 144 | 97 | 67\% | 11 | 47 |
| myrica | 167 | 132 | 79\% | 54 | 35 |
| seaoats | 18 | 13 | 72\% | 0 | 5 |
| typha | 44 | 18 | 41\% | 59 | 26 |
| water_nshore | 206 | 206 | 100\% | 0 | 0 |
| submerged_nets | 34 | 34 | 100\% | 0 | 0 |
| Total | 2743 | 2108 | 77\% | 635 | 635 |

- Number of frame elements: $s=69$
- Type of coefficients: sparse

Statistical results for Smith Island trial 1 B can be found in table 4.7. Figure 4.37 shows the class map for this trial, while figures 4.40 and 4.41 show the individual class maps. Figures $4.47,4.48,4.49$, and 4.50 show the coefficient maps for each of the frame elements.

Table 4.7: Smith trial 1 B ground truth results

|  | \# | \# correct | \% correct | \# false positives | \# false negatives |
| :---: | :---: | :---: | :---: | :---: | :---: |
| phrag | 196 | 129 | 67\% | 64 | 67 |
| scirpus | 246 | 160 | 65\% | 45 | 86 |
| juncus | 184 | 112 | 61\% | 54 | 72 |
| patens | 66 | 54 | 82\% | 31 | 12 |
| distichlis | 97 | 86 | 87\% | 35 | 11 |
| andropogon | 57 | 39 | 68\% | 6 | 18 |
| ammophila | 32 | 25 | 78\% | 36 | 7 |
| mud | 70 | 64 | 91\% | 23 | 6 |
| alterniflora | 200 | 190 | 95\% | 69 | 10 |
| borrichia | 90 | 85 | 94\% | 6 | 5 |
| salicornia | 76 | 57 | 75\% | 2 | 19 |
| iva | 58 | 32 | 55\% | 52 | 26 |
| pine | 166 | 112 | 67\% | 75 | 54 |
| hardwood | 328 | 189 | 58\% | 53 | 139 |
| pond_water | 105 | 70 | 67\% | 1 | 35 |
| sand | 159 | 159 | 100\% | 0 | 0 |
| wrack | 144 | 95 | 66\% | 9 | 49 |
| myrica | 167 | 107 | 64\% | 70 | 60 |
| seaoats | 18 | 13 | $72 \%$ | 0 | 5 |
| typha | 44 | 19 | 43\% | 75 | 25 |
| water_nshore | 206 | 206 | 100\% | 0 | 0 |
| submerged_nets | 34 | 34 | 100\% | 0 | 0 |
| Total | 2743 | 2037 | 74\% | 706 | 706 |

### 4.1.4.3 Smith Island Competing Results

Table 4.8 contains the overall results of the competing Smith Island results. We note that the LLE and SVDD results were obtained at the following points:

- LLE only (trial 1 ): $d=43$
- SVDD (both coefficient cubes): $s=8$

Table 4.8: Smith competing overall results

|  | $\#$ | \# correct | \% correct | \# false pos/neg |
| :--- | :---: | :---: | :---: | :---: |
| Raw data | 2743 | 1957 | $71 \%$ | 786 |
| LLE only (trial 1) | 2743 | 2211 | $81 \%$ | 531 |
| SVDD only (min $\ell^{2}$ error coeffs) | 2743 | 2088 | $76 \%$ | 655 |
| SVDD only (mixed $\ell^{2}-\ell^{1}$ coeffs) | 2743 | 1497 | $55 \%$ | 1245 |

### 4.1.4.4 Smith Island Class Maps



Figure 4.36: Smith trial 1 A class map


Figure 4.37: Smith trial 1 B class map

### 4.1.4.5 Smith Island Individual Class Maps



Figure 4.38: Smith trial 1 A individual class maps 1-9

(a) borrichia

(d) pine

(g) sand

(j) seaoats

(b) salicornia

(e) hardwood

(h) wrack

(k) typha


(c) iva

(f) pond_water

(i) myrica

(l) water_nshore
(m) submerged_nets

Figure 4.39: Smith trial 1 A individual class maps 10-22

(a) phrag

(d) patens

(g) ammophila

(j) borrichia

(m) pine

(b) scirpus

(e) distichlis

(h) mud

(k) salicornia

(n) hardwood

(c) juncus

(f) andropogon

(i) alterniflora

(l) iva

(o) pond_water

Figure 4.40: Smith trial 1 B individual class maps 1-15


Figure 4.41: Smith trial 1 B individual class maps 16-22

### 4.1.4.6 Smith Island Coefficient Maps



Figure 4.42: Smith trial 1 A canonical coefficients 1-12


Figure 4.43: Smith trial 1 A canonical coefficients 13-30


Figure 4.44: Smith trial 1 A canonical coefficients 31-48


Figure 4.45: Smith trial 1 A canonical coefficients 49-66


Figure 4.46: Smith trial 1 A canonical coefficients 67-69


Figure 4.47: Smith trial 1 B sparse coefficients 1-18


Figure 4.48: Smith trial 1 B sparse coefficients 19-36


Figure 4.49: Smith trial 1 B sparse coefficients 37-54


Figure 4.50: Smith trial 1 B sparse coefficients 55-69

### 4.1.5 Conclusions

The numerical statistics show that, for the most part, our algorithm improves upon the raw data and the SVDD method, while remaining even with LLE, at least on the Urban data set. The one exception comes from the LLE coordinates of Smith Island, which when $d=43$, LLE alone attains the best classification results. It should be noted, though, that when restricted to the range $10 \leq d \leq 25$, the LLE coordinates on Smith Island attain a maximum classification percentage of $75 \%$, below what our algorithm attains in the same range for $d$. Similarly, for the Urban data set, larger values of $d$ seem to generate the best classification results for LLE alone, but values of $d$ near the number of classes or below seem to work best when incorporated into our algorithm. Given that both data sets have 22 distinct classes, it would seem plausible that they would lie closer to a manifold with dimension somewhere near 22 as opposed to one with a dimension in the 40 's. Yet with LLE alone the spectral angle classifier desires more and more dimensions, as opposed to our algorithm which seems to prefer a more 'appropriate' low dimensional space. Of course this could also be due to the fact that as $d$ increases, the redundancy of the frame lessens, thus reducing its effectiveness.

Another point to make, however, is the argument of storage versus speed. What we have gained in reducing the size of our data set, we have lost in speed. It is rather quick to run SVDD on the original data sets, even for a large number of different parameters. However, even with the speed ups employed in the kernel process, such as landmarking and the out of sample extension, computing a kernel
on Urban, and especially Smith, is a time consuming process that can take upwards of 24 hours even on a 8 core computer with 16 gigabytes of RAM. Concerning our algorithm, one must weigh the benefits of reduced storage and increased precision at the cost of time. Of course our algorithm does manage to do more with less, at least compared to LLE, and so perhaps can serve as a compromise between endmember algorithms and kernel methods.

Another point in favor of frames, though, comes from figures 4.2, 4.3, and 4.35. These graphs clearly show a drop off in classification results as one goes from having an over-complete frame to an under-complete endmember set, at least when dealing with the reduced coordinates $Y$. Perhaps the same is true in the original space $X$, but given the high dimension that $X$ lies in it is hard to construct a frame with the same redundancy of those constructed for $Y$.

One final comparison comes from the type of coefficients. In one group we have the minimum $\ell^{2}$-error coefficients (for endmembers) and the canonical coefficients (for frames), and in the other group we have the mixed $\ell^{2}-\ell^{1}$ coefficients (for endmembers) and the sparse coefficients (for frames). Despite the added complexity and increased visual appeal of the sparse types of coefficients, it is the simpler $\ell^{2}$ and canonical coefficients that did better in terms of classification. Perhaps for material classification the $\ell^{2} /$ canonical coefficients a preferable, but for material identification, one should go with the sparse coefficients.

### 4.2 Multispectral Retinal Data

The purpose of this experiment is to aid in research concerning age related macular degeneration (AMD), which is one of the leading causes of blindness in the elderly population. One of the indicators of AMD is the presence of irregular lipofuscine deposits, also known as drusin. Using our techniques developed in chapter 3, we present an automated method for the early detection of drusin in retinal imagery.

In order to apply our techniques, we need a high dimensional data set. Through the National Institute of Health, we have obtained a multispectral retinal imagery data set, known to contain drusen. We apply the Laplacian eigenmap kernel to obtain low dimensional coordinates $Y$. We then construct a maximally separated frame $\Phi$ for $\operatorname{span}(Y)$, and represent each $y_{i} \in Y$ in terms of sparse coefficients $C$. The goal is to have certain frame elements correspond to the drusin. We have one trial on the retinal data to illustrate this process.

### 4.2.1 Description of the Retinal Data Set

We work on $500 \times 500 \times 20$ patch of the retinal data set; that is $500 * 500=$ 250000 pixels and 20 spectral bands. A color image of the entire data set is displayed in figure 4.51, with the patch that we work on cut out; that patch is magnified in figure 4.52 . A sample band of the data set is given in figure 4.53 . The only class that we interested in finding is the drusen; there is no ground truth.


Figure 4.51: Color image of entire retinal data set


Figure 4.52: Magnified color image patch


Figure 4.53: Sample band of retinal data set

### 4.2.2 Retinal Data Trial 1

The results of retinal data trial 1 were obtained with the following settings:

- Data set: $X=$ retinal data
- Kernel: Laplacian eigenmaps
- Number of neighbors: $k=12$
- Laplacian eigenmaps sigma parameter: $\sigma=1$
- Frame construction: maximally separated frame
- Number of reduced dimensions: $d=7$
- Number of frame elements: $s=15$
- $t$ parameter in (3.5.64): $t=12$
- $\varepsilon$ parameter in (3.5.64): $\varepsilon=.02$
- Type of coefficients: sparse

Figures $4.54,4.55$, and 4.56 show the coefficient maps for each of the frame elements.

### 4.2.3 Retinal Data Coefficient Maps



Figure 4.54: Retinal data trial 1 sparse coefficients 1-4


Figure 4.55: Retinal data trial 1 sparse coefficients 5-12


Figure 4.56: Retinal data trial 1 sparse coefficients 13-15

### 4.2.4 Conclusions

The first thing to note is that we set $t=12$ in (3.5.64). The idea behind this choice is that the drusin in terms of area in the image are small, and they also have low spectral intensity. Given this knowledge, we allowed 12 of the 15 frame elements to have as large as correlation with the data as they liked, but limited the remaining three to only a $2 \%$ correlation $(\varepsilon=.02)$. Given the physical nature of the drusin and the fact that they are unique, yet make up an extremely small part of the data, it seemed logical that those three restricted frame elements would gravitate towards
the drusin, as opposed to some larger feature. Indeed, of the last three coefficient maps, two of them rather clearly mark drusin (the top two maps in figure 4.56). Furthermore, these two maps seem to mark two separate categories of drusin, the left map marking drusin on the left side of the image, the right map marking drusin on the right side of the image. Perhaps this speaks to a difference between early and later stage drusin, and/or chemical differences. Further investigation is necessary to know for sure.

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[^0]:    ${ }^{1}$ While this section is almost certainly not original, it was independently co-authored by David Widemann, University of California at Davis, and the author.

[^1]:    ${ }^{1}$ All material in this section can be found in [19]

