Oregon Undergraduate Research Journal

Vol. 1 No. 1 (2011): Inaugural Issue

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Vol 1, No 1 (2011): Inaugural Issue

Welcome to the first issue of the Oregon Undergraduate Research Journal!

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Welcome

Lucy Gubbins, Executive Editor

Dear Reader,

It is with great excitement that I welcome you to the inaugural issue of the *Oregon Undergraduate Research Journal*, the University of Oregon's student-run, open access, peer-reviewed scholarly publication. The *OUR Journal* student editorial board has worked for months to produce this first issue and to create a platform for showcasing exceptional undergraduate research.

In creating a publication that makes undergraduate research more accessible to students, my hope is that *OUR Journal* becomes a symbol for what is possible. The value of higher education lies in its ability to spark excitement and intellectual curiosity, and yet few students have the opportunity to explore those passions beyond the term-by-term grind of classes and grades. Students have the right to know that research is not just possible, but a powerful, formative part of the university experience. Beyond this, it is important to stress that research can be valuable to readers when it is published and shared. In this way, I believe *OUR Journal* will not only stand to honor achievement, but also to encourage it.

This issue is representative of the high quality research produced by undergraduates across a wide array of academic disciplines, including history, geology, psychology, political science, communication disorders, and economics. William Goodling's article, "Railroad Antitrust Immunity: Clarification, Discussion, and Evaluation," offers the reader an opportunity to explore the legal ramifications of antitrust immunity abolishment. "Forward Modeling to Assess and Improve Gravity Network Geometry at Kilauea Volcano, Hawai'i," written by Patricia MacQueen, is a fascinating discussion of gravity network geometry in Hawaii and future improvements to volcanic activity monitoring. Tracy Zapf's research explores the important issue of immersion language learning among students with speech disorders in her article, "Acquisition of Second Language Vocabulary for Kindergartners with Speech Sound Disorders." Lauren Dickey provides a timely and politically relevant analysis in her article, "Weapons for Oil: An Analysis of Contemporary Chinese Weapons Sales to Africa in Exchange for Oil." In "Exploring the Adaptiveness of Moderate Dissociation in Response to Betrayal Trauma," Janae Chavez outlines the relationship between dissociation, betrayal trauma, and attachment using a large data set collected through an online survey. In our last article, Neil Cronkrite and Ian O'Gorman, in their article, "Signaling for Attention: Mobility and Student Performance in United Way's Promise Neighborhoods," discuss how student mobility might impact educational development, and how this information might be utilized for non-profit projects.

While *OUR Journal* seeks to highlight undergraduate work, its importance stems from the fact that it is a peer-reviewed, student-led publication. Throughout the summer and fall,

editorial board members worked closely with authors, who had the chance to critically analyze their manuscripts through numerous stages of peer review and editing. Authors had the experience of submitting their essays for publication, and learned how to revise a paper according to anonymous peer comments and copyediting. Student editors learned to adapt to various roles throughout the publication process, from reviewer and editor to author liaison. From ensuring stylistic accuracies during the copyediting phase to creating consistent, appealing article layout, every step was a learning experience. In other words, *OUR Journal* provided an opportunity for both editors and authors to experience the entire process of academic publication, from beginning to end.

I want to thank the members of the editorial board, whom I am incredibly lucky to have the opportunity to work with. I have never before worked with individuals so hard working, energetic, and positive, nor more willing to put up with my many frantic emails. I also want to acknowledge our publisher, journal manager, and faculty advisers for their constant help and support in all aspects of this publication. Finally, I must thank our contributors and their faculty mentors. From July, the *OUR Journal* authors have worked with patience and excitement to be a part of this issue, and for you I am grateful. Your dedication to undergraduate research will inspire a new generation of passionate, curious students.

Sincerely,

Lucy Gubbins,
Executive Editor
Oregon Undergraduate Research Journal Editorial Board

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Some Background on this New Journal

JQ Johnson, University of Oregon Libraries[1]

It is exciting to be present at the creation of a new journal, and especially the new *Oregon Undergraduate Research Journal*. Establishing a peer reviewed journal is a substantial undertaking, and the work that the student editors and various UO faculty and staff have put in has been quite impressive.

Planning for a UO undergraduate journal began more than a year ago, in June 2010, with meetings among interested faculty and administrators. In that early period there were more questions than answers: Should such a journal be student- or faculty-edited? Should it be peer-reviewed, and should it be open access? Which departments on campus would take responsibility for sponsoring the journal? How would it relate to other UO publications that feature student research such as the *UO McNair Journal* or to discipline specific academic journals that might be alternatives for the publication of student work? Etc.

Answers to most of the questions came quickly. Whether this should be a student-produced publication became a non-issue when a group of students came forward and announced that they were planning an undergraduate journal. It's not clear if the administrators co-opted the students or vice versa, but a synergistic collaboration quickly developed. The outcome was a student edited and managed publication with support and sponsorship from the Office of the Vice President for Research and the UO Libraries, plus encouragement and assistance from Student Affairs, Undergraduate Studies, and the Clark Honors College.

Whether this new publication should be a peer-reviewed journal wasn't hard either, since the student editors were clear in their resolve to focus this new journal on the highest quality UO student work. Peer review works well in mainstream academic scholarly publication, and the editors realized – correctly – that by mirroring the process of other academic journals they had the best chance of identifying the most outstanding undergraduate research. It provided a great learning experience for the editors, too; it's a good bet that many of them will become professors and maybe edit their own journals or serve on editorial boards in their chosen disciplines.

For the UO Libraries, the new journal was a great opportunity. The library already had an annual Undergraduate Research Award that recognizes undergraduate students who demonstrate skill and creativity using library resources in their research. In addition, the library has a strong commitment to open access – online, free distribution of scholarly work – as a way to take advantage of modern Internet technology, improve scholarly communications, and make scholarship available to a wider public audience than is possible with traditional journals that may only be subscribed to by a handful of academic research libraries. The Oregon Undergraduate Research Journal plans dovetailed nicely with a new library program to serve as

publisher for UO open access journals using the Open Journal System developed by Simon Fraser University. This journal is the second one published by the UO Libraries, the first being *Humanist Studies and the Digital Age* (available on the same website as *OUR Journal*).

All of the participants in *OUR Journal* are excited about this inaugural issue and looking forward to future issues. We anticipate continued annual publication of high-quality undergraduate research articles, often based on term papers or senior theses. We are also excited by the possibility of taking advantage of technology to widen our net, and to publish student works in the creative and performing arts, multimedia "papers," online computer-based tools that disseminate the results of student programming projects, student-generated datasets, and much more. The core mission will remain, though, to showcase the best University of Oregon undergraduate work.

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[1] Note: JQ Johnson is Director, Scholarly Communications and Instructional Support at the University of Oregon Libraries, and leads the library's efforts to serve as publisher for open access journals such as this one.

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U.S. Railroad Antitrust Immunity: Clarification, Discussion and Evaluation

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Abstract

The U.S. Congress regulated the railroad industry in 1887, and over the course of the 20th century also granted the industry significant antitrust immunities. Antitrust immunities are laws that expressly exempt an industry from prosecution under antitrust laws, such as the Sherman Act. Presumably, the rationale for railroad antitrust immunities was that because railroads were stringently regulated, the regulators alone would uphold antitrust principles and make antitrust litigation unnecessary. However, culminating in the passage of the Staggers Rail Act of 1980, the railroad industry was largely deregulated, yet retained many antitrust immunities. This has raised concerns among shippers and consumers that railroad companies, which often face neither regulation nor antitrust liability, can freely commit anticompetitive abuses. Given these concerns and currently proposed legislation to abolish railroad antitrust immunities, the purpose of this paper is to evaluate the efficacy of legal outcomes in a counterfactual situation where antitrust immunities are abolished. To reach this end, I will first clarify railroad regulation and deregulation, antitrust laws as they apply to all other industries, and the poorly understood railroad antitrust immunities.

> Antitrust laws … are the Magna Carta of free enterprise. They are as important to the preservation of economic freedom and our free enterprise system as the Bill of Rights is to the protection of our fundamental personal freedoms.
>
> Justice Thurgood Marshall (U.S. v. Topco Associates, 1972)

1. Introduction

In a purely competitive market, the number of buyers and sellers is so large that any one of them cannot affect the market price, which through competition is driven down to equal the marginal cost of production. If certain additional assumptions are valid, a competitive market maximizes efficiency and welfare for both producers and consumers (Bernanke & Frank, 2007).

Railroad transportation markets, however, deviate in several ways from the conditions necessary for purely competitive, efficient markets. First, due to the large expenditures required to start a railroad company, railroad companies often exhibit economies of scale. Economies of scale, which are often associated with behemoth firms, violate a requirement for competitive markets and mean that as production expands, average costs decrease. Firms with economies of

scale can drive smaller competitors into bankruptcy by using their low costs to temporarily decrease prices—only to raise them once the competition has been eliminated (a practice known as predatory pricing). Second, in any particular railroad market there is often an insufficient number of competitors for the market price to be driven down to its competitive equilibrium. A railroad market is defined as transporting commodity *X* from discrete locations *A* to *B*. It is usually not in a rail firm's interest to expend a significant outlay to build a rail line directly beside a competitors' rail line, so many railroad markets have a severe dearth of competition—often only one firm. Third, in addition to specific railroad markets, the overall railroad industry is exceedingly concentrated with four firms controlling nearly 90% of the freight traffic in the United States (S. Rep. No. 112-38, 2011). When these industry conditions are present, there is not sufficient competition to make the market efficient, and firms possess market power. Market power is the ability to charge prices above marginal costs, and it implies market inefficiency and welfare losses for consumers (Bernanke & Frank, 2007). Market power can be expressed through prices exceeding marginal costs or though other forms of anticompetitive conduct that will be further discussed in this paper.

When market power is present, there are two ways public policy can counteract the harmful consequences for consumers. First, statuary regulation can prescribe what is acceptable, such as direct price regulation. In Section 2, Railroad Regulation, I will discuss in greater detail the history of the statutory regulation of the railroad industry. Second, parties who are harmed by anticompetitive conduct can use the antitrust laws to sue for damages and injunctions in court. The antitrust laws are a collection of statutes that are intended to promote competition and thus reduce prices, and in Section 3, The Antitrust Laws, I will explain how they apply to virtually all industries.

In 1887, Congress created the Interstate Commerce Commission (ICC), a federal agency that, among other things, had the power to directly regulate railroad rates. The ICC's purpose was to stop firms with market power from raising prices above what would have prevailed in a competitive market. Throughout the 20th century, the railroad industry also accumulated antitrust immunities, which are laws that exempt an industry from prosecution under the antitrust laws. I will further explicate the railroad antitrust immunities, which have been called the "most convoluted story in American antitrust," (Saggers, 2009) below in Section 4, Railroad Antitrust Immunities.

According to the U.S. Supreme Court, the rationale for railroad antitrust immunities was that the heavy ICC regulation alone would be "an effective safeguard against the evils attending monopoly, at which the [antitrust laws] are directed" (*Mclean Trucking Co. v. U.S.*, 1944). However, culminating in the passage of the Staggers Rail Act of 1980, Congress largely revoked the ICC's regulatory power over railroads and gave railroads more flexibility to set their own rates. Despite undermining the rationale for the immunities' existence by deregulating the industry, Congress retained the antitrust immunities. Without regulation or applicable antitrust laws, shippers and consumers are now concerned that railroad companies can freely exploit their market power to engage in anticompetitive conduct and raise prices above marginal costs.

Given these concerns, the purpose of this paper is to analyze the legal ramifications of a counterfactual situation in which antitrust immunities are abolished. To that end, I will first clarify railroad regulation and deregulation (Section 2 below), antitrust laws as they apply to other industries (Section 3 below), and the railroad industry's antitrust immunities (Section 4 below). This is an important and timely investigation because several industry developments are consistent with market power abuse, such as significant upward trends in rates and industry profitability since 2001 and several railroad practices that would arguably merit antitrust liability if not for the railroad antitrust exemptions (S. Rep No. 111-9, 2009). These concerns of market power abuse prompted members of the U.S. Congress to propose current legislation, the Railroad Antitrust Enforcement Act of 2011, to abolish all railroad antitrust immunities (S. 49, 2011).

2. Railroad Regulation

The first federal regulation of interstate railroads began with the Interstate Commerce Act of 1887, which established the Interstate Commerce Commission (ICC) that regulated railroad mergers, routes, and most important, rates (i.e., prices). Through direct regulation, the ICC attempted to set rates consistent with the public interest, and to curtail rate discrimination based on person, place, or length of haul considerations.

Though the authority was implied by the Interstate Commerce Act of 1887, the Hepburn Act of 1906 explicitly granted the ICC power to set maximum rates, and the Transportation Act of 1920 explicitly empowered the ICC to set minimum rates and to regulate all entry and exit of firms from railroad markets. By 1945, the ICC's regulation was expansive, including rates, line abandonment, service discontinuities, mergers, car flow, and interchange rules (Wilson & Burton, 2003), all of which Viscusi et al. (1998) note was "extensively used" (554). Therefore, in this era, federal regulators closely managed nearly all aspects of the railroad industry and widely used their authority to set maximum rates to counteract market power abuse, such as rates exceeding marginal costs. With such dominant regulation, antitrust laws were considered duplicative and unnecessary, so the railroad companies were exempted.

In the 1970s, the rise of competing alternative modes of transportation (truck and barge) and the inflexibility of ICC regulations – which constrained railroads from adequately adapting to the changing market conditions – lead to problems in the railroad industry. Railroad companies had deteriorating tracks, low productivity, poor financial performance, and many went bankrupt. In an attempt to revitalize the industry, Congress deregulated the railroad companies under the Rail Revitalization and Regulatory Reform Act of 1976 (4-R Act) and most importantly under the Staggers Rail Act (1980).

The 4-R Act established a "zone of reasonableness" within which railroads could freely set their rates, created easier procedures for abandoning routes, and eased merger restrictions. However, if the ICC found a firm to be "market dominant" (defined as controlling 70% of all traffic in a rail market or having revenue-to-variable cost ratio exceeding 1.8), then the ICC gained jurisdiction to conduct a rate reasonableness inquiry. If the rate was found unreasonable, the ICC prescribed rate regulation (Wilson, 1996). However, Winston, Corsi, Grimm and Evans

(1990) demonstrated that under the ICC's narrow application of the 4-R Act in the late 1970s, the legislation did not cause a significant change in the railroad regulatory regime, meaning the ICC still had overwhelmingly regulatory control of the railroad industry.

The Staggers Rail Act of 1980 most significantly deregulated the railroad industry. The preamble of the Staggers Rail Act states that it is the policy of the United States Government to "allow, to the maximum extent possible, competition and the demand for services to establish reasonable rates for transportation by rail" (49 USC §10101). Unless the ICC found a railroad to be "market dominant" and then found its rates unreasonable, railroads gained freedom to set their own rates. The Staggers Act also eased merger guidelines, facilitated abandonment of unprofitable segments of network as well as entry into new markets, and encouraged private, confidential contracts in rate setting. Essentially, for the first time since 1887, railroad companies gained the freedom to set their own rates. Coupled with a dearth of competition or the presence of economies of scale or both, this was an invitation for the railroad companies to use their market power to raise rates beyond marginal costs and therefore reduce consumer welfare.

The passage of the ICC Termination Act of 1995 replaced the ICC with a much smaller federal agency, the Surface Transportation Board (STB). The STB has jurisdiction to (1) adjudicate rate reasonableness cases brought by shippers against railroads, (2) prescribe new regulations, and (3) approve rail transactions, such as line sales, line construction, line abandonment, and mergers (Government Accountability Office [GAO], 2006). The STB does not directly regulate rates, so the only venue for regulatory relief from market power abuse is STB rate reasonableness adjudications. However, there is wide concern that the proceedings are "largely inaccessible," "expensive, time consuming, and complex" and may be financially prohibitive (GAO, 2006). Johnstone (2009) and Pittman (2010a) argue that the complexity, expense, and likelihood of failure in these proceedings have rendered them useless for most shippers. Indeed, on average it costs $3.3 million over three years to litigate a rate reasonableness case (GAO, 2006), a price that is prohibitive for all but the biggest shippers.

In sum, the railroad industry transitioned from being heavily regulated in the first half of the 20th century to being significantly deregulated by 1980 and onwards. The ICC's forceful direct regulation once was a justification for antitrust immunities, but today the STB's ineffectual venue of rate reasonableness adjudication rarely constrains railroad company's pricing decisions. Coupled with antitrust immunities, this implies that railroad companies can exploit their market power to raise prices and engage in anticompetitive conduct.

3. The Antitrust Laws

3.1. Statutory Fundamentals

The first major antitrust law was the Sherman Act of 1890, which has two foundational prohibitions:

Section 1: "every contract, combination … or conspiracy in restraint of trade …[is] illegal."

Section 2: "every person who shall monopolize, or attempt to monopolize, or … conspire with any other person … to monopolize … shall be guilty of a misdemeanor."

Section 1 governs mutual conduct, especially acts that coordinate sellers by the use of formal agreements. Mutual seller conduct, such as agreements to fix prices, is an expression of market power antithetical to competitive markets; instead of competing to lower prices, firms are conspiring to jointly raise prices. Likewise, monopolization is repugnant to competitive markets because monopolization is the process of forcing out competitors through illicit means, such as using economies of scale to engage in predatory pricing or undergoing mergers designed to reduce competition.

In 1914, the Federal Trade Commission Act (FTC Act) established the Federal Trade Commission (FTC), an agency designed to perform both investigatory and adjudicative functions. Section 5 of FTC Act created a foundational antitrust law, stating that "unfair methods of competition … and unfair or deceptive acts" are illegal. The FTC has the power to determine what is "unfair" and issue cease and desist orders, which are binding in 60 days unless appealed in court. Only the FTC can bring cases under the FTC Act, and it can only apply civil penalties. However, this law is subject to a railroad antitrust exemption that is further discussed in Section 4: in the railroad industry, only the STB can enforce the FTC Act, not the FTC (15 USC §21[a]).

Perhaps in an attempt to define the vagueness of the Sherman Act, the Clayton Act (1914) explicitly prohibits four forms of conduct that were arguably implied by Section 1 of the Sherman Act's prohibition on "restraint of trade." The four forms of conduct prohibited by the Clayton Act are:

Section 2: price discrimination

Section 3: exclusionary practices (i.e., excluding a competitor by making it a condition of a purchase that the customer cannot later buy from a competitor)

Section 7: mergers "where the effect may be substantially to lessen competition, or to tend to create a monopoly"

Section 8: interlocking directorates

The Clayton Act also increased the available damages for violations of either the Sherman Act or the Clayton Act to treble damages (damages triple what makes the harmed party "whole") and allowed for injunctions to prohibit specific conduct.

3.2. Judicial Interpretation

Based on these broad phrases and prohibitions, antitrust law in the U.S. is created through a common law process of precedents from courts and enforcement agencies to determine what practices are illegal in what situations. While private parties can sue under the Sherman Act and the Clayton Act, the main burden of enforcement falls on two government agencies: civil and criminal actions via the U.S. Department of Justice Antitrust Division, and exclusively civil actions via the Federal Trade Commission (Caves, 1994). Due to the nonspecific nature of the statutes, enforcement has gone through periods of high scrutiny to lax scrutiny. Below is a

survey of the law of (A) monopolization, (B) conspiracies in restraint of trade, such as price-fixing, and (C) select issues in exclusionary conduct.

A. Monopolization (Section 2 of the Sherman Act)

As Viscusi et al. (1998) notes, Section 2 of the Sherman Act has been the most difficult antitrust law for courts to interpret, because the Sherman Act prohibits the process of monopolization but not necessarily the market position of being a monopolist. The market position of being a monopolist is traditionally defined as having a large market share, such as producing 60% of a particular good. Courts must balance the tension between discouraging anticompetitive monopolization without discouraging superior performance and efficiency that may lead to a monopoly.

In 1911, the Supreme Court had its first major test of how it would interpret Section 2 of the Sherman Act. Standard Oil Company and American Tobacco Company both controlled around 90% of their markets, and both firms used a wide range of aggressive conduct towards their rivals. For example, Standard Oil was accused of engaging in predatory pricing to drive competitors out of business by buying pipelines to foreclose crude oil supplies to rivals, securing discriminatory rail freight rates, and conducting business espionage. In *U.S. v. Standard Oil Co.* (1911) and in *U.S. v. American Tobacco Co.* (1911), the Court found both firms guilty under Section 1 and Section 2 of the Sherman Act and ordered their dissolution into several independent firms.

The most important consequence of these cases was the doctrine that the Court established to interpret the Sherman Act. The Court reasoned that the plain text of the Sherman Act was too extreme to be enforceable, and decided that not "every" restraint of trade or attempt to monopolize is illegal, but only those that are "unreasonable" or "injurious" (*U.S. v. American Tobacco Co.*, 1911). For adjudicating Sherman Act cases, determining what is a reasonable or injurious restraint of trade is known as the Rule of Reason doctrine. In *State Oil Co. v Kahn et al.* (1997), the Court summarized the Rule of Reason:

> … [Sherman Act] claims are analyzed under a 'Rule of Reason,' according to which the finder of fact must decide whether the questioned practice imposes an unreasonable restraint on competition, taking into account a variety of factors, including specific information about the relevant business, its condition before and after the restraint was imposed, and the restraint's history, nature, and effect.

In *U.S. v. American Tobacco Co.* (1911) and *U.S. v. Standard Oil Co.* (1911), the Court ruled that the companies' use of their large market share to engage in coercive actions against their rivals was unreasonable and therefore illegal, but also emphasized that high seller concentration alone need not necessarily be unreasonable.

In this era of monopolization jurisprudence, Sherman Act Section 2 violations were difficult to prove, prompting one commentator to call the law a "dead letter" (Gellhorn & Kovacic, 1994). For example, although U.S. Steel Corporation held 60% of the nation's iron and steel capacity, in *U.S. v. U.S. Steel Corp.* (1920) the Court ruled that its acquisitions of competing companies were not an illegal attempt to monopolize. The Court's reasoning was that "the law does not make

mere size an offense," meaning that they interpreted Section 2 to not prohibit the market structure of monopoly, but only to prohibit the conduct of unreasonable monopolization.

The ruling in *U.S. v. Aluminum Co. of America (U.S. v. Alcoa)* (1945) effectively ended the previous Sherman Act Section 2 doctrine. Alcoa controlled 90% of U.S. aluminum production, and even without engaging in overtly unreasonable anticompetitive conduct, its size alone constituted a violation of Section 2 of the Sherman Act. Despite that Section 2 prohibits monopolizing but not monopolies, the Court argued, "A firm can gain and protect a monopoly position in ways more subtle than taking the bloody axe to its competitors." They reasoned that Alcoa engaged in practices that were not overtly predatory, but still effectively resulted in monopolization of the market, such as building capacity ahead of demand to foreclose potential rivals. This ruling suggested that all companies with substantial market share were suspect, and perhaps presumably liable, for monopolizing under Section 2 of the Sherman Act.

Since the early 1970s, the courts have retreated from the encompassing approach of *U.S. v. Alcoa* (1945). In addition to a large market share, the courts are increasingly requiring that the conduct itself be an overt attempt to monopolize. Two major types of conduct that are potentially illegal monopolization are predatory pricing and refusal-to-deal.

Predatory pricing occurs when a firm, supported by economies of scale and a large market share, charges a price below marginal cost to force out competitors. This is illegal monopolization because the only possible benefit of charging a price where the firm is losing money is to bankrupt a competitor so as to later gain a monopoly position and then excessively raise prices. Refusal-to-deal monopolization occurs when a business reduces competition by refusing to sell to another business for the purpose of rending them inoperable. Refusal-to-deal monopolization occurs if the plaintiff can prove:

1. The monopolist's control of an essential facility (or product)
2. A competitor's inability to reasonably duplicate the essential facility (or product)
3. The monopolist's denial of use of the facility (or product) to a competitor
4. The monopolist's feasibility of providing the facility (or product) (Gellhorn & Kovacic, 1994).

This is known as the essential facility doctrine, and it is intended to stop monopolists from exploiting their control of a resource to obstruct competition—such as a railroad firm owning the only track in a rail market and not letting others pay to use it. There is a fifth overriding, case-by-case consideration in all Rule of Reason analyses: whether or not the defendant advanced a reasonable business justification for denying access. If the defendant did advance a reasonable business justification for denying access, then they are not guilty of monopolizing under Section 2 of the Sherman Act.

In sum, monopolization jurisprudence is continually evolving. During the *U.S. v. Alcoa* (1945) era, having a dominant market share (as the four major railroad companies currently do) was sufficient for violation of Section 2 of Sherman Act. More recently, however, merely having a dominant market share is not enough to constitute monopolizing; the firm must also engage in

specific conduct deemed an injurious attempt to monopolize, such as predatory pricing or refusal-to-deal conduct.

B. Conspiracies in Restraint of Trade (Section 1 of the Sherman Act)

Unlike the ebbing judicial interpretation of Section 2 of the Sherman Act, interpretation of Sherman Act Section 1 prohibition on conspiracies (agreements among would-be competitors) to restrain trade has been "the most unambiguous antitrust rule of law" (Viscusi et al., 1998). Per se illegal restraints of trade include:

1. Agreements to fix or maintain prices (U.S. v. Trenton Potteries Co., 1927; U.S. v. Socony-Vacuum Oil Co., 1940)
2. Agreements to limit output or productive capacity (U.S. v. Trenton Potteries Co., 1927; U.S. v. Socony-Vacuum Oil Co., 1940)
3. Agreements to share or divide markets (U.S. v. Topco Associates, Inc., 1972)

A per se violation of Section 1 of the Sherman Act means that it is automatically condemned as illegal, and is not subject to Rule of Reason deliberations about whether the act amounted to an unreasonable restraint of trade. These forms of conduct are so unambiguously adverse to competition and consumer welfare that courts automatically rule them illegal restraints of trade.

C. Exclusionary Conduct (Section 3 of the Clayton Act)

Section 3 of the Clayton Act makes exclusionary conduct illegal. Exclusionary conduct occurs when a company requires buyers of its goods to refrain from purchasing goods from its rivals. Types of exclusionary conduct include tying agreements and exclusive-dealing arrangements.

Tying agreements occur when a company gives a buyer access to one of its goods only if the buyer takes others as well. The courts have enforced violations of tying agreements very strictly (Caves, 1994). To prove a tying agreement is illegal, the plaintiff must prove:

1. There are two distinct products
2. The seller has required the buyer to purchase the tied product in order to obtain the tying product
3. The seller has market power in the market for the tying product
4. The tying arrangement affects a substantial amount of commerce in the market for the tied product (*Grappone, Inc. v. Subaru of New England, Inc.*, 1988).

For example, in *Northern Pacific Railroad Co. v. United States* (1958) the Court found Northern Pacific Railroad guilty of a tying agreement because its sale of land adjacent to its rail lines required the purchasers to ship commodities only via their company.

Exclusive-dealing arrangements occur when a seller gives the buyer access to its goods only if the buyer agrees to not buy goods from any of the seller's rivals. The courts have enforced violations of exclusive-dealing arrangements if dominant sellers use it to place newcomer sellers at a disadvantage (Caves, 1994), and in general, the Courts have treated them harshly (Viscusi et al., 1998). To evaluate the reasonableness of exclusive dealing, courts usually consider three factors: (1) the extent of market foreclosure, (2) the duration of the exclusive dealing, with

agreements less than one year "presumptively lawful" (*Roland Machinery Co. v. Dresser Industry*, 1984), and (3) the height of entry barriers (Gellhorn & Kovacic, 1994). If entry barriers to the market are small, the exclusive-dealing arrangement is likely to be legal; if entry barriers are significant (such as the large expenditures required to start a railroad), exclusive-dealing arrangements are likely to be found illegal.

4. Railroad Antitrust Immunity

The railroad industry was exempted from antitrust laws in an era of direct railroad regulation under the rationale that the regulators alone would promote competition and avoid market power abuse. The antitrust exemptions for the railroad industry are disparate and complex, but nevertheless the exemptions are all in the United States Code and case law. Railroad antitrust exemptions apply only if the Surface Transportation Board (STB) regulates the firm or if the STB approves the action.

4.1. Transactional Immunity

The Transportation Act of 1920 gave the STB exclusive purview over specified railroad transactions, and states that if a transaction is approved by the STB, it is exempt from antitrust laws (49 USC §11321[a]). Transactions that can be "approved or exempted" by the STB include mergers, acquisitions, leases, trackage rights (the right to use but not own rail lines), and joint ownership of rail lines (49 USC § 11323). Therefore, among other transactions, if a merger is approved by the STB, it is expressly exempt from challenge under Section 7 of the Clayton Act. This is a powerful immunity because it has permitted the industry to become extremely concentrated without allowing the Department of Justice to sue for antitrust violations. With four large railroads currently controlling 90% of the railroad market (S. Rep. No. 112-38, 2011), this immunity almost surely has shielded otherwise illegal mergers that "tend to create a monopoly" (Clayton Act).

4.2. Rate Agreements

Enacted by the Reed-Bulwinkle Act of 1948, railroad rate agreements can be exempted from antitrust laws with approval by the STB (49 USC §10706[a][2][A]). If this exemption did not exist, the conduct would likely be a per se illegal violation of Section 1 of the Sherman Act, which prohibits conspiracies in restraint of trade, such as price fixing. Nevertheless, the antitrust exemption permits firms to coordinate their actions. When firms are permitted to coordinate their actions, they have an incentive to mutually raise prices.

4.3. Rate Bureaus

The Reed-Bulwinkle Act of 1948 also allowed the creation of 'rate bureaus.' Rate bureaus are organizations of railroad firms where the companies discuss rates jointly. While rate bureaus are still legal (49 USC §10706[a][2][A]), their activities have been restricted to ameliorate the potential for anticompetitive collusion. Currently under 49 USC §10706(a)(3)(A), once the STB approves a rate, a member in that bureau cannot participate in discussions concerning another firm's rates. Rail firms are also prohibited from discussing rates that another firm proposes for interline services "unless that carrier practicably participates in the movement" (49

USC §[a][3][A][ii]). Rate bureaus must report transcripts and audio recordings of all discussions to the STB. While these restrictions help reduce the potential for anticompetitive conduct in rate bureaus, it is nevertheless questionable why rate bureaus – would-be competitors jointly discussing rates – are exempt from antitrust laws.

4.4. Interlocking Directorates

Under the Transportation Act of 1920, if the STB approves interlocking directorates for railroad firms – a practice otherwise prohibited by Section 8 of the Clayton Act – it is legal and exempt from antitrust challenge (American Bar Association [ABA], 2007). Interlocking directorates are individuals who serve on the board of directors of multiple corporations, which can lead to serious issues of conspiracies in restraint of trade among firms that are ostensibly competitors.

4.5. Line Sales

The STB is authorized to review all sales, creations, and abandonments of railroad lines, and its approval immunizes the transaction from the antitrust laws (49 USC §10901). From the perspective of promoting competition, this immunity is problematic because it can allow a dominant firm to exploit smaller railroad companies to enhance its dominant position. For example, this exemption is the source of exclusionary conduct known as 'paper barriers.' In many line sales, major railroads divest track to regional operators that, after the sale, are connected with the seller's main lines. In the line sale contract, the seller mandates that the buyer only interchange its traffic from the divested line to the seller, precluding the ability to interchange traffic with other railroads (ABA, 2007). If there were no antitrust immunities, this would be a patent violation of Section 3 of the Clayton Act, which prohibits just such exclusionary conduct that artificially reduces competition.

4.6. Pooling Arrangements and Division of Traffic

The STB can approve combinations to pool or divide traffic, services, or revenues between carriers, which indirectly approves a rate agreement (49 USC §11322). If these divisions and pooling of revenues are approved, they are exempt from the antitrust laws (49 USC §11321[a]). If it were not for antitrust immunity, pooling arrangements would be prosecuted as conspiracies in restraint of trade under Section 1 of the Sherman Act. For all other industries, agreements between competitors to pool or divide business are per se illegal because they are adverse to competition (*U.S. v. Topco Associates, Inc.*, 1972).

4.7. No Injunctive Relief for Private Parties

For other industries, under the Clayton Act (15 USC §26), private parties threatened with loss or damage by a violation of the antitrust laws can sue for injunctive relief in a civil action to enjoin an illegal activity. However, Section 16 of the Clayton Act prohibits private parties from suing railroad firms for injunctive relief under any antitrust law.

4.8. The *Keogh* Doctrine: No Treble Damages for Private Parties

Under the *Keogh* doctrine (*Keogh v. Chicago & NW Railway Co.*, 1922), treble damages are unavailable for private shippers who challenge the reasonableness of rates submitted to the STB.

Along with the railroad's immunity from private party injunctions described above in Section 4.7, the *Keogh* doctrine narrows the scope of remedies that private parties can obtain against railroads under the antitrust laws to only singular monetary damages.

4.9. Secretary of Transportation Conferences

Conferences among railroads, shippers, labor organizations, consumer representatives and government agencies may be convened by the Secretary of Transportation, and agreements entered into with the Secretary's approval through these conferences are exempted from antitrust laws (49 USC §333). This immunity grants opaque – and therefore dangerous – power to the Secretary of Transportation.

4.10. Federal Trade Commission Act Enforcement

The Federal Trade Commission Act (FTC Act) prohibits "unfair methods of competition … and deceptive acts" in commerce, and establishes that the Federal Trade Commission (FTC) is the exclusive enforcer of the law. However, Section 5 of the FTC Act (15 USC §45) states that the FTC cannot enforce the law against railroads; rather, the STB has sole authority to enforce compliance with the FTC Act against railroads (15 USC §21[a]). This sweeping antitrust exemption entirely removes one of the foundational antitrust laws from enforcement by any party besides the STB.

To summarize, railroad antitrust exemptions apply if a firm is under STB jurisdiction or if the STB approves the action. Therefore, the STB essentially has exclusive control of anticompetitive issues in the railroad industry, which is why the STB has been described as a "surrogate" to antitrust laws (ABA, 2007). However, the deregulated STB is "largely inaccessible," "time consuming, and complex" (GAO, 2006), and its proceedings are financially prohibitive for the vast majority of shippers (Johnstone [2009] and Pittman [2010a]). Given this, it is problematic that there is no recourse for those who believe that the STB's decisions will be adverse to competition except general rules of administrative procedures (Brennan, 2009).

5. Analysis of the Counterfactual: No Antitrust Immunities

I have described the historical and current regime governing the railroad industry. Because presumably some form of government intervention is necessary to police the naturally monopolistic railroad industry, the efficacy of this particular regime of government intervention in railroad markets is an important investigation. There is a wide literature on the positive effects of the Staggers Rail Act deregulation (summarized in GAO, 2006), so it is not efficacious to have a federal agency heavy-handedly manage industry and establish rail rates (as the ICC did prior to the Staggers Rail Act). However, what is less clear is the desirability of antitrust immunities. What would happen in the railroad industry if there were full antitrust liability, as in virtually all other industries? This is an important consideration because members of the U.S. Congress recently proposed the abolishment of all railroad antitrust immunities in the Railroad Antitrust Enforcement Act of 2011 (S. 49, 2011).

The literature addressing what would happen if antitrust immunities were repealed is limited (Pittman [2010b], Brennan [2009], Massa [2001], and Saggers [2009]). The Staggers

Rail Act deregulation, in combination with continued antitrust immunities, has engendered several deleterious anticompetitive issues for a minority of 'captive shippers.' Even though deregulation under the Staggers Rail Act increased welfare in significant ways, it does not necessarily imply that application of the antitrust laws could not further increase welfare by ameliorating specific anticompetitive conduct that is harming certain shippers. Two important anticompetitive issues are refusal-to-deal in 'bottleneck' markets, and exclusive-dealing contracts known as 'paper barriers.' A third crucial consideration is the overall concentration of the railroad industry. In what follows, I consider what would occur in the counterfactual situation of abolishing antitrust immunities.

5.1. Refusal-to-Deal in 'Bottleneck' Markets

Many rail customers, known as "captive shippers," are served by only a single railroad at either their origin or destination. However, over some portion of the captive shipper's route, another railroad could compete for the shipment service. This market is illustrated in Figure 1:

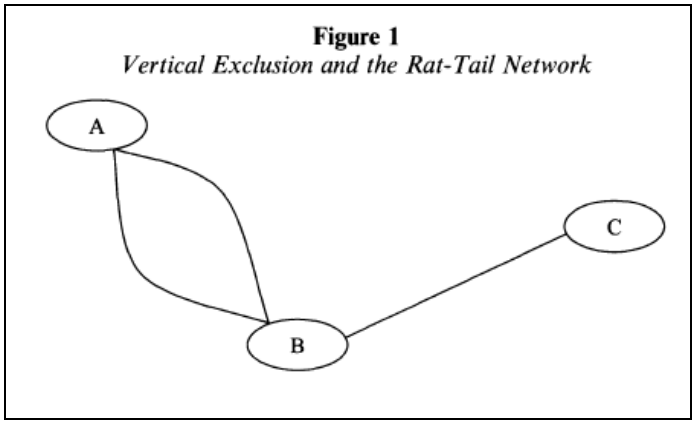

Figure 1. Such rail markets are called "bottleneck markets" because the portion from B to C narrows to just one rail provider, like the neck of a bottle narrows. Source: Wilson and Burton (2006).

Consider a shipper at *C* who needs to ship a product to *A*. From *C* to *B* the shipper is captive to a pure monopolist, but from *B* to *A* there could be two firms competing to lower rates. Currently, the monopolist who controls *C* to *B* forecloses the possibility of competition on the segment from *B* to *A*. The monopolist achieves this by either refusing to grant trackage rights over *C* to *B* to other railroads, or refusing to offer a route that would stop at the point where the carrier could be switched to allow for competition (i.e., only offering *C* to *A* and refusing to offer *C* to *B*). The STB approved these tactics to hold a shipper captive over an entire route, so the conduct gained antitrust immunity (*Central Power and Light Co. v. Southern Pacific Transportation Co.*, 1997).

There is a consensus in the literature that captive shippers pay higher rates than non-captive shippers, which suggests captive shippers' rates are affected by market power abuse (GAO [2006], S. Rep. 111-9 [2009], Ellig [2002], Saggers [2009], Brennan [2009], Pittman [2010b], and S. Rep 112-38 [2011]). Grimm and Winston (2000) estimate that captive shippers pay rates which are 20.9% higher than non-captive shippers. If there were no antitrust immunities, these

practices would be evaluated as refusal-to-deal monopolization violations of Section 2 of the Sherman Act, or as tying arrangements violations of Section 3 of the Clayton Act.

Refusal-to-deal violations of Section 2 of the Sherman Act occur when a business refuses to sell to another business or consumer for the purpose of driving the other business into bankruptcy or to raise prices for the consumer. However, the Court ruled in *Verizon v. Trinko* (2004) that refusal-to-deal would be illegal only if the firm was stopping a practice it had formerly been undertaking, and the firm would lose profits by stopping the practice (except for longer-run monopolist benefits). In addition, in *Monsanto Co. v. Spray-Rite* (1984), the Court ruled that a seller "has a right to deal, or refuse to deal, with whomever it likes, as long as it does so independently." Under this case law, refusal-to-deal violations are rare and difficult to prove (Pittman, 2010b).

In *Verizon v. Trinko* (2004) and *Monsanto Co. v. Spray-Rite* (1984) the essential facility doctrine, which proves refusal-to-deal monopolization, was not relevant; in bottleneck railroad markets, however, the essential facility doctrine is highly relevant. The essential facility doctrine holds that for refusal-to-deal monopolization to be present, the plaintiff must prove (1) the monopolist controls an essential facility, (2) competitors are reasonably unable to duplicate the essential facility, (3) the monopolist denies use of the facility to a competitor, and (4) the monopolist could feasibly provide the facility (Gellhorn & Kovacic, 1994). Based on these criteria, if antitrust immunities were abolished, the refusal-to-deal conduct of the initial rail monopolist would be illegal monopolization. That is, the railroad firm that is a monopolist over *C* to *B* controls an essential facility (the monopoly route); there are prohibitive costs for a competitor to construct a competing line over that segment; the monopolist denies the potential competitor trackage rights; and the monopolist could feasibly offer those trackage rights. In sum, if antitrust immunities were abolished and the plaintiff could successfully argue that the essential facility doctrine is the relevant authority, the conduct in bottleneck markets would be illegal under Sherman Act Section 2 as refusal-to-deal monopolization.

An alternative way to prosecute conduct arising in bottleneck markets is through a Clayton Act Section 3 lawsuit that alleges a tying agreement. Tying agreements occur when a seller gives a buyer access to the seller's service only if the buyer takes other services as well. For the shipper in Figure 1, a railroad only offering shipping from *C* to *A* could be framed as a tying arrangement. If a shipper uses *C* to *B*, then they are forced by necessity to also use *B* to *A*. This is a plausible argument, and tying agreements supported by market power have been enforced strictly (Caves, 1994). As outlined in *Grappone, Inc. v. Subaru of New England, Inc.* (1988), for a tying agreement to be found illegal, the plaintiff must prove: (1) there are distinct services, (2) the seller has required the buyer to purchase the tied product in order to obtain the tying product, (3) the seller has market power in the market for the tying product, and (4) the tying arrangement affects a substantial amount of commerce in the market for the tied service. The monopolist railroad company has market power, and a substantial amount of commerce is involved in any given railroad market, so conditions (3) and (4) are satisfied. What is debatable is whether the railroad is offering two services (*C* to *B* and then *B* to *A*) or only one service (*C* to *A*). If the railroads were to be found in violation of the Clayton Act Section 3 for a tying agreement, the plaintiffs would need to successfully portray the two segments as distinct

services that were tied together solely for the purpose of excluding the competitor. Given the empirical evidence of higher rates for captive shippers, it is a convincing argument that two independent lines are joined for the purpose of exploiting market power to increase rates.

In sum, railroads holding shippers captive over portions of track where competition could prevail is anticompetitive. This conduct could be prosecuted as a refusal-to-deal offense under the essential facility doctrine (Sherman Act Section 2) or a tying arrangement violation (Clayton Act Section 3). If antitrust immunity were removed, litigation would expose the details of these issues to determine legality of conduct. Moreover, as Pittman (2010b) suggests, the mere threat of this litigation could force railroads to negotiate lower rates rather than risk going to court for potentially exploiting captive shippers.

5.2. Exclusive-Dealing Contracts

One goal of the Staggers Rail Act was to decrease regulations that prevented the "rationalizing" of rail lines, which means selling or abandoning unprofitable lines. Due to these deregulations, large and dominant railroads have leased or spun-off portions of track to form smaller regional short line railroads. These transactions often come with contractual conditions – called 'paper barriers' – that require the new buyer of the line to only engage in transit with the previous owner. Line sales are under the STB's jurisdiction, so it has the power to reject these transactions if it believes they are contrary to the public interest. However, the STB has regularly approved these contracts, which then immunizes them from antitrust laws. These contracts are anticompetitive because they decrease shipper's options and expose them to a firm with increased market power. The contractual barriers are additionally inefficient because they result in idle short line capacity even though market demand would otherwise provide productive opportunities.

If antitrust immunity were removed, these paper barriers could be prosecuted under Section 3 of the Clayton Act as tying arrangement or exclusive-dealing violations. As outlined in the bottleneck market context, for a tying agreement to be found illegal, the plaintiff must prove: (1) there are distinct services, (2) the seller has required the buyer to purchase the tied service in order to obtain the tying service, (3) the seller has market power in the market for the tying service, and (4) the tying arrangement affects a substantial amount of commerce in the market for the tied service (*Grappone, Inc. v. Subaru of New England*, 1988). Based on these criteria, paper barriers would be illegal tying arrangements because the seller is tying the sale of its line to its transit service.

A second way paper barriers would be illegal is as exclusive-dealing contracts in violation of Section 3 of the Clayton Act. Exclusive-dealing arrangements occur when a seller gives the buyer access to its goods only if the buyer agrees to not buy goods from any of the seller's rivals. The courts have enforced violations of exclusive-dealing arrangements if dominant sellers use it to place newcomer sellers at a disadvantage (Caves, 1994), and in general the courts have treated them harshly (Viscusi et al., 1998). To evaluate the reasonableness of exclusive-dealing contracts, courts usually consider three factors: (1) the extent of market foreclosure, (2) the duration of the exclusive dealing, and (3) the height of entry barriers (Gellhorn & Kovacic, 1994). Under these criteria, with antitrust immunities abolished, paper barriers would be illegal

exclusive-dealing violations because they force the buyer of the short line to not do business with the seller's rivals; they indefinitely foreclose the market; and they establish significant barriers of entry for competition.

Whether as a tying agreement or an exclusive-dealing agreement under Section 3 of the Clayton Act, if antitrust immunities were abandoned the railroad industries' paper barriers would face antitrust liability. Judicial scrutiny and any appropriate remedies, such as injunctions enjoining the practice, would be welcomed. Shippers who are currently forced to use a monopolist provider to gain access to short lines under paper barriers would enjoy the benefits of increased competition—principally, lower rates.

5.3. Mergers and Industry Concentration

Under the STB's approval of mergers, the U.S. railroad industry has been extensively consolidated: four major railroad companies currently control nearly 90% of the market's revenues (S. Rep. No. 112-38, 2011). This is an astonishing lack of competition, and suggests that railroad companies are not robustly competing to bid down rates. Indeed, in the last decade the four largest railroad firms have annually raised their rates 5% above inflation and have increased their profit margin to 13%, making railroads the fifth most profitable industry in the U.S. (S. Rep. No. 112-38, 2011). This directly harms shippers but also consumers by raising the price of the goods that are shipped, such as foodstuffs and coal for electricity.

The mergers that created the concentration of the railroad industry are exempt from antitrust laws because they were approved by the STB. However, if antitrust immunities were abolished, there would likely be colossal confrontations in court brought by the Department of Justice (Pittman, 2010b) to determine whether the mergers amounted to monopolization under Section 2 of the Sherman Act, or violated Section 7 of the Clayton Act, which prohibit mergers where "the effect may be substantially to lessen competition, or to tend to create a monopoly." A plain reading of this law implies that the mergers would be found illegal, and the firms could be ordered to break into smaller firms that compete with each other. This would help avoid the market power abuse that leads to increased prices for all consumers.

5.4. Other Legal Effects of Antitrust Immunity Repeal

If antitrust immunities were revoked from the railroad industry, their rate agreements and rate bureaus exemptions would be illegal under antitrust law. For example, this would permit a shipper who alleged a rate was excessive due to a collusive rate agreement to sue the railroad under Section 1 of the Sherman Act. Additionally, the current railroad arrangement to pool and divide traffic would be illegal under antitrust law, namely as an illegal market allocation under Section 1 of the Sherman Act. With antitrust immunity abandoned, the FTC would be allowed to prosecute railroad companies under the FTC Act, which prohibits "… unfair or deceptive acts or practices in or affecting commerce…" (15 USC § 45). Given the breadth of the phrase "unfair or deceptive acts," this could entail nearly any conduct that the FTC deems anticompetitive, such as bottleneck market tactics designed to make shippers captive to one company. Lastly, if antitrust immunities were abandoned, the remedies available to private parties would be expanded to include injunctive relief and treble damages, which would encourage more antitrust lawsuits and thus less market power abuse.

6. Conclusion

The regime governing the railroad industry since 1980 – a deregulated federal agency coupled with broad antitrust immunities – in some ways has presided over a period with significant improvements in railroad industry conditions, such as productivity gains and restored financial health. However, extensive industry consolidation is likely responsible for rising prices, and some specific complaints remain for shippers who have excessively limited transportation options. If antitrust immunities were repealed, shippers with grievances (especially captive shippers in bottleneck markets and shippers subject to paper-barriers from line spin-offs) would be able to sue railroad companies for violations of antitrust laws. If the facts substantiated the allegations in the courts, then the result would be a more competitive and efficient railroad market because courts would enjoin railroad conduct that reduces competition.

Antitrust immunities were initially implemented because at the time there was a federal agency that resolved anticompetitive concerns with aggressive regulation. However, now that the industry has been largely deregulated, there is not a sufficiently accessible or responsive mechanism for responding to anticompetitive conduct. By removing antitrust immunities, as members of the U.S. Congress have currently proposed in the Railroad Antitrust Enforcement Act of 2011 (S. 49, 2011), the industry would couple deregulation with a more logically consistent application of the antitrust laws. Under this new regime, as in virtually all other markets, market forces would determine prices, entry, and exit. Furthermore, in the spirit of Justice Thurgood Marshall's principle that antitrust laws are "the Magna Carta of free enterprise" that are fundamental "to the preservation of economic freedom and our free enterprise system" (*United States v. Topco Associates, Inc.*, 1972), if any firm used market power to abuse the competitive marketplace, antitrust lawsuits would be an available remedy to make the marketplace efficient and fair.

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Forward Modeling to Assess and Improve Gravity Network Geometry at Kilauea Volcano, Hawai`i

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Abstract

Scientists have been using campaign gravity surveys to monitor volcanic activity at Kilauea's summit for decades, yet we have a poor understanding of the ability of the existing network to resolve sources of magma accumulation with different mass changes and depths. We also do not yet have a fully quantified measure of the relative importance of the stations in the network. This research tests the network using a simple forward modeling approach over a range of likely source volumes and depths. The analysis determines network sensitivity to three different likely source locations, calculates the relative importance of stations in the network, and examines the problem of signal distortion imposed by network geometry. This work finds that the current network is least sensitive to south caldera sources, and investigates the location and number of stations that will resolve this problem most effectively.

Introduction

Measurements of gravity changes at the summit of a volcano can yield valuable information about subsurface mass movements as magma rises toward the surface. However, how well the network of gravity monitoring stations resolves these changes in mass determines the value of the information a campaign gravity survey provides. A campaign gravity survey measures changes in gravitational acceleration over time at a network of gravity stations whose locations remain constant for all surveys. Since gravitational acceleration is determined by the distribution of mass in the subsurface, measuring the change in gravity over time at these gravity stations gives researchers an idea of mass change over time. Ideally, such a network of stations would be arranged on a grid, to minimize spatial signal distortion that arises from irregular grid spacing. Unfortunately, this is usually effectively impossible due to the complex topography around an active volcano and considerations of time and money. A network has existed on the summit of Kilauea Volcano in Hawaii, United States, for decades (Johnson et al., 2010). Gravity stations are co-located with an existing and more extensive network of benchmarks used for annual leveling surveys.

Kilauea Volcano is a shield volcano located at the southeastern end of the Hawaiian volcanic chain on the island of Hawai'i (Figure 1). Kilauea has been erupting continuously since 1983; its most recent phase of eruption consists of a summit lava lake, a lava pond at Pu`u O`o, effusive

flank eruptions, and a new fissure eruption near Napa`au crater ("HVO Kilauea Status Page," May 2011).

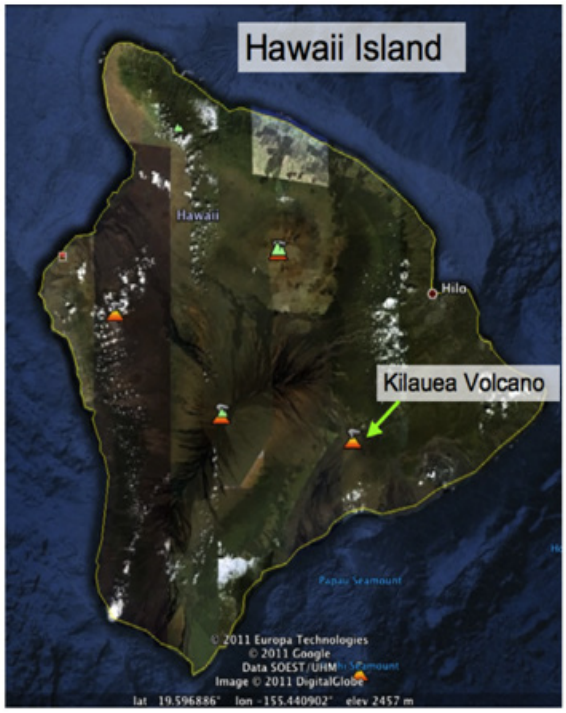


Figure 1. Map of Hawai'i Island with the location of Kilauea Volcano marked.

Although its eruptions are primarily effusive, the United States Geological Survey (USGS) considers Kilauea to be the most dangerous volcano in the United States because of its ongoing activity and the high potential for persons and property to be exposed to volcanic activity from Kilauea and the volcano's east rift zone. The many tourists who flock to Kilauea make hazard mitigation at the volcano difficult (Ewert et al., 2005). In addition to effusive eruptions, Kilauea has erupted explosively in the past. Kilauea's caldera formed in a massive eruption in the 16th century, killed visitors to the caldera in an explosive eruption in 1924, and more recently, a smaller explosion in March 2008 accompanied the formation of the current lava column in Halema`uma`u (Jaggar, 1924; Swanson, 2008; Poland et al., 2009; Fee et al., 2010; Houghton et al., 2011). Recent research has uncovered several other explosive eruptions in Kilauea's history, including an eruption that left tephra sheets blanketing the south caldera region, producing what is known as the Kulanaokuaiki tephra (Fiske et al., 2009). Less dramatic than these explosive eruptions but no less concerning is "vog" (volcanic fog). The gas plume from Kilauea's current summit eruption, full of noxious sulfur dioxide, trails westward over the Kailua-Kona area, creating vog which causes many respiratory problems (Sutton et al., 2000).

Eruptive events at Kilauea are commonly preceded by months to years of magma accumulation in a buried magma reservoir beneath the summit. During eruptive events, this magma drains rapidly, sometimes in a matter of hours, to erupt from vents on the volcano's rift zones. Most monitoring efforts at Kilauea are focused on the summit area as it is theorized that

this is where magma rises from great depths through the crust and into the volcanic edifice (Decker, 1987; Johnson, 1987). GPS, leveling, tilt, seismographic, and gravity instruments all monitor these inflation and deflation events.

Using gravity to monitor volcanoes is crucial for constraining other data sources and providing greater advance warning. Because gravity detects changes in mass, it provides an important constraint on the ascent of magma. The gravitational acceleration measured at any one point on the surface of the earth depends on the distribution of mass around that point. Accordingly, by measuring gravitational acceleration at several points in a given area, we can gain some idea of what lies beneath the surface. If we do such surveys repeatedly in the same area using the same measurement points we can detect changes in gravitational acceleration over time. Changes in gravitational acceleration allow us to detect changes in the distribution of mass over time (Dzurisin, 2007; Battaglia et al., 2008). Gravitational acceleration is determined by either continuous, permanent meters or a series of campaign measurements. A continuous gravity meter gives better temporal resolution but the limited number of stations gives poor spatial resolution. The number of continuous meters is limited by the substantial bunkers needed to protect a meter running continuously in an active volcanic region. A campaign gravity survey consists of many single, discrete measurements of gravity at a set network of benchmarks, which are then repeated at a given interval, usually every few months or years. This method gives better spatial resolution but poor temporal resolution, since surveys are time intensive and performed infrequently, but the campaign stations cover a wider spatial area than is possible with continuous meters (Battaglia et al., 2008).

The distribution of mass that produces a given gravity signal is non-unique, as is the distribution of volume change that produces a deformation signal. Coupling the two independent observations can better resolve the source. Interpretation of deformation data, in turn, benefits from gravity data, as gravity data can help constrain the source of deformation. Deformation can tell us that something is inflating, its location, and its volume, but it cannot tell us anything about the source's density, which can be a key factor in establishing the identity of the source (e.g. magma or hydrothermal). For example, gravity studies at Long Valley Caldera were able to determine that a source of inflation was from basaltic magma saturated with hydrothermal fluids, rather than hydrothermal fluids alone, which has vastly different hazard implications. Hydrothermal activity is typically harmless to humans, whereas the presence of basaltic magma may imply an eruption is possible in the near future (Battaglia et al., 1999, 2003).

On Hawai'i, a gravity survey could help distinguish between a source that is inflating from subsurface accumulation of magma versus a source that is inflating due to vesiculation of an existing magma body. Figure 2 depicts the hypothetical results from gravity and deformation surveys under these two scenarios. In the former scenario, a deformation survey and a gravity survey across the source would show an increase in gravity and uplift at the source, indicating both an increase in volume and an increase in mass from the addition of magma. In the latter scenario, a deformation survey across the source would show an increase in uplift at the source, but a gravity survey (after corrections for the change in elevation) would show no change,

indicating a volume increase without a corresponding increase in mass. This would suggest a decrease in density of source, possibly from vesiculation (Figure 2).

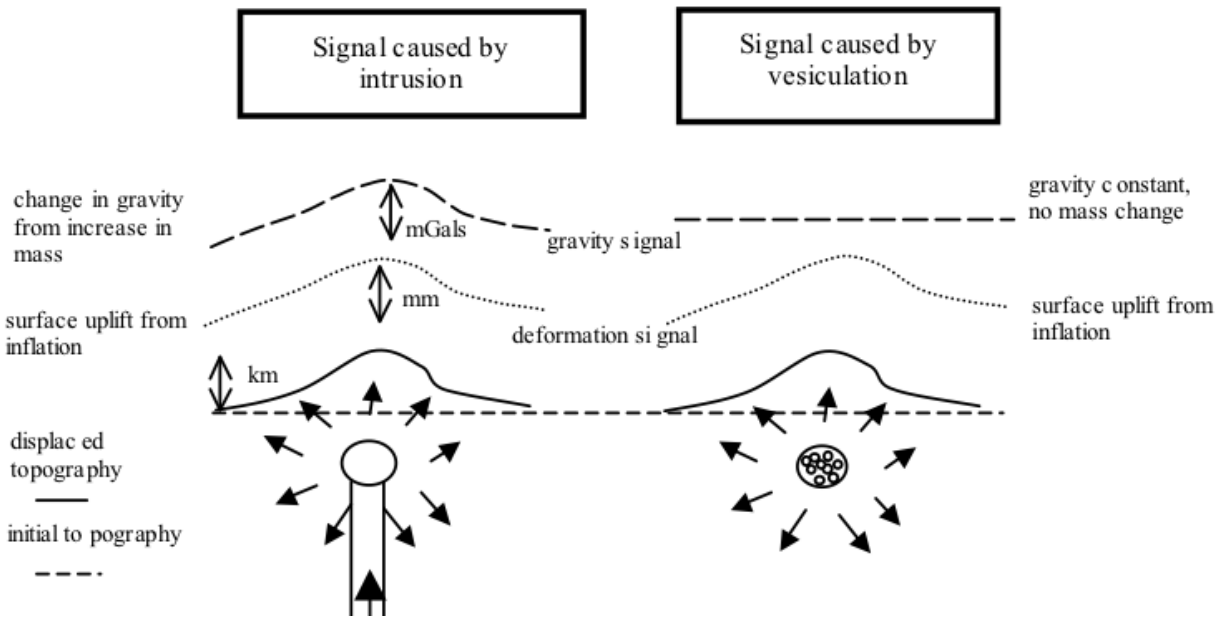


Figure 2. On the left, an intrusion causes increases in gravity and uplift in deformation across the source. On the right, a vesiculating source causes uplift across the source, but no change in gravity.

From a monitoring standpoint, changes in gravity are sometimes detectable long before other more traditional precursors. During a volcanic crisis on the island of Fogo, a subsurface accumulation of magma was detected through campaign gravity surveys months before changes in deformation or seismic activities were detected (Fonseca et al., 2003). Gravity data can also help to clarify the eruption mechanism. During a flank eruption on Mt. Etna, a continuous gravity meter recorded a sharp decrease in gravity, followed by a sudden but more gradual increase in gravity. With this continuous gravity data, seismic data, and petrological evidence, scientists were able to deduce that the mechanism for this flank eruption was a dry fault opening in the volcanic edifice, which then created a path for magma to erupt at the surface (Carbone et al., 2007). Studies at Campi Flegrei, Mayasa, and Merapi have also used gravity to monitor volcanic activity (Jousset et al., 2000; Gottsmann et al., 2003; Camacho et al., 2007; Battaglia et al., 2008; Williams-Jones et al., 2003, 2008). At Kilauea, work by Daniel Johnson suggested that long-term subsidence at the summit may be due to deflation of magma bodies through gas release. Gravity measurements would be crucial to distinguish this activity from the rapid deflation associated with eruptions on the rift zones (Johnson, 1992).

My primary objectives for this study are to assess the capabilities of the current network geometry and suggest improvements where necessary. Kilauea's complex volcanic terrain limits station location, so it is important to understand the capabilities of a gravity network on the summit, given the limitations on station location and density. I will determine network capabilities by examining the minimum depth and volume combinations necessary for the

network to detect a source. I will also examine the relative importance of the different stations in the network. Since network geometry can distort a spatial signal, I will also look at this problem for Kilauea's network. Using what I know about network capabilities, I will suggest improvements as needed.

Methods

Gravity monitoring at Kilauea currently consists of two continuous gravity meters and a network of campaign gravity benchmarks that are co-located with leveling benchmarks for height control. These campaign gravity benchmarks are typically located along paved roads for easy access and surveying, with exceptions including a small loop of benchmarks in the caldera northeast of Halema`uma`u crater and three stations in the south caldera region (Figure 3).

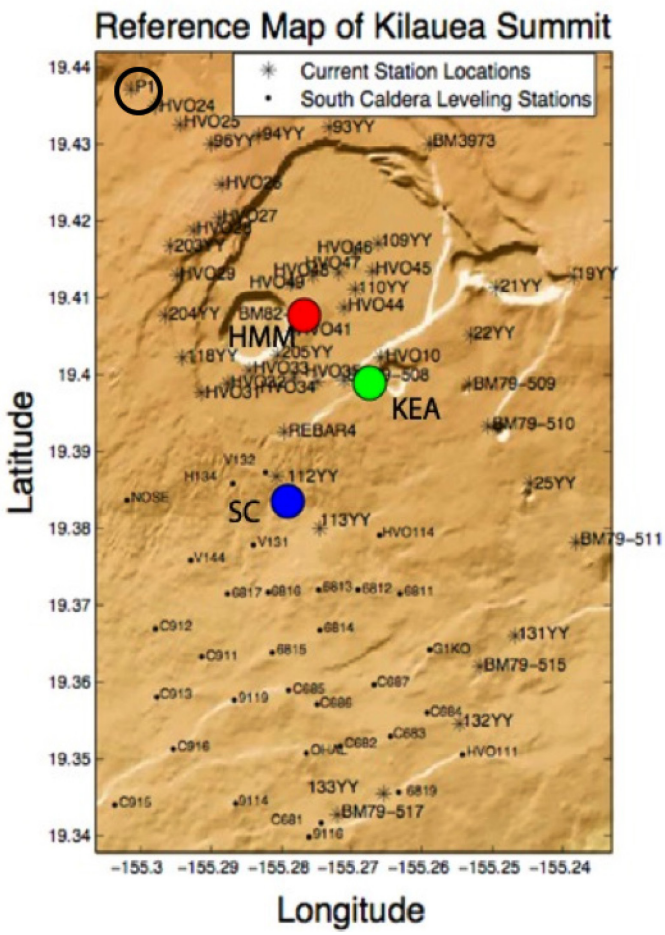


Figure 3. Reference map of gravity and deformation stations in the summit area. Current gravity stations are marked with stars, and deformation stations in the south caldera region are marked with dots. The three colored dots represent source locations for this study. P1, the base station, is circled in black.

In an actual campaign gravity survey, gravity values for all stations are calculated relative to P1, a benchmark presumed to be stable relative to the volcano. These benchmarks are surveyed using the double-loop method: first the base station (P1) is surveyed, followed by a few stations,

the base station again, the same stations as before, and finally the base station a third time to end the day's survey. Gravity meters drift over time, such that a reading at a stable benchmark early in the day will be lower than a reading at the same benchmark later in the day. This effect increases with time. Surveying the base station three times constrains the drift of the instrument as well as permitting an error calculation for each station. Each survey day is as short as possible, as long survey days mean larger drift on the meter. For the campaign gravity surveys, HVO uses a Scintrex CG5 relative gravity meter, which (under Kilauea field conditions) has an error of approximately 15μGal (1 Gal is equal to 1 cm/s^2).

To test the capabilities of the current network of campaign gravity benchmarks, I forward modeled a series of very simple scenarios. The simplest possible model for magma accumulation and withdrawal uses an analytical point source. An analytical model relies on simple equations and is useful because it is computationally simpler and thus requires minimal computing power and time. Although a point source does not accurately represent volcanic activity at Kilauea, it is a computationally simple model that actually fits real data quite well in many cases (Battaglia et al., 2008). My model uses equations for a point source of expansion and contraction in an elastic half-space adapted from Mogi (Dzurisin, 2007). I assumed a density of 2800 kg/m^3 for basaltic magma, and a density of 3100 kg/m^3 for basaltic crust. The total gravity change is calculated as follows (Figure 4):

$$\Delta g = \frac{G\Delta V d}{\left(r^2 + d^2\right)^{3/2}}\left[\left(\rho_m - \rho_c\right) + 2\rho_c\left(1-\nu\right) + \rho_c\left(1-2\nu\right)\right]$$

Figure 4. The analytical equations used in my model, adapted from a basic Mogi model.

where Δg is gravity change, G is the gravitational constant, ΔV is the change in volume of the source, d is the depth of the source, r is the radial distance to the source, $\boldsymbol{\rho}_m$ is the density of magma, $\boldsymbol{\rho}_c$ is the density of the crust, and ν is Poisson's ratio (2.5). The first term in the brackets calculates gravity change due to the increase in mass. The second term calculates gravity change due to uplift of the inflating point source. The third term calculates gravity change due to the elastic expansion of the crust changing the density of the crust (Dzurisin, 2007).

A cross section of gravity change across a Mogi source shows a steady increase in gravity as horizontal distance to the source approaches zero (Figure 5). Gravity reaches a peak above the source location (Figure 5). As the source gets more shallow, the peak becomes sharper, and as the source becomes larger in volume the whole curve moves up.

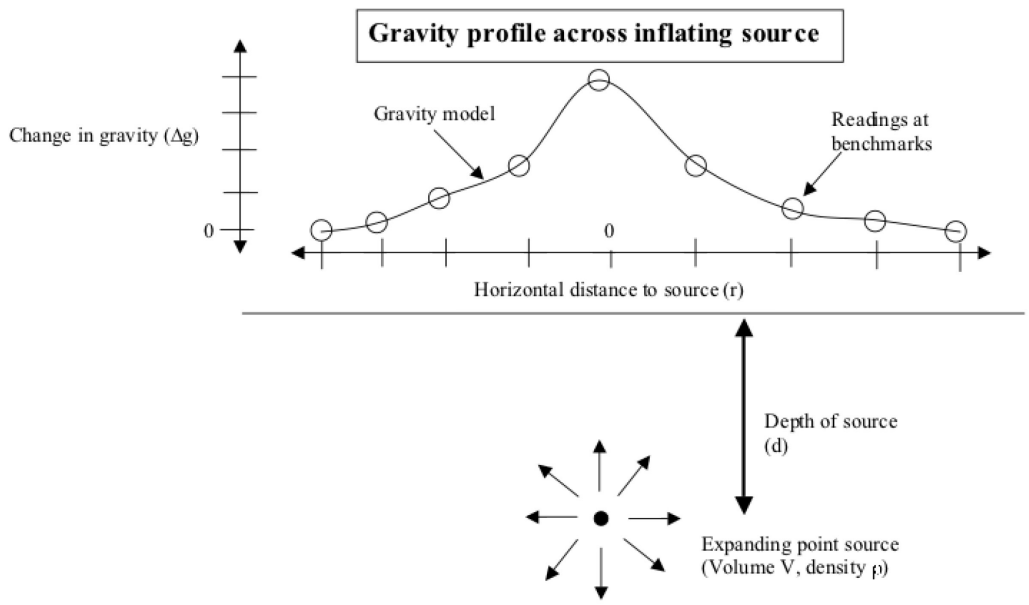


Figure 5. Change in gravity versus distance from source for an expanding point source (Mogi model). The curve peaks smoothly above the source, and the sharpness of the peak increases as the source depth decreases.

In map view, the signal from a Mogi source is radially symmetric about the source location (Figure 6).

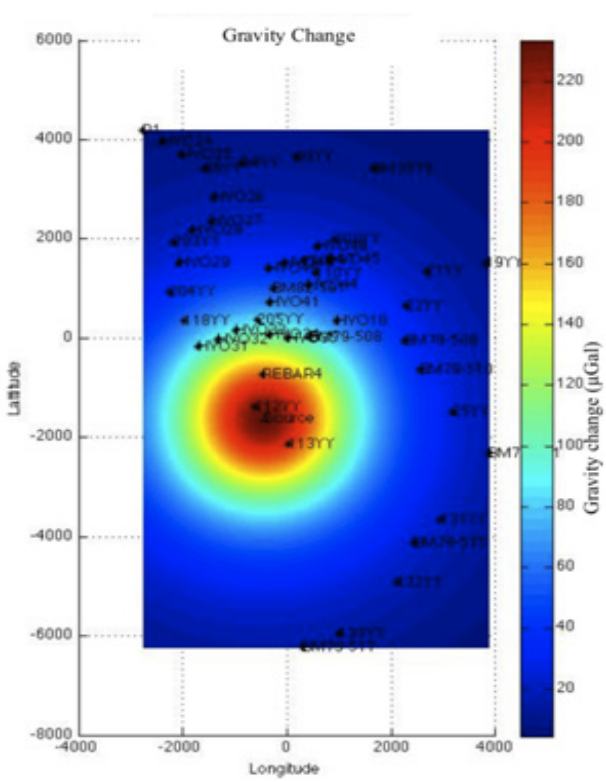


Figure 6. Map view of an expanding Mogi source. Note how gravity change is greatest over the source and decreases uniformly with distance from the source.

For this study, I focused on three source areas: Halema`uma`u crater (HMM), Keanakakoi crater (KEA), and the South Caldera (SC) region (near station 112YY), all known areas of magma storage (Decker, 1987; Johnson, 1987; Johnson, 1992; Johnson et al., 2010) (Figure 3). Most recently there has been inflation and deflation at HMM, and deflation at SC (Johnson et al., 2010). Inflation and deflation episodes have also occurred at KEA in the past few decades of observation (Decker, 1987). Whether or not the gravity change at the stations in the existing network is above the noise level of the gravimeter can be a way to assess network capabilities quantitatively.

Under Kilauea field conditions, the noise level of the Scintrex CG5 spring gravimeters used at HVO is about 15 μGal. If the model calculates a gravity change at a station that is above 15 μGal, this measurement is considered reliable and useful. If it is below 15 μGal, it is not, because the measurement cannot be distinguished from the noise of the meter. The number of stations above this noise level for a given depth and volume of source provides a measure of how well the network detects that source. Accordingly, to assess the effectiveness of the network, I ran a series of inflation simulations over a range of depth (0 to 10 km) and volume (5.5×10^5 to 10^9 m^3) combinations. For each volume/depth pair, I kept track of how many stations were above the noise level for each simulation and of how many times each individual station was above the noise level over all the simulations. To quantify station value, I established a ranking system based on how many times each individual station measured above the noise level. I ran different depth/volume simulations for each of the three sources, and then added these results to get a cumulative result in addition to the results from each individual source. For each of these four data sets (three single sources and one cumulative), I identified the highest score (number of times a station measured above the noise level), and called this the "best possible value" (BPV). I then arbitrarily decided that stations that scored over 80% of the BPV in that data set were important stations, and those that scored below 80% were less important. This arbitrary cutoff was based on map-view contours of station scores. Each map had 5 to 6 contours spaced evenly between the lowest and highest score. As the third contour (roughly half way between highest and lowest score) usually fell close to 80% of the BPV, I decided this would be an acceptable cutoff point.

To gain some understanding of how the network geometry distorts the spatial gravity signal, I compared gravity changes contoured from a hypothetical grid of points to gravity contoured from data at the station sites alone. In an ideal situation, a gravity survey would use a grid pattern of stations to minimize signal distortion, but rough volcanic terrain and considerations of survey time makes this virtually impossible at Kilauea. As Figure 7a illustrates, gaps in coverage can distort the spatial gravity signal. The top figure shows even coverage for a given source. The inflating source causes a given gravity change at each station, and when these gravity changes are contoured they create a smooth Mogi curve that peaks over the source. The bottom figure (7b) shows that when some of these stations are removed, even though the gravity change at the remaining stations is still the same, the gaps in coverage alter the contour of the data. Now instead of a high peak over the source, the apparent peak is shifted to the left and reduced in magnitude.

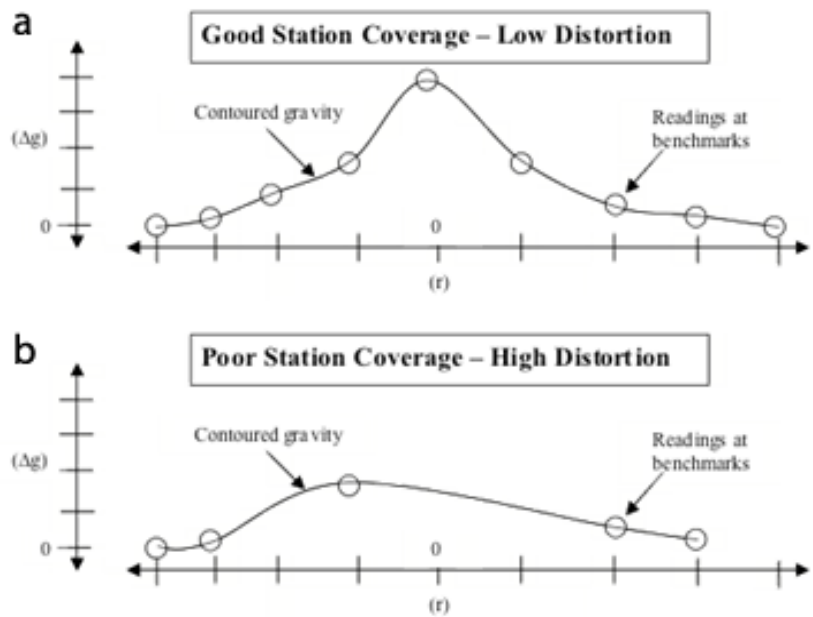


Figure 7. Schematic showing distortion of inferred gravity signal from poor station coverage. In the top figure, even station coverage allows contouring of gravity data to produce a smooth curve and provide a reliable indicator of source location (at r=0). In the lower figure, poor station coverage distorts the gravity signal when data is contoured, skewing the indication of source location.

To address this problem at Kilauea, I calculated the gravity change first for a hypothetical 500 by 500 grid of points and contoured the data to produce an ideal scenario. I also calculated gravity change at only the existing station locations, and contoured those data, using Matlab's "griddata" function and the default "v4" option. I then subtracted the grid-calculated map from the station-calculated map to obtain a map of the distortion, and calculated a numerical residual value between the two maps for a qualitative image of signal distortion. This method provides for a visual means of evaluating signal distortion, but lacks the quantitative results of an actual inversion. However, there are data in the calculated residual, so I deemed this method sufficient for this preliminary assessment of network capabilities.

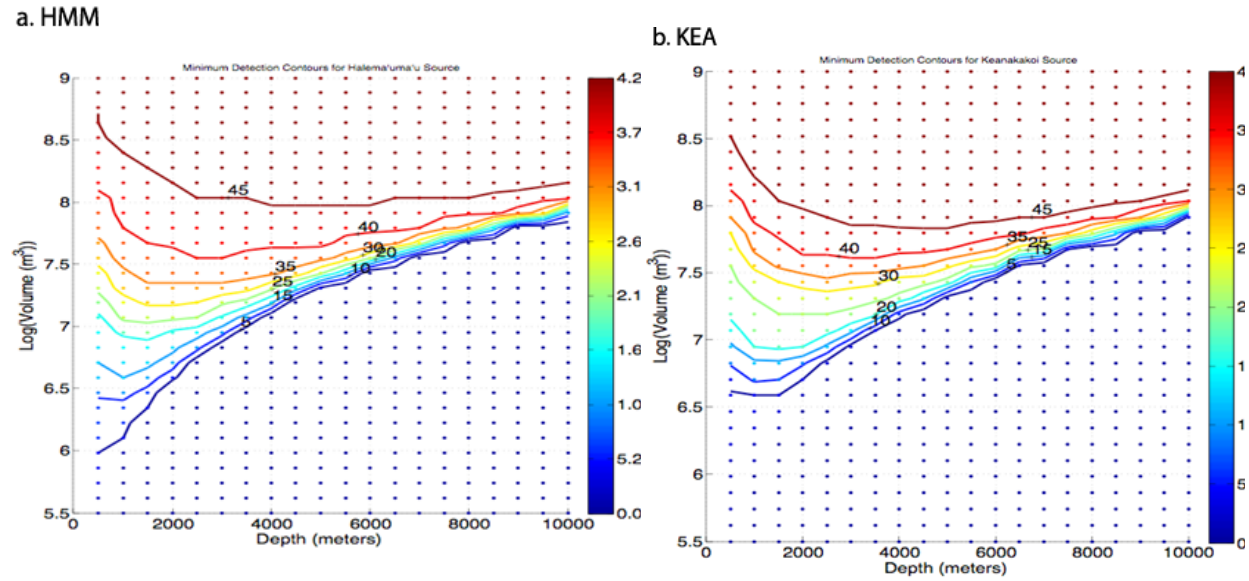
Results

One method for assessing the network's ability to detect different sources is to keep track of how many stations are above the noise level for a range of different source depths and volumes. In Figure 8, I have plotted the number of stations above the noise level for intrusion simulations at each of the three sources. Each colored dot represents a scenario.

Source depth is plotted along the x axis, and source volume is plotted along the y axis with a base 10 logarithmic scale. The color of each dot corresponds to the number of stations above the noise level for that depth and volume combination. These dots are contoured for number of stations. The lowest line, which is dark blue, represents 5 stations above the noise level, and the highest, dark red, represents 45 stations above the noise level (there are 47 total stations in the current network). For all figures, the contours are parallel except at shallow depths, where they

diverge slightly because of station density close to the source. Figure 8a depicts detection threshold contours for a source at HMM, 8b at KEA, and 8c at SC.

From Figure 8a we can see that the network begins to detect a source at HMM at the 5 station level when it reaches just below $1x10^6$ m^3 in volume, and for 20 stations to detect a source at HMM, the source must be at least $6x10^6$ m^3. Sources at KEA are detectable by 5 stations when they reach $1.7x10^6$ m^3. Sources are detectable by 20 stations when they reach $6x10^6$ m^3, as at HMM (Figure 8b). Figures 8a and 8b show that the network is, for the most part, equally sensitive to sources at HMM and KEA.



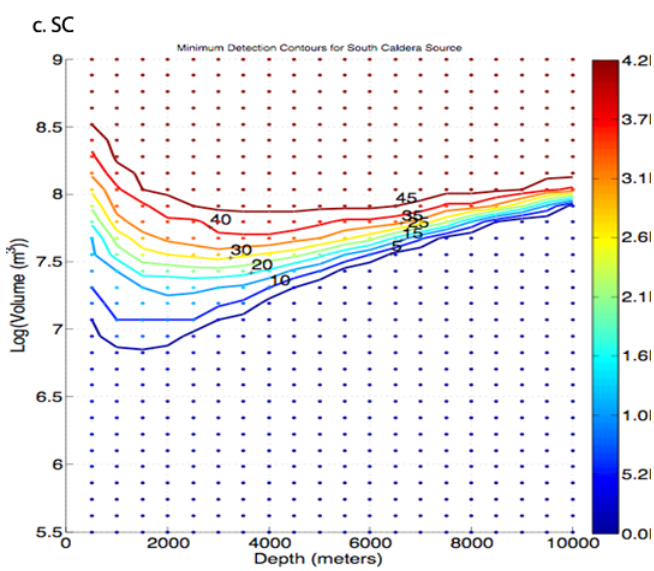


Figure 8(a-c). Minimum detection thresholds of gravity changes for a source at HMM, KEA, and SC. Each colored dot represents an intrusion scenario with a given depth (x axis) and volume (y axis). The color of the dot represents how many stations were above the noise level (15μGal) for that scenario. This data is contoured to show patterns more clearly. The lowest line represents 5 stations above the noise level, and the highest line represents 45 stations above the noise level.

Figure 8c shows that the network is much less sensitive to the SC source. A source of $6x10^6$ m^3 is necessary for a minimum of 5 stations to be above the noise level, whereas a source of this volume would be detectable by a minimum of 20 stations for a source at HMM or KEA (Figure

8a-b). For 20 stations in the network to detect a source at SC, the source needs to be at least 2.5×10^7 m^3.

Figure 8 shows that for all sources examined, the network will detect a source with a volume greater than 3×10^8 m^3 for all depths tested (this volume is slightly higher at HMM, at 5×10^8 m^3). For comparison purposes, the May 1973 east rift zone eruption produced 1.2×10^6 m^3 of lava, the 1977 east rift zone eruption produced 3.5×10^7 m^3, the July 1974 summit eruption produced 6.5×10^6 m^3, and the 1969 Mauna Ulu eruption produced 3.4×10^8 m^3 ("Global Volcanism Program," May 2011).

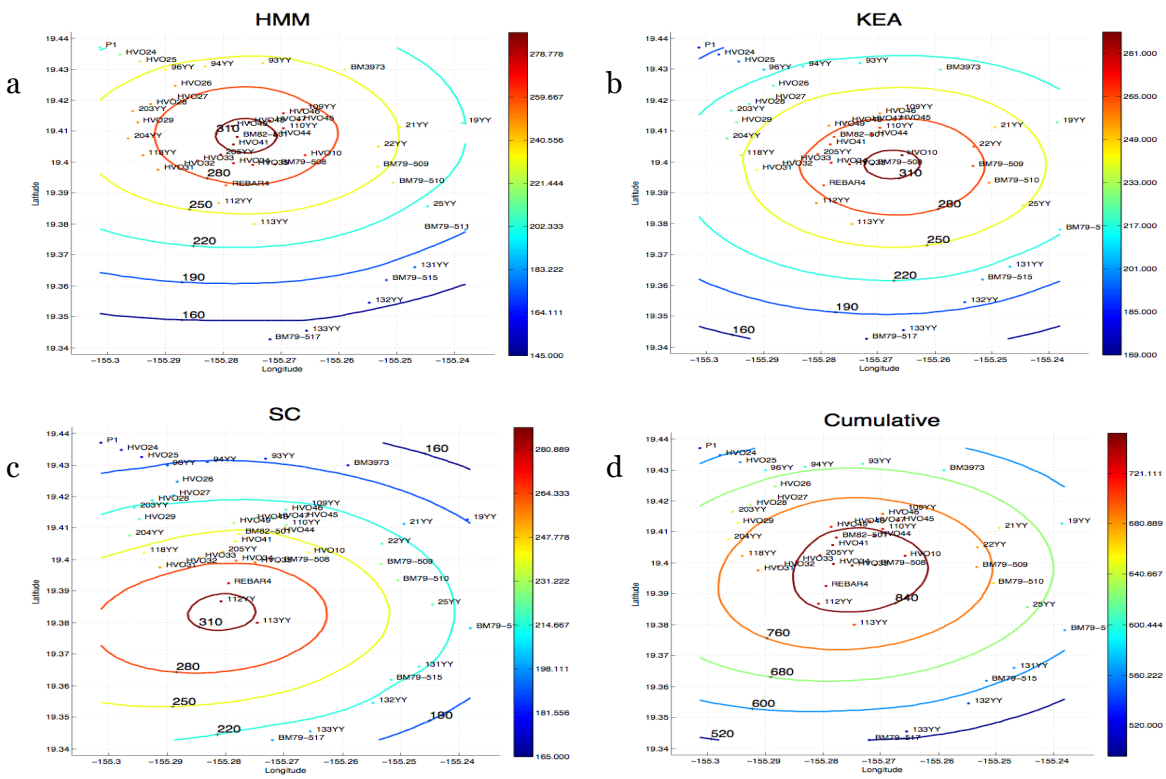


Figure 9(a-d). Relative station importance for detecting a source at HMM (a), KEA (b), and SC (d). Figure 9d is relative station importance using all three sources. Stations are plotted in their relative positions using latitude and longitude, and station color corresponds to the number of times that an individual station was above the noise level across all the scenarios (intrusions of different depths and volumes) calculated. Stations that fall within the third contour (80% BPV) are considered important to the network, and those that fall outside the third contour are less important to the network.

Figure 9a-d shows the relative importance of stations for a source at HMM, KEA, SC, and all sources combined, respectively. Each station is represented by a colored dot that represents relative station importance, with red colors representing stations with higher scores (meaning the station was above the noise level for more scenarios) and blue colors representing stations with lower scores. These dots are then contoured to show the spatial distribution of station importance. For a source at HMM, KEA, SC, and all sources cumulatively, the best possible values (BPVs) are 317, 313, 314, and 882, respectively, out of 600 simulations for the individual

sources, and 1800 simulations for the cumulative score (600 simulations for each of the three sources). The BPV for all sources is higher because this cumulative data set is the sum of the scores from the three individual sources.

Using the arbitrary definition that stations that score above approximately 80% of BPV are important and those that score less than 80% are less important, stations falling within the third highest contour in Figures 9 a-d are important (77-79% BPV), and those falling outside the third highest contour are less important. Figure 9a shows that the stations that are important for detecting a source at HMM are essentially inside the caldera, encompassing 32 of 47 total stations. Results for a source at KEA are similar (Figure 9b), with more stations to the east and south of the main caldera falling within the third contour and 26 stations total. Figure 9c shows that only 14 stations are important for detecting a source at SC, as only 14 stations fall within the third contour, and 12 of these are between the second and third contours. Looking at the cumulative map (Figure 9d), we see that stations in the caldera and directly outside of it are most useful, while stations along the Mauna Loa road (P1 to 96YY), the Hilina Pali road (BM79-511 to BM79-517), and Crater Rim Drive (94YY to 19YY) are less useful.

In summary, the network resolves sources at HMM and KEA fairly well, with little to no distortion of the signal. The source, however, shows considerable distortion. Figure 10 shows maps of residual for HMM (10a), KEA (10b), and SC (10c), al lcalculated for a source $5x10^7$ m^3 in volume and 2 km deep.

Stations are plotted in map view, and color SC corresponds to the residual between the grid calculated gravity map (the ideal network) and the station calculated map (the actual network). Reds are positive residual, blues are negative residual (the large negative residuals that occur on the edge of

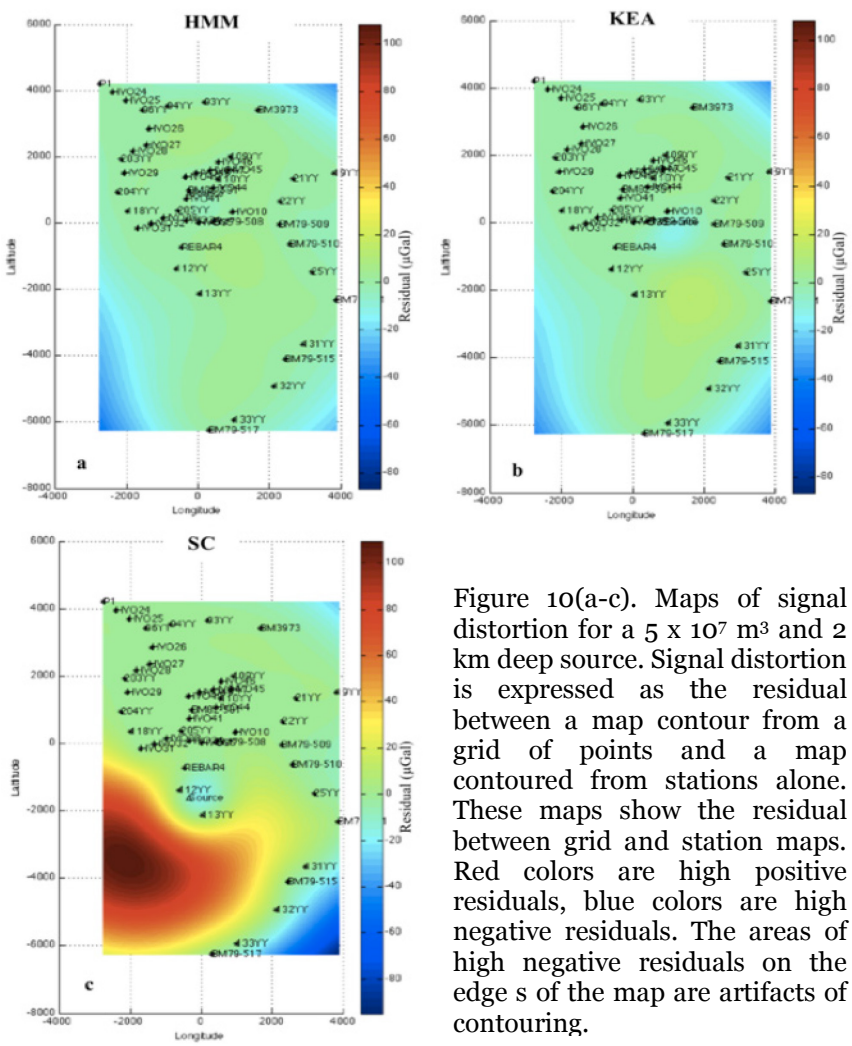


Figure 10(a-c). Maps of signal distortion for a 5 x 10^7 m^3 and 2 km deep source. Signal distortion is expressed as the residual between a map contour from a grid of points and a map contoured from stations alone. These maps show the residual between grid and station maps. Red colors are high positive residuals, blue colors are high negative residuals. The areas of high negative residuals on the edge s of the map are artifacts of contouring.

the map are an artifact of contouring). The maximum residual for the SC source is near 100 µGal (Figure 10c), whereas the residual for HMM and KEA is below 20 µGal (Figure 10a-b). These residuals increase as the source volume increases and/or depth decreases.

Figure 11 shows grid and station calculated maps for a HMM source of $5x10^7$ m^3 and 1 km depth (Figure 11a-b), and grid and station calculated maps for an SC source of $5x10^7$ m^3 and 2 km depth (Figure 11c-d). A shallower source at HMM (1 km) has a higher residual, but the spatial signal is not noticeably distorted between the grid map and the station map (Figure 11 a-b). In contrast, the spatial gravity signal from a source at SC is noticeably distorted between the grid map and the station map (Figure 11 c-d), spreading the circular Mogi signal in an oval that encompasses part of the southwest rift zone, an area of frequent volcanic activity.

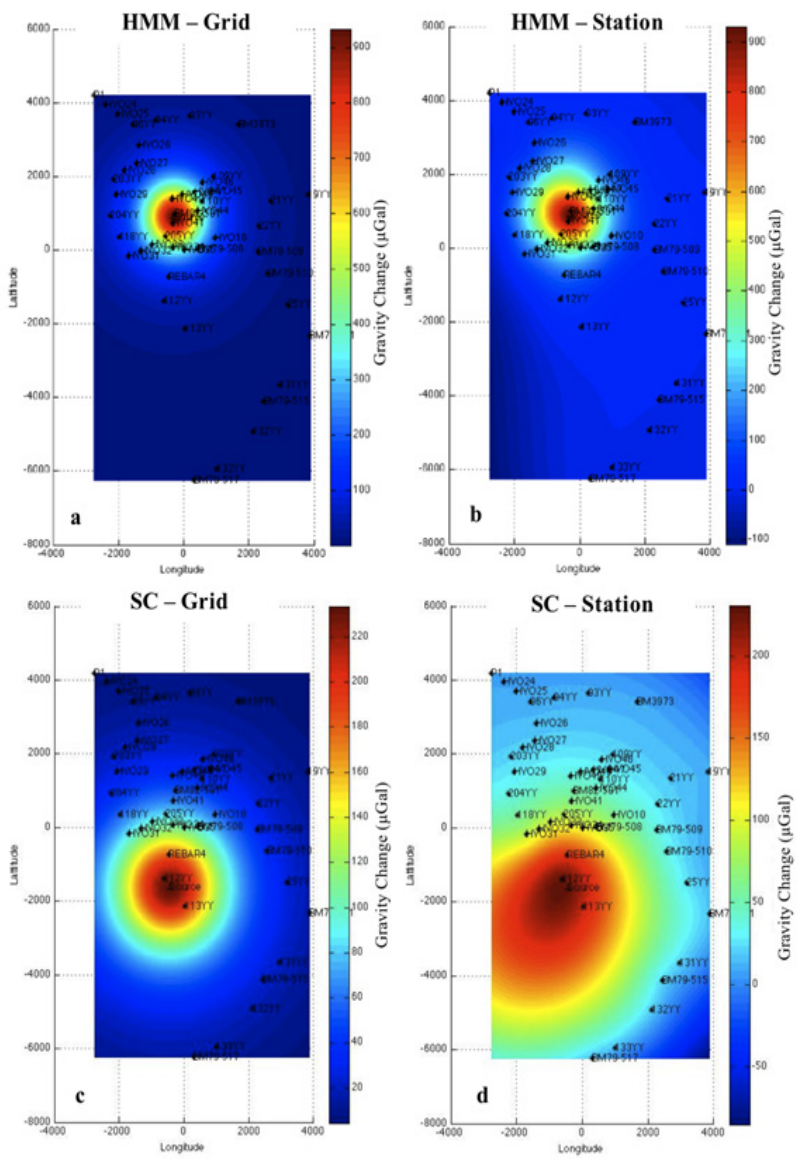


Figure 11 (a-d). Comparison of spatial gravity signal distortion between a source at HMM and a source at SC. Figures 11a and 11c show the ideal signal expected from a Mogi source. The ideal signal is created by calculating gravity at a hypothetical grid of points across the network area, and the n contouring the data produced by the grid of points. Figures 11b and 11d are produced by contouring only gravity changes calcuated at current sites, showing how the network does not appreciably distort the spatial gravity signal for a source at HMM, but causes significant distortion of a signal for source SC.

Discussion

The current network is able to detect most large intrusion volumes (>$2x10^7$ m^3) at depths less than 5 km, although smaller or deeper intrusions are less likely to be detected, especially at the SC source. For each source, detection criteria contour lines begin converging at deeper depths (Figure 8). This is because as the source gets deeper, the spatial gravity signal becomes longer wavelength, and the detection criteria saturates. This means that for each source, it will be difficult to establish a definite source location until the source is shallower than 8 km, because differences in gravity change between stations that would show source location may be obscured by the noise level of the instrument.

Conversely, at shallow depths these detection criteria contour lines are more widely spaced along the volume axis, because at shallower depths the signal is much more spatially localized (Figure 8), such that it takes larger volumes to make the small wavelength spatial gravity signal detectable at more stations. The spacing of the contours at shallow depths is a measure of station density around each source. For the HMM source, this spacing is fairly regular, indicating that stations are present at relatively uniform densities at all distances from the source (Figure 8a). By contrast, for the SC source, the contours for 5, 10, and 15 stations are very widely spaced and the higher contours are more closely spaced (Figure 8c). This reflects very sparse station coverage near the source, and denser coverage farther away from it. This means that a source must be larger than $1.7x10^7$ m^3 before it can be detected beyond the few stations close to the SC source. The KEA source is somewhere in between HMM and SC; its lower contour lines are densely spaced and its higher contour lines are more widely spaced (Figure 8b). However, the closely spaced lower contour lines ensure that smaller sources can be reliably detected by numerous stations.

Most of the detection level contours reach a local minimum at some depth and volume combination, meaning that for a given volume, the same level of detection (say, 20 stations) is possible for two different depths. This is because of the change in the shape of the Mogi model signal with depth: for a constant volume, at shallow depths the signal peaks sharply, and at greater depths, the signal peak spreads out. This means that at certain detection level (the contours on the figure), for a given volume there is an optimum depth for detection. For example, an HMM source of 10^7 m^3 will be detected by 25 stations at 1.5 km depth, but will only be detected by 20 stations at 0.75 km depth.

The results from the station value maps are encouraging. For all stations evaluated cumulatively, 33 of 47 total stations are important to the network, scoring above 680 (which is 77% of BPV) (Figure 9d). The stations that fall outside of 81% BPV are typically the less easily accessible stations (the stations along Mauna Loa Road and the Hilina Pali Road, and 19YY) or less reliable stations (93YY, 94YY, and BM3973). Interestingly, P1, the base station, does not have the lowest score, which suggests that for very large or very shallow events P1 may not be an ideal base station (figure 9d). Stability of the benchmark and accessibility are other considerations for choosing base stations. Conversely, the stations that score lower than P1 may not require frequent surveying. Continuous gravimeters could be placed in areas of greatest station value for each of these three possible sources to maximize effectiveness.

My analysis also suggests that it might be useful to prioritize surveys for different coverage areas and different time intervals. It takes about 1.5 weeks to survey the entire current network, but it would only take 1-2 days to survey all the stations inside the first contour on the cumulative station value map (Figure 9d). Surveying all the stations within the third contour in Figure 9d would probably take 5-6 days. A potential survey schedule could be a yearly survey of the entire network, a survey of third contour stations every 3 to 4 months, and a survey of the first contour stations every 1 to 2 months.

The station value maps for each of the three sources (Figure 9 a-c) suggests that the network is best configured for sources at HMM and KEA. These maps also suggest that the network is not as well set up for a source at SC, a fact that is troubling since this area has shown marked deflation and gravity decreases in recent leveling and gravity surveys (Johnson et. al. 2010). Only 14 stations are important for this source, as compared to HMM and KEA, which have 32 and 26 important stations, respectively.

Once I had determined through forward modeling that the south caldera region showed the greatest degree of bias from the current station network, I used forward modeling to determine possible new station locations and assess their potential for improving the residual error. I began with a map of current leveling stations superimposed on the map of existing stations, and chose new trial station locations based on which of these leveling stations I knew to be accessible by road (or nearly so). Figure 12 shows the trial station locations I chose.

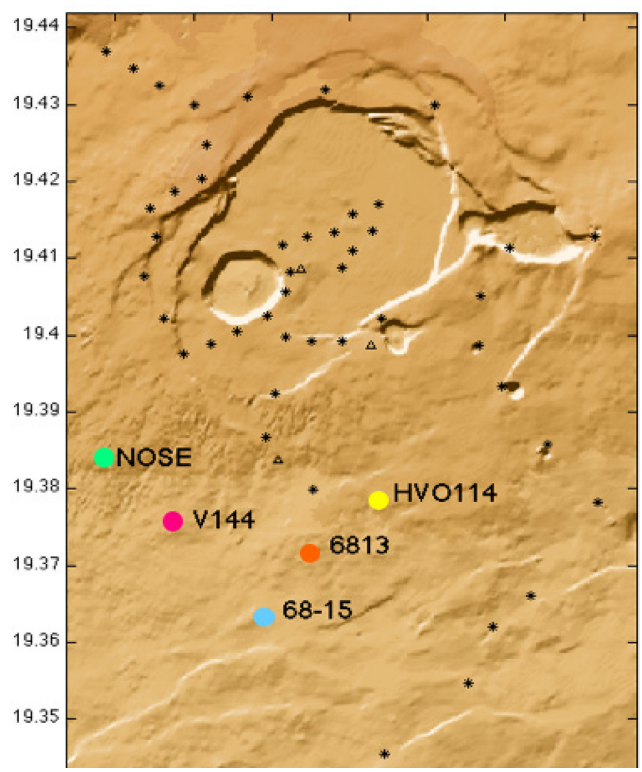


Figure 12. Proposed locations for new gravity stations. These sites are co-located with existing leveling benchmarks (see figure 3). Stations in the existing network are marked with stars, and source locations are marked with triangles. (x-axis is latitude, y-axis is longitude).

I then optimized station order by first trying only one station, then seeing which station of the five possibilities gave the greatest improvement in residual. Then, using the best station of those five as the first station, I would then try two stations, and find which of the remaining four gave the greatest improvement, and use these two best stations when trying three stations, and so on, until I had determined the best order for station addition. Figure 13 shows several alternate station orders, plotting percent residual improvement as a function of number of stations added. Table 1 shows the order in which stations were added for station orders A through F. Figure 13 shows that adding stations using order A (V144, 6813, HVO114, NOSE, 68-15) gives residual improvement the fastest with each station added.

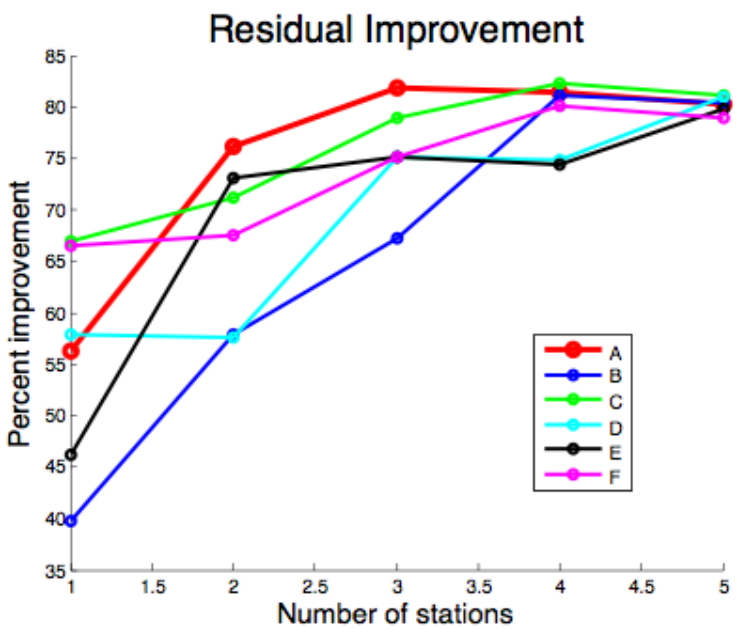


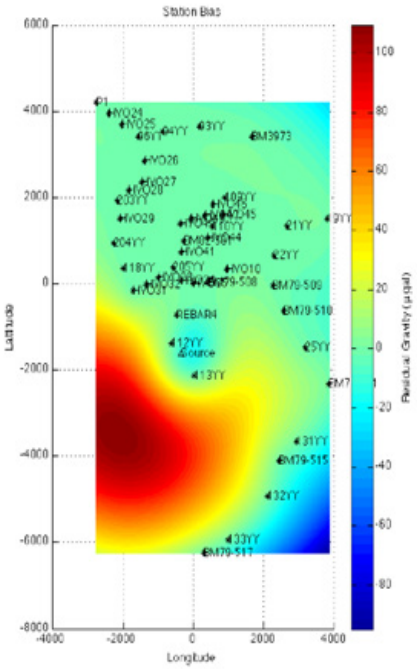
Figure 13. Plot of percent residual improvement versus number of stations added for a variety of station orders. Although some orders (C,D, and F) have a greater improvement than order A initially, these other orders do not improve as fast as A after this first point. Order A shows the greatest improvement fastest, and is therefore the optimal order for adding new stations to the network.

Table 1. Table showing the order of stations in each order in Figure 13

Station Orders

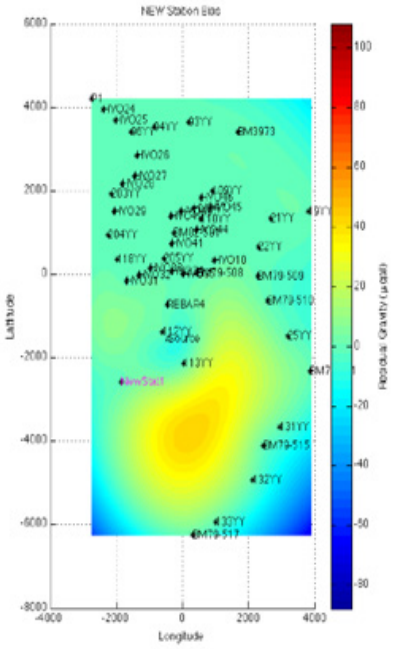
Order name	1st station	2nd station	3rd station	4th station	5th station
A	V144	68-13	HVO144	NOSE	68-15
B	NOSE	V144	HVO114	68-13	68-15
C	68-15	NOSE	HVO114	68-13	V144
D	V144	NOSE	68-13	68-15	HVO114
E	68-13	NOSE	V144	68-15	HVO114
F	68-15	V144	68-13	HVO114	NOSE

a



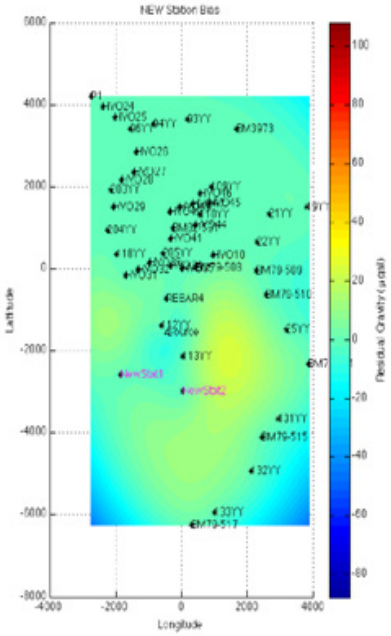
0 new stations

b



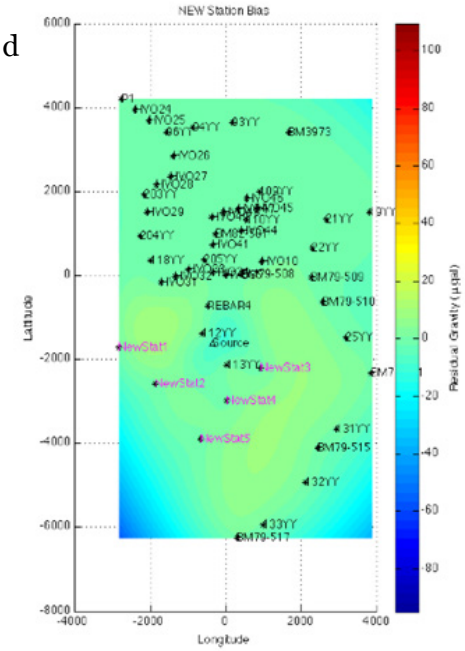
1 new station

c



2 new stations

d



3 new stations

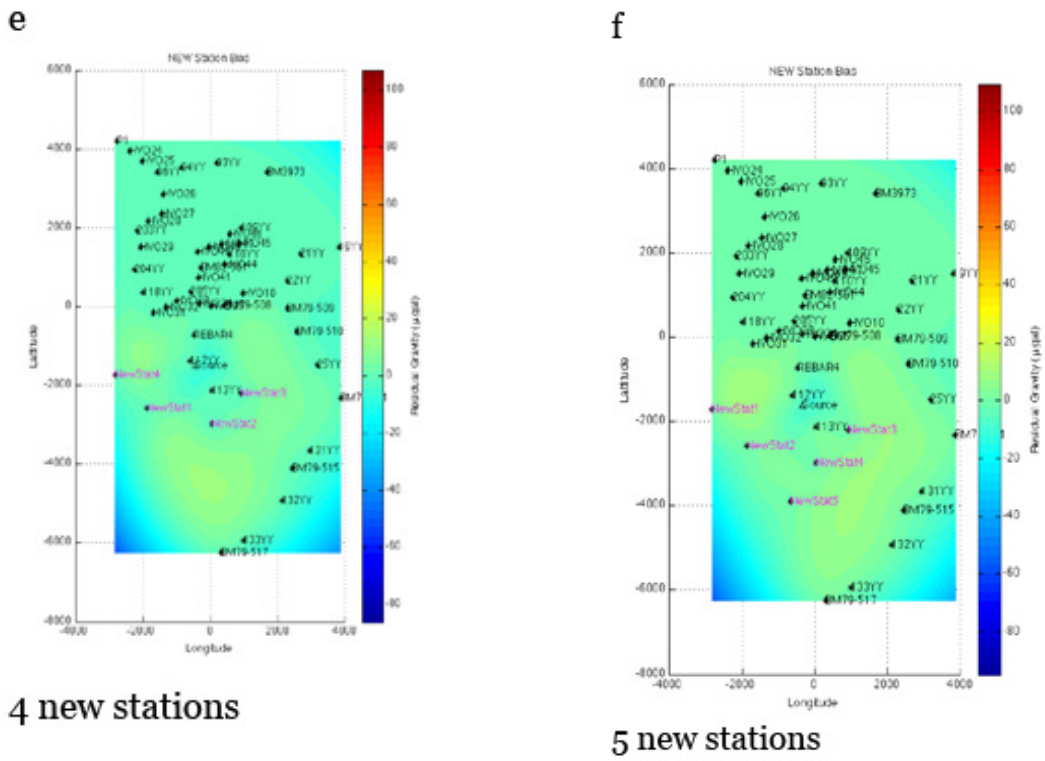


Figure 14(a-f): Maps of improving residual with the addition of new stations in the south caldera area. These maps represent the difference between a map of gravity change contoured using a grid of points and a gravity map calculated using data points only at station locations. High positive residual is in red, and high negative residual is in blue (the high negative residual on the edges of the maps is an artifact of the contouring). Figures a-d show dramatic improvement, but figures d-f show little change.

In general, there is dramatic improvement in the residual for the first three stations added, but after this the reductions in the residual seem to level off or decrease slightly (Figure 14). Figure 14 shows maps of residual after 0-5 stations have been added. The difference between Figure 14a (zero new stations), and Figure 14d (three new stations) is dramatic – with no new stations, residual reaches nearly 100 μGal, but with three additional stations in the south caldera area, maximum residuals drop to below noise levels (below 15μGal). However, there is little difference between Figures 14d-f (four and five new stations), where maximum residuals remain at around the same level. Future work should include a more comprehensive set of possible sources. These sources could include locations on the east and southwest rift zones, and also northeast of Halema`uma`u crater in the caldera. It would also be very useful to forward model from existing real data to see how the current network and any proposed alterations to the current network would detect gravity change in more realistic situations. Using inverse methods would better quantify network distortion and improvements to the distortion through the addition of new stations. Also important would be additional forward modeling to determine how well the network could detect vesiculation of existing intrusions.

Conclusions

Using campaign gravity surveys to track subsurface changes in mass is a useful technique for monitoring Kilauea's constant activity. Since eruptions at Kilauea are often preceded by intrusions of magma at the summit, gravity surveys can effectively track this subsurface activity long before surface activity commences, providing longer warning time for purposes of hazard mitigation (Decker, 1987; Johnson, 1987). Campaign gravity surveys can provide good spatial coverage, and additional continuous meters, carefully placed, can provide good temporal coverage (Battaglia et al., 2008).

However, this monitoring technique is useful as long as the data it provides is of good quality. Forward modeling magma intrusions using a simple point source model is an effective way to understand the capabilities and limitations of the campaign gravity network on Kilauea. Although a vast simplification of a much more complicated reality, these models give the researcher a first look at what the network can detect, which stations are most important to the network, and where the network needs additional station coverage.

The current gravity network on Kilauea's summit is well suited for detecting sources at HMM or KEA, and by extension any sources within the caldera. However, with the current network configuration, sources in the SC area are much harder to detect and locate. Installing new stations in this area would greatly alleviate this problem; in fact, only three new stations may be necessary. To simplify future gravity surveys, some stations from the current network (HVO24–96YY, 94YY–19YY, and BM79-511–BM79-517) could be surveyed less frequently, as this forward modeling shows that they are of lesser importance to the network. Conversely, some stations (mostly those in the caldera along the Chain of Craters Road) could be surveyed with greater frequency based on their importance to detecting likely sources.

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Acquisition of Second Language Vocabulary for Kindergartners with Speech Sound Disorders

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Abstract

Researchers and educators alike have raised concerns over the potential exclusion of children with speech and language disorders in immersion programs. The purpose of this pilot study is to determine whether children with speech sound disorders acquire Spanish vocabulary at a rate comparable to typical peers when learning in an immersion program, and if rates of acquisition differ for expressive versus receptive vocabulary. Five kindergartners attending a partial, early elementary immersion school were studied: one control participant and four participants with speech sound disorders. This investigation used expressive (spoken) and receptive (understood) vocabulary probes to test the number of Spanish vocabulary words children could produce and comprehend within an eight-week period. Results show that children with speech sound disorders acquire expressive Spanish vocabulary at a rate comparable to their typical peers but had lower levels of acquisition overall, while rates of receptive vocabulary acquisition varied across participants. The results of this pilot study, which are not yet comprehensive, suggest that children with speech sound disorders are able to acquire Spanish vocabulary and, as a result, should continue to be included in immersion programs in the future.

Introduction

The American Speech-Language-Hearing Association (ASHA) website states that speech and language disorders (S/LD) are among the most common disabilities in the United States ("Incidence and Prevalence," 2008). The number of children who received services for speech or language disorders in fall 2003 was estimated at 1,460,583, not including children who had speech or language disorders secondary to another disorder (U.S. Dept. of Education, 2005); while the total number of children enrolled in elementary school in the United States in 2009 was 32,238,000 (U.S. Census Bureau, 2009). Given that close to 4.5% of children in elementary schools received services for speech and language disorders, it is clear that many children in the United States (US) are affected by these disorders such that their speech production, speech use, academic success, and even social participation may be disrupted (McCauley, 2004).

Second language learning represents one facet of academic success in the US educational system, which is a prerequisite for high school students applying to college, and is generally a requirement for university Bachelor of Arts students. Although the majority of children and young adults learning a second language participate in traditional foreign language education (such as foreign language in elementary school programming), an alternative method called

immersion schooling has proven to be a successful method of educating children in a second language (Lambert & Tucker, 1972).

Immersion schools have frequently been shown to benefit children from the regular educational track (e.g., Campbell, Gray, Rhodes & Snow, 1985; Lambert & Tucker, 1972; Turnbull, Lapkin & Hart, 2001), yet the benefits of immersion schooling for children with S/LD, who could be considered atypical language learners, remain largely unstudied. Genesee (1987), an advocate for inclusion of children with S/LD in immersion schools, states, "If immersion is not good for all, it should not be offered to only a few; therefore, it becomes available to none" (Genesee, 1987). Consequently, it is important for children with S/LD, who make up a large number of the children in our schools, to have equal access to immersion schooling. Currently, it is not entirely clear what the benefits of immersion schooling are for such children, or the potential difficulties.

This pilot study aims to assess Spanish vocabulary acquisition over an eight-week period by English-speaking children with speech sound disorders (SSD) attending an immersion school. Although the scope of this study is not large enough to show definitively the success of children with SSD in immersion schooling, understanding their ability to acquire vocabulary is an important step toward addressing their academic success—or difficulties—in learning a second language. Sinatra (2008) states that understanding of vocabulary is strongly related to language comprehension and reading tasks and, ultimately, academic success. Therefore, assessments of vocabulary acquisition of students learning a second language are crucial to understanding whether these students will be academically successful in learning a second language. The aim of this pilot study is to analyze second language vocabulary acquisition for children with SSD in comparison to typical peers. The research questions that direct this investigation include:

1. Do kindergarten children with SSD acquire Spanish vocabulary at a rate similar to typical peers when learning in an immersion program?
2. If children with SSD acquire Spanish vocabulary through an immersion program, do their rates of acquisition for receptive (that which is spoken) versus expressive (that which is audibly understood) vocabulary differ?

Speech and Language Disorders in Children

Children with speech and language disorders (S/LD) are a heterogeneous group of children who experience difficulties with speech production and/or language comprehension (Dodd & McIntosh, 2008). Speech is "the production…of the sounds that convey phonetic structure" while motor movement of body parts such as the tongue and mouth allow production of the elemental sounds, or phonemes, that make up the speech signal (Liberman & Whalen, 2000). Language is a culturally shared code that allows any group of people to communicate with one another utilizing higher-level cognitive processing (Liberman & Whalen, 2000). In order for both speaker and listener to recognize certain sounds as language, both must have a cognitive understanding of the language being spoken; their brains are trained to understand a speech sound of their cultural code, or language, over a sound that is not a part of their code (Liberman & Whalen, 2000). Children with S/LD experience disruptions to speech, language, or both with varying degrees of severity and differences in etiology (cause), recovery, and social effects

(Leonard, 1998; Webster & Shevell, 2004; McCauley, 2004). Due to the integral role of speech and language in educational tasks, S/LD affect scholastic success as well as communication and socialization (Varley, 2008).

For a typically developing child, acquisition of speech and language occurs easily and without extensive scaffolding from parents. Young children still make some mistakes such as incorrect use of tense markers and pronunciation of some late developing sounds (such as /l/, /r/, /s/, /sh/, /ch/, /y/, /v/, /z/, and /th/) but the majority of their speech can be well understood ("National Institute," 2000). By age four or five, 80% to 90% of speech by a typically developing child can be understood (Hamaguchi, 2010). For children with S/LD, the process of acquiring and using speech and language is a challenge that requires aid beyond that of parents and classroom teachers. These children must meet with a speech-language pathologist (SLP) to work on tasks such as picture naming drills, joint book reading, dramatic play, computerized language comprehension exercises, and writing assignments (Leonard, 1998; Webster & Shevell, 2004). These tasks encourage further interaction with language and speech than the child receives in their regular classroom experience (McGregor, n.d.).

Speech Sound Disorders in Children

Speech Sound Disorders (SSD) are some of the most common disorders of speech among school-aged children (Lewis et al., 2006), with an estimated prevalence of 2-13% (Peterson & McGrath, 2009). SSD encompass difficulties with phonetics (articulatory components) or phonemics (cognitive-linguistic processing) and affects populations with structural, perceptual, or neuromotor disorders (Bowen, 2009). There are two main subgroups within SSD that describe a child's SSD more specifically: articulation disorders and phonological disorders. Although current research combines these two terms into SSD, the names 'articulation' and 'phonological' are still used as lay terms to describe these disorders. Changes in terminology are common in the field of communication disorders and sciences due to new discoveries in research (Bowen, 1998). In general, an articulation disorder is associated with difficulties at a phonetic level (motor production of speech sounds), while a phonological disorder implies difficulties at the phonemic-linguistic (cognitive) level, which indicates trouble in organizing and understanding the speech sound system (Bowen, 2009; McCauley, 2004). Children with phonological disorders will typically experience difficulty with word production or will have an inadequate understanding of the phonological system that makes up speech, causing speech comprehension problems for the child. Children will sometimes produce sounds correctly in certain contexts and incorrectly in others. Often, they will shorten their sentences to minimize confusion experienced by the listener, and substitute or omit sounds in their speech production, affecting intelligibility (Dockrell & Messer, 1999; Peterson & McGrath, 2009). When a child has a purely articulatory problem, the cause is typically an oral-motor disorder, which manifests as incorrect production of sounds (Dockrell & Messer, 1999; Peterson & McGrath, 2009).

The etiology of SSD and other speech disorders is unknown for many individuals; however, there are some hypotheses to account for these differences. For example, researchers have recently identified a link between genetics and SSD, as many children diagnosed with SSD have SSD in their family history (Lewis et al., 2006; Dockrell & Messer, 1999). Other potential causes

of SSD include differences in IQ, socioeconomic status, speech development, motor skills, auditory discrimination, oral structures, hearing loss, and neuropathologies (Hegde, 1995). Yet most children with SSD have no known cause for what is considered their "functional speech disorder" (Bowen, 2009).

Impacts of SSD include problems with speech production and use, reduced intelligibility, risk for broader S/LD, academic difficulties, and social stigma (McCauley, 2004). SSD causes stress for the child struggling through speech production, as well as for parents and siblings who must listen closely to decipher and understand the child's speech (Bowen, 2009). Aside from these difficulties, studies have linked SSD with lower academic success. Felsenfeld, Broen, and McGue (1994) studied the long-term effects of a moderate phonological disorder in children whose disorders persisted past first grade, and compared them against a cohort of students who had typical articulation abilities in childhood. During the participant identification process, children who performed at least 1.5 standard deviations below the mean on the Prekindergarten Imitation Articulation Test—which assesses production of word-initial and word-final consonant singletons, as well as word-initial consonant clusters—were included in the phonological disorder cohort. Control participants had to score at or above the mean and have received no speech or language treatment. After 28 years, the investigators interviewed these students, recording their educational and occupational accomplishments. It was found that participants who had a moderate phonological disorder received lower grades in high school, required more academic help, and completed fewer years of formal education. In addition, this group tended to occupy jobs considered semiskilled or unskilled at a higher frequency than the control group. Although these results do not predict that any single child with a phonological disorder will be less successful in adulthood, Fesenfeld et al. (1994) conclude that a phonological disorder may be one facet in understanding how childhood speech problems might manifest in adulthood.

Immersion Schooling

Immersion schooling began in Quebec, Canada in 1965 at the St. Lambert French immersion program (Genesee, Paradis & Crago, 2004), an educational initiative created to address the prevalence of both French and English language use in Canada. The first immersion program in the United States opened in 1971 as a Spanish total-immersion elementary school in Culver City, California (Genesee, 1987), created to address the increase of Spanish language use in the United States. Spanish is currently the most popular target language in U.S. immersion schools, although interest in other target languages is increasing (Christian, 1996). At immersion schools, children learn a majority or all of their school subjects in a second language, leading to the assumption that children will learn the second language much in the same way they learned their first language (Cummins & Swain, 1986; Genesee et al., 2004). Teachers do not introduce the second language as a separate subject, but use it as a "vehicle" for instruction (Stewart, 2005). Participation in immersion schooling is usually voluntary, mostly based on parental decision (Christian, 1996). There are various models of immersion programs, including early, middle, or late immersion, as well as partial and total immersion. Early immersion typically begins in kindergarten or grade 1, middle immersion in grades 4 or 5, and late immersion in grades 6, 7, or 8 (Lazaruk, 2007). For partial immersion, students learn about 50% of the day in

English, and 50% in the target second language. Total immersion schools typically begin with 100% of instruction in the target second language, and gradually add classes taught in English over the course of following years (Genesee et al., 2004).

Immersion schooling differs from traditional foreign language classes in instructional methods. Students in immersion programs learn a second language through the instruction of other subjects given in that language, with little direct language instruction. Students in foreign language in elementary schools (FLES) programs learn through traditional foreign language instruction. As a result, students in immersion programs spend less time learning grammatical rules and vocabulary through rote memorization, yet are proven to have a superior understanding of their target second languages (Cummins & Swain, 1986).

Children with Speech and Language Disorders in Immersion Schools

Studies conducted over the past 50 years have shown the overwhelming benefits of immersion schooling for the education of mainstream children in both native and second languages (Campbell, Gray, Rhodes & Snow, 1985; Lambert & Tucker, 1972; Turnbull, Lapkin & Hart, 2001). By contrast, the suitability of immersion schooling for children with S/LD is less widely researched. Genesee, a leading advocate for the advantages of immersion schooling for all students, states, "It is imperative that educational decisions concerning exclusion of subgroups of students from immersion be founded on systematic objective investigation, and not on speculation or 'common sense'" (Genesee, 1987). Children with S/LD have never been systematically excluded from immersion schooling, though concerns of parents and educators have limited their enrollment in immersion schools, since parents and educators worry that having regular academic subjects taught through another language may be an unnecessary burden for these children (Genesee, 1987). Cummins and Swain (1986) also explain that "A commonly held view is that immersion education is only for children of above average intelligence" (Cummins & Swain, 1986) because of the assumed difficulty of learning all school subjects in a second language. Research has proven, however, that this system does not burden children academically, and actually helps them achieve at high levels in all school subjects (Campbell et al., 1985; Lambert & Tucker, 1972; Turnbull et al., 2001).

Bruck (1978) has conducted the most prominent research concerning language-impaired children in immersion settings. She conducted her research in Montreal, Quebec, and studied children with language difficulties in Canada's French immersion schools. Bruck's study included four groups of native English speaking children: (a) children with language impairments in French immersion classes, (b) children with language impairments in English-only classes, (c) typically functioning children in French immersion classes, and (d) typically functioning children in English-only classes. For this study, language screenings performed by the school identified those children with language impairments. The investigator studied these four groups from kindergarten to grade 3 by testing them on second language abilities, cognitive development, and school achievement. The results of this study revealed that children with language impairments were not only successful in their immersion program, but also outperformed children with language impairments in English-only settings. They did not succeed in academic achievement and second language acquisition as rapidly as those students

without language impairments (both in the immersion setting and not), but this finding was expected given their pre-existing condition. Bruck concluded that because the language-impaired children were successful in an immersion school, they should have the opportunity to learn in an immersion setting and not be excluded based on speculation. Bruck (1982), in a later commentary, notes that children with language impairments had higher levels of second language proficiency than non-immersion students, both with and without language impairments, in conventional second language courses. Additionally, she states that language-impaired children cannot typically cope with traditional second language courses, as learning linguistic rules and structures by rote directly exposes their weaknesses.

Not all studies on children with language difficulties in immersion schools have yielded positive results. Trites (1981) examined the suitability of immersion schooling for children with a maturational lag of the temporal lobe, a region of the brain typically associated with language function. The participants in this study were students who possessed high IQs as well as good motor and sensory function, and yet performed poorly on the Tactual Performance Test (TPT), a psychomotor problem-solving test in which children place blocks into a foam board while blindfolded, alternating the blocks using the dominant hand, non-dominant hand, and then both hands. Trites concluded that poor performances on the TPT is a result of a maturational lag in the temporal lobe of the brain, and that students who performed poorly were also those students having trouble succeeding in French immersion schools. This conclusion counters the belief that all children are best suited to learn new languages at an earlier age: these deficits in second language ability imply that if these children had been taught in their native language, such problems would not have occurred. Trites repeated the study a year later with different students, and reached the same conclusion: despite normal intelligence and good socioeconomic background, some students are not fit for early immersion schools. Follow-up was done with students who had transferred out of immersion schooling, and it was determined that they were academically successful in an English-only educational setting. However, Genesee (1987) criticized Trites' study because it only dealt with English-speaking students having difficulty in French immersion schools, and did not find minority language students with the same profile having trouble learning English. Genesee (1987) pointed out that Trites' study suggested that this disability is specific to English-speaking children in French immersion schools, which seemed highly unlikely. For these results to be applicable to a greater population of students learning second languages, a study that yielded similar results and tested children learning a second language other than French would be necessary. As it is, Trites' conclusions (1981) were based on a narrow study sample and cannot be applied to a wide variety of students.

Summary

Research on immersion schooling for children with S/LD has generally implied that students with S/LD are successful in learning a second language through immersion schooling, and should therefore have the opportunity to learn in such settings. Immersion schooling has proven academic benefits for most children, so children with S/LD should be considered for any academic advantage possible, since S/LD negatively affect academic achievement. It is necessary for researchers to conduct more studies concerning the success of children with S/LD in

immersion schools for consideration by parents and educators and to broaden the body of knowledge of this subject for speech language pathologists.

Methods

Participants

All participants in the study were kindergartners attending a partial-early immersion elementary school. All children were native English speakers learning Spanish as a second language. The investigator selected participants based on: (a) enrollment at the test school, (b) willingness of child and parents to participate, and (c) having English as a native language. The investigator recruited children with SSD based on their eligibility for speech therapy with a licensed speech-language pathologist (SLP) at the school and nomination of probable candidates by the SLP. The investigator called parents of children for recruitment, and parents completed a permission form drafted for the purpose of the study. The Institutional Review Board (IRB) of the University of Oregon approved all consent documents and study procedures.

Summarized participant information is found in Table 1. The participants in the study were all five to six years of age and had speech and/or language disorder diagnoses with the exception of a control participant, designated AB. AB was a five-year-old male with no speech or language difficulty. DC was a five-year-old female diagnosed with a SSD and a unilateral hearing impairment. DC started kindergarten with errors in the production of the following sounds – /s/, /z/, /s/ blends, /f/, /v/, /sh/, /ch/, /j/, /th/, final /t/, and /d/. At the time of the study, she could produce most of these sounds correctly in conversation. DC struggled with peer social interaction that affected her learning. EF was a six-year-old female diagnosed with a SSD. She made errors in the following sounds – /k/, /g/, /f/, /l/, /th/, /r/ blends, and /s/ blends, and had more difficulties with multisyllabic words. GH was a five-year-old male diagnosed with a SSD and a language disorder. The following sound errors characterized his articulation – /m/, /r/, /v/, /l/, /th/ and /s/ blends. His language scores were at the seventh percentile with his receptive scores being much higher than his expressive scores. He used simplified sentences with many pronoun and verb tense errors. JK was a six-year-old male diagnosed with a SSD. At the beginning of kindergarten, his language scores were at the seventh percentile and he made errors in the following sounds – /k/, /v/, /ng/, /y/, /th/, /ch/, /z/, /s/, /r/, and /l/ blends. His phonological patterns showed consonant deletion (e.g. “no” for “nose”), consonant cluster reduction (e.g. “net” for “nest” and “tee” for “tree”), and occasionally consonant stopping (e.g., “tun” for “sun”). At the time of the study, the only sound he was still receiving therapy for was the /th/ sound.

“Class” listed below in Table 1 signifies whether children were in the same Spanish class or not; students in Class A learned together and students in Class B learned together. All four students in Classes A and B had the same Spanish teacher but had Spanish class at opposite times of day. Likewise, students in Class A and B had English with the same teacher but at opposite times of day (this schedule maintained a 50:50 ratio of Spanish to English instruction time). The student in Class C had different Spanish and English teachers and classrooms.

All kindergarten Spanish teachers used the same Spanish class curriculum. The school sought to maintain an equal number of native Spanish and English speakers enrolled in each classroom to reinforce peer language assistance. Kindergartners learned Spanish and English literacy (in each respective language) at all times during the school year, and mathematics, science, and writing were taught alternately in English and Spanish, switching between the two languages every five units.

Table 1. Participant Profiles.

Child	**Age (years; months)**	**Sex**	**Diagnosis**	**Class**
AB	5;11	Male	No SSD	Class A
	5;11	Female	SSD	Class A
			DC	Unilateral Hearing Impairment
EF	6;3	Female	SSD	Class B
	5;9	Male	SSD	Class B
			GH	Language Disorder
JK	6;0	Male	SSD	Class C

In terms of the population at the study school, a total of 332 students were enrolled in grades K-5 at the time of the study. Among the students at the school, 80.7% were considered economically disadvantaged, as indicated by the percent of students receiving free or reduced price lunches (Council of Chief State School Officers: School Matters, 2011). Demographic statistics from the school website show that 53.9% of the students identified as White, 1.8% as Black, 33.7% as Hispanic, 0.9% as Asian/Pacific Islander, and 2.1% as American Indian/Alaska Native ("Council," 2011).

Procedures

Measures

Measures for the study included a list of 25 Spanish vocabulary words chosen from the students' Spanish class curriculum. All vocabulary words from the first three units of kindergarten Spanish curriculum (those in the vocabulary word banks) were written down, and the words used more frequently in the classroom (determined through observation by the investigator) were chosen for the final list of 25 words. The investigator also chose words based on the criteria that they are high frequency (commonly used), concrete nouns, and fit into one of two categories: items found in school and body parts. See Appendix A for a complete list of words. The investigator created picture flashcards with clip art images (see Appendix B) of each word used in the three vocabulary probes (vocabulary testing sets).

Data Collection

Participants completed four biweekly testing sessions over an eight-week period, with each session lasting about 15 minutes and taking place in the hallway or child's classroom in the school. At the first testing session only, the investigator administered an English expressive vocabulary probe as a data collection exercise and at all four sessions, two vocabulary probes were administered to measure receptive and expressive Spanish vocabulary acquisition. These two probes are the dependent variables of the study, and represent Spanish language expressive and receptive abilities of the children. An investigator, conversationally proficient in Spanish, administered the probes.

English vocabulary probe

The investigator administered the English vocabulary probe at the first testing session only, to determine if each child knew the vocabulary words in English. In order to be considered for continuation in the study, all students needed to know each Spanish vocabulary word in English; students were not expected to learn words in Spanish without previous knowledge of the English equivalent. For this probe, children were shown one picture flashcard at a time and asked to say the word in English. The investigator repeated this exercise for each word.

Expressive vocabulary probe

The investigator used the expressive vocabulary probe at each testing session to test the child's expressive, or spoken, abilities in Spanish. Children were shown one picture flashcard at a time, and asked to say the word in Spanish. The investigator repeated this exercise for each word.

Receptive vocabulary probe

The investigator administered the receptive vocabulary probe in a second exercise at each testing session, in order to test the child's receptive knowledge, or auditory understanding, of Spanish. The investigator would set out five picture flashcards at a time, stating a word in Spanish and asking the child to point to the picture of that word. The investigator repeated this exercise five times until all 25 words had been shown.

Scoring

For data calculation, the number of words correctly produced (expressive) and understood (receptive) by the child was divided by the total number of correct responses possible, 25. This created a percentage accurate for each probe (excluding the English vocabulary probe) that exemplifies the percentage of expressive or receptive vocabulary words the child appeared to know. The investigator considered expressive answers correct when the child produced a word with the correct syllabic structure; specifically, with the correct number of syllables, correctly stressed syllables, and with CV (consonant-vowel) structure intact. Words were sufficiently near target and considered correct when their production met the above criteria and no more than one phonemic insertion or substitution of either a vowel or consonant was made. The investigator considered receptive vocabulary correct when the child pointed clearly at a flashcard in response to the stated word.

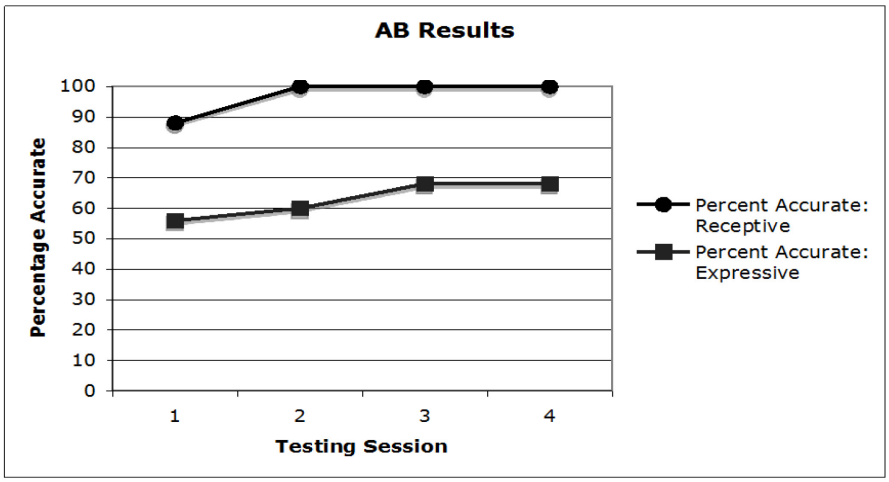
At one testing session, a reliability coder (a secondary person tasked with scoring the expressive and receptive abilities of the children alongside the primary investigator) calculated point-by-point agreement for expressive and receptive vocabulary accuracy. The reliability coder, together with the investigator, scored whether a child expressed or demonstrated comprehension of words accurately, following the same inclusion criteria. The reliability coder and investigator compared scores given independently to each child during each probe, determining if both felt the production reflected that the child had learned the word. The reliability coder was a native Spanish speaker and certified SLP employed at the study school. The reliability coder participated at only one testing session to verify that the inclusion criteria employed by the investigator was an accurate representation of the children's vocabulary understanding. If the same inclusion criteria were followed in a subsequent study, baring human error, the same inter-rater reliability percentage would most likely be achieved. The investigator calculated inter-rater reliability for this pilot study at 100%.

Results

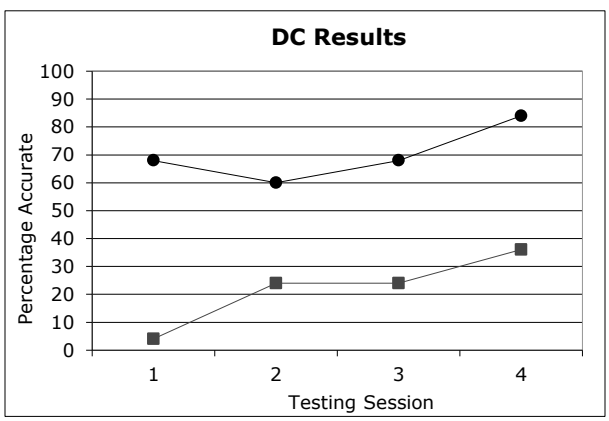
Data from the probe sessions shows the percentage of vocabulary words (out of 25) each child could accurately express and understand receptively per testing session. Scores on the two dependent variables are calculated as percentages by dividing the number of words correctly expressed or receptively understood by the total number of words possible, and then multiplying by 100. Each figure lists percentage of words accurately expressed or receptively understood per probe session per child. As Figure 1 shows, control participant AB increased 12% (from 56% to 68% accuracy) on the expressive vocabulary probe. On the receptive probe, he also increased 12% (88% to 100%). Figure 2 shows that DC increased 32% on the expressive vocabulary probe (4% to 36%). On the receptive probe, she increased 16% (68% to 84%). According to Figure 3, EF increased 8% on the expressive probe (36% to 44%), and increased 4% on the receptive vocabulary probe (96% to 100%). As Figure 4 illustrates, GH increased 8% on the expressive vocabulary probe (0% to 8%), and decreased 8% on the receptive vocabulary probe (68% to 60%). The investigator should note that data for GH was used from his second and final session instead of first and final session due to his absence from school on the first testing day. Figure 5 shows that JK increased 8% on the expressive vocabulary probe (12% to 20%), and increased 0% on the receptive probe (60% to 60%). Overall, the participants with SSD increased in the number of words produced accurately on the expressive vocabulary probe, an average of 14% from the first to the final testing session. On the receptive vocabulary probe, all children but one increased in their knowledge of receptive vocabulary, an average of 3% from the first to the final testing session (or 6.67% excluding the child who decreased from the first to the final session).

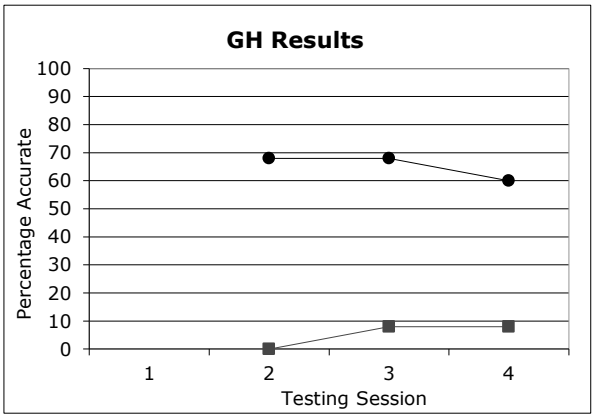
The investigator calculated a paired *t*-test, which assesses whether the means of two groups of data are statistically different from one another, to determine whether the number of words learned from the first to the final session represented a significant increase for the children with SSD. The results were not significant for the expressive vocabulary ($t(3)=2.33$; $p=.10$) or receptive vocabulary ($t(3)=.60$; $p=.59$) scores.

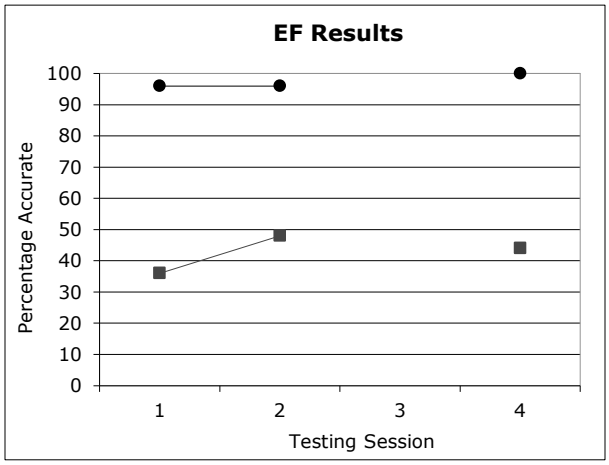
Control participant:



Children with speech sound disorders:







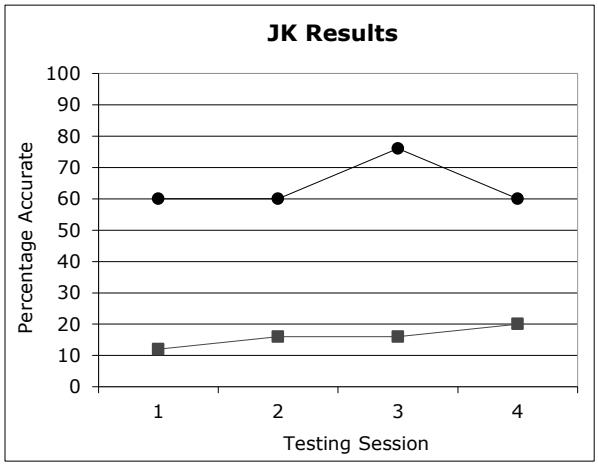


Figure 1. Results of Expressive and Receptive Vocabulary Accuracy for Control Participant and Children with Speech Sound Disorders.

Discussion

This study investigates whether children with SSD acquire Spanish vocabulary similarly to typical peers over an eight-week period. The first research question addresses whether kindergartners with SSD would acquire Spanish vocabulary at a rate comparable to a typical peer. Acquisition rate was determined by comparing change in acquisition accuracy from session one to the final session of the typical peer against the average change in acquisition from session one to the final session for the students with SSD. Results showed that participants learned expressive vocabulary at a similar rate (14% compared to 12%), but did not learn receptive vocabulary at a similar rate (3% compared to 12%). Interestingly, this comparable rate of expressive vocabulary acquisition occurred despite the fact that the control participant began the study demonstrating a higher number of expressive vocabulary words. The children with SSD began at a lower level of expressive ability and increased from that level, at a similar rate to the control participant.

The second research question addressed the differences between expressive and receptive vocabulary acquisition for participants with SSD, and in comparison to their typical peer. All children, including the control participant, showed superior accuracy in their receptive vocabulary comprehension over their expressive vocabulary abilities throughout the eight-week study. The investigator expected these results, given that most children learning their first language comprehend more words than they initially express (Hamaguchi, 2010).

Most of the participants with SSD began the study below the control participant in expressive and receptive abilities (with the exception of EF) yet all but one child increased over time in both the number of words acquired expressively and receptively. Interestingly, although there was a difference between the percent of expressive and receptive vocabulary acquired by all children, the difference between expressive and receptive vocabulary acquisition for those with SSD was even greater than that of the control participant (a difference of 32% between expressive and receptive vocabulary at the final probe session for the control participant compared to 40-56% for the children with SSD). Most children, when learning a language, experience a phenomenon called the "silent period," in which they focus on listening and comprehension and speak very little in order to put their efforts in first understanding the language (Roseberry-McKibbin & Brice, n.d.). It appears that the children in this study excelled in their receptive abilities over their expressive abilities, which conforms to this 'silent period' phenomenon.

The higher receptive and expressive abilities of the control participant from the onset of the study, as well as greater gains by some participants with SSD, are open to interpretation. Factors that may have impacted overall acquisition include: (a) exposure to Spanish outside of the classroom, (b) the possible link between children who were late talkers at age two and those diagnosed with S/LD during school-age years, (c) vocabulary word preference and (d) degree of vocabulary exposure within each classroom. It is important to note that the investigator administered the initial study probe at a time when children had had no previous exposure to Spanish. Children were four months into their Spanish education at the time of the study, so the rates and impact of previous learning is unknown. Furthermore, although all children were

selected based on their limited exposure to Spanish (all kindergartners beginning Spanish education for the first time), those children whose parents were willing to practice Spanish with them outside of the classroom and encouraged their use of Spanish would potentially be more capable of understanding and using Spanish vocabulary, although it is not possible to measure this factor within the scope of the study. Along with parental help, it is possible that the control participant's overall higher performance was due to experience previous to the study that the other children did not have; an example would be watching popular television programs like *Dora the Explorer*, which contains Spanish vocabulary words. Alternatively, there may have been a link between rates of acquisition and their SSD. Lower rates of acquisition experienced by these participants could be the result of a possible, as-yet unidentified, language difficulty. For example, late talkers, or children diagnosed at a young age with delayed expressive abilities, have been shown to have lower scores on language measures through age 8 when compared to typically-developing children (Rescorla, 2002). Identification of subsequent language problems, which late talkers are at risk for, might not occur until after a child's kindergarten year, yet would lead to poorer vocabulary acquisition in the first years of school. It is possible the children with SSD in this study had been late talkers, although this was unknown at the time of the study. Additionally, the investigator selected the vocabulary words used in this pilot study from a predetermined set within Spanish curriculum. It is possible that individual children prefer some words to others, and the words included in this list were not those that the children particularly liked or remembered best. Finally, although the investigator used a standard curriculum, students may have experienced different levels of vocabulary exposure in different classrooms.

The study sample was heterogeneous, as each child with a SSD learned the vocabulary at different rates. These results suggest some link between extent of the communication problem and acquisition. For example, at the beginning of the school year, DC was diagnosed with a unilateral hearing impairment and SSD, yet at the time of the study she was able to produce most of her mispronounced sounds correctly in conversation. DC showed the greatest rate of improvement out of all the children; it may be that the speech sound errors of the other participants with SSD had a greater impact on their learning of new vocabulary. GH, for example, had the most impacted communication as he had concomitant speech and language diagnoses. He began and remained lowest out of the participants in expressive vocabulary abilities, yet his receptive understanding was comparable to other children with SSD. This possible link between the extent of SSD problems and rate of acquisition is conjecture, given the heterogeneity of the sample.

Limitations of this pilot study included sample size and composition, study design, and methodology of the Spanish receptive vocabulary probe. Because only five participants were involved in this study, it is difficult to generalize these results over a large population of children. The number of participants was limited for this study based on both the limited number of kindergartners with S/LD at the test school and the amount of time available to test the participants. As previously noted, the sample of children with SSD was heterogeneous, further limiting the scope of the results. Additionally, a longitudinal study design may have been more successfully tracking change over a longer period of time and allowing for a study follow-up. Finally, the Spanish receptive vocabulary probe was a somewhat flawed methodology; it did

not eliminate the possibility that children could simply guess by pointing at the correct picture flashcard without actually understanding the word spoken to them. Fluctuations in receptive test scores per child suggested this flaw; it is possible that scores decreased on any given day if the child was less successful at guessing. Given this knowledge, a more successful receptive vocabulary probe would include all 25 cards laid on the table at one time for the children to choose from, with selected correct or incorrect answers not revealed to the child. With this methodology, the chances of guessing would greatly decrease. A more thorough study design would be needed in order to further address the success of children with SSD in immersion schools.

The results of this study suggest that although children with SSD do not acquire as many Spanish vocabulary words as their mainstream classmates—a student following a normal school trajectory without extra academic help—they still acquired vocabulary words and increased the amount of words acquired over time. For some participants with SSD, their receptive abilities were comparable to the control participant, and the lag in their expressive abilities is not necessarily reason enough to exclude these children from immersion schooling. Genesee (1992) addressed a number of issues that place a child "at risk" in school, listing poor first language ability as one of those risk factors. Genesee (1992) stated that, "...without valid evidence concerning the suitability of immersion for children at risk in school, there is the danger that immersion programs could become elitist" (Genesee, 1992). Further research is needed to make immersion schools entirely accessible to all children, specifically research concerning the suitability of immersion schooling for children with S/LD, who might stand to benefit from such curricula as much, if not more, than their peers. Immersion schools are not intended to be elitist environments but are intended to share world languages with students to promote a more culturally aware society. The exclusion of certain groups of children was not the original aim of immersion schools and should not be a trend projected into the future.

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Appendix A

Table A.1. Vocabulary

Spanish Vocabulary Words	**English Equivalent**
Niño	Boy
Niña	Girl
Escuela	School
Mochila	Backpack
Basura	Trash/garbage
Maestra	Teacher
Baño	Bathroom
Zapato	Shoe
Ventana	Window
Puerta	Door
Ojo	Eye
Nariz	Nose
Cabeza	Head
Avión	Airplane
Brazo	Arm
Lápiz	Pencil
Pies	Feet
Papel	Paper
Chaqueta	Jacket
Libro	Book
Silla	Chair
Mesa	Table
Estrella	Star
Mano	Hand
Boca	Mouth

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Weapons-for-Oil Exchanges in Contemporary Sino-African Relations

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Abstract

Since 1949, China has gradually increased the scope of its weapons sales overseas to include 40 countries, 23 of which are located in Africa. With Chinese weapons sales in 2009 reaching 946 million U.S. dollars, the global community cannot help but pay attention to China's overseas weapons sales model. Over the past few years, resource-rich African nations have become the center of China's new geopolitical strategies and the starting point for oil development and extraction programs. With decades of cooperative experience in the energy sector, Africa has become an important area through which China is able to further diversify its energy resources. While China helped Africa develop an oil market, many other forms of aid and investment were also necessary. The most important form of aid can be best seen through "goods for goods" bartering transactions, especially in the form of Chinese weapons for African oil. This article will examine the realities of weapons-for-oil transactions as well as other Chinese involvement in Africa's natural resources and domestic economies, paying particular attention to how China is defining its interactions with African nations as those of a socialist leader. Through research and analysis of both historical and contemporary Sino-African relations, this article will illustrate the implications continued weapons-for-oil exchanges have for the future of Sino-Africa relations as well as for the international community.

China's Weapons Sales to Africa: Historical and Existing Circumstances

The People's Republic of China (PRC) began to pay attention to and work alongside the countries of Africa much earlier than the 21st century. Contemporary Sino-African relations can be traced back to the late 1950s when China signed bilateral trade agreements with Algeria, Egypt, Guinea, Morocco, and Sudan. From the very beginning of China's relations with African countries, China also offered developmental aid to support the socialist governments or anti-colonialist movements of these African states (Bitzinger, "Arms to Go"). According to research from the Stockholm International Peace Research Institute (SIPRI), from 1961-1971 China carried out a number of small light weapons sales to Africa (e.g., weapons that can be carried by an individual soldier), with sales totaling approximately $2.34 million[1] during this first decade (SIPRI Yearbook, 1960-1975). By 1976, however, China had quickly expanded the scope of its

[1] All numbers in this article are adjusted for inflation.

weapons sales to Africa, and sales profits increased to $537 million within five years alone (SIPRI Yearbook, 1976-1980).

Early weapons sales were heavily limited by restrictions on the timing of such sales as well as the numbers of weapons that could be sold. China was initially only able to supply basic light weapons as well as patrol boats, small tanks and fighter planes to African nations because of Chairman Mao Zedong's demands. Because Mao did not want China to be perceived by Western countries as a "Merchant of Death" (死亡商人), China chose to provide African countries with the most basic and affordable low-priced weapons and no-strings-attached military aid (Bitzinger, "Arms to Go" 92).

While Mao certainly hoped China could profit from weapons sales to Africa, the potential damage to China's international reputation from becoming a "Merchant of Death" at a critical point in the country's history outweighed the benefits of material gain. Mao knew quite well that weapons sales would provide a means to spread Chinese socialist ideology and further strengthen relations with developing African countries. But because China could not afford to damage its reputation in the eyes of Western powers, Mao chose to avoid weapons sales to African nations in the mid-1960s and early 1970s (Bitzinger, "Arms to Go" 93).

However, within a few years' time, China began to feel pressed for capital domestically and thus opted to increase weapons sales from just Egypt, Algeria and Sudan to include Tanzania, Zaire, Tunisia, Zambia, Sierra Leone and the Republic of Congo (Bitzinger, "Chinese Arms Production and Sales to the Third World" 28). These added sales included a small number of Soviet-style patrol boats, fighter planes, and tanks in addition to light weapons and ammunition (SIPRI Yearbook,1955-1980). Several countries among those receiving weapons from China did use these Chinese weapons in domestic or transnational conflicts, while others restricted the use of Chinese-bought weapons to military training, a trend that has continued to the present (Bitzinger, "Arms to Go").

Contemporary Chinese weapons sales to Africa began in the 1980s during the era of Deng Xiaoping's "Four Modernizations," a set of goals designed to make China a great economic power vis-à-vis development in domestic industry, national defense, agriculture and science and technology (Downs 31). For China's overseas weapons sales, the "Four Modernizations" paved the way for the redevelopment of weapons export policies. China not only increased the number of overseas weapons sales but also began to diversify the kinds of weapons it sold in foreign markets during this period; these changes are discussed later on in this paper. However, despite the changes in weapons sales policies, China became more interested in developing new markets in the Middle East and Southeast Asia, resulting in a decrease in the number of weapons exports to Africa (Bitzinger, "Arms to Go" 96). The weapons China sold to Africa in the 1980s were similar to those of the last decade, with only an increase in the numbers of weapons manufactured and minor improvements in technology (Bitzinger, "Arms to Go" 96).

Research from the Brookings Institute shows that Chinese and African interactions as well as China's domestic politics were highly interrelated. Even if China was ostensibly placing equal importance upon domestic and international affairs, domestic interests always took priority as China's own development continued to remain the basis of Chinese foreign policy (Downs 31).

For example, China's continued ability to access energy resources continues to be a top priority in order to ensure sustained domestic economic growth. Even at a time when the "Four Modernizations" were really just beginning to kick into action, it is clear that domestic interests fueled China's endeavors at the international level (Downs 31).

As a result of the Tiananmen Square incident in June 1989 and changes in the Chinese domestic political environment, the period from 1990 to 2002 witnessed a turning point in China's weapons sales to Africa (Bitzinger, "Arms to Go" 97). Ross, Whiting and Harding note that the sanctions imposed by Western countries on China after the Tiananmen incident generally involved the suspension of official visits, development assistance and export credits, and sales of military and police equipments. Ross, Whiting and Harding also found that the sanctions spurred by the events of June 1989 re-defined China's foreign relations, having a substantial "net impact … on China's foreign relations … and China [was] forced to look elsewhere in developing substantial and profitable military relations" (12).

While Africa was not China's largest weapons buyer at the time, African weapons purchases increased exponentially during the 1990s as compared to sales levels of the previous decade (Bitzinger, "Arms to Go" 98). According to statistics from the U.S. Department of State, from 1989 to 1999, China exported more than $1.6 billion worth of weapons to Africa, including $260 million worth of weapons to Northern African countries, $777 million to Central African countries and $647 million to Southern African countries (SIPRI Yearbook, 1980-2010). Weapons sales during the mid- to late-1980s were comprised mainly of patrol boats, tanks, older fighter plane models and light arms; in the 1990s, China began to sell fighter planes, transport aircraft, helicopters, military trucks, newer models of tanks and light arms, bombs, land mines and other similar weapons (Bitzinger, "Arms to Go" 99). Not only did the 1990s see a dramatic increase in Chinese weapons sales, but this era clearly signaled an increase in heavy armaments and aircraft sold overseas. It is also worth noting that China also began to invest increasing funds in Sudan (now known as the Republic of Sudan and South Sudan) during the 1990s and conducted large transactions of bombs and land mines to Sudan in addition to building three weapons manufacturing plants in Khartoum (Bitzinger, "Arms to Go" 101).

At the same time, while Chinese weapons sales to Africa increased dramatically during the 1990s, Russian and American weapons sales to African nations were both higher than those of China. Russian weapons sales to Africa, mainly to Egypt, totaled around $2.6 billion, whereas American weapons trade and related military agreements were valued at $9.8 billion (Shinn, "Military and Security Relations" 180).

According to a *New York Times* report, China supplied $1.5 billion in weapons sales to countries around the world between 2003 and 2006, representing only 2.9% of the global level of weapons sales during those years (Shanker). During this same time, China was ranked third in the list of countries selling weapons to Africa, moving to first place by 2008 (Zhang, "Analysis of Chinese Aid to Africa"). Most weapons sales were comprised predominantly of light weapons; however, China did supply numerous armored transport vehicles, tanks, frigates, artillery, fighter jets, training and transport aircraft and other heavy weapons to some African countries ("An Incomplete Record of Chinese Weapons Sales"). As recently as 2010, Chinese weapons

sales to Africa have also been rumored to include other types of jet aircraft, warships and/or spy planes ("Weapons Gone Wild").

As Chinese weapons sales to other nations have developed, the Chinese have actively sought out more cost-effective opportunities for continued transactions. One frequently debated change in Chinese aid to Africa is the decision to construct and staff weapons manufacturing plants near the Sudanese capital of Khartoum. Weapons made locally could thus be transferred directly to use in the Sudanese North-South Civil War or Darfur region without China bearing the logistical burdens of overseas transport (Shinn, "Chinese in African Conflict Zones" 7-8). With the recent formation of South Sudan, future relations between China and the new South Sudanese government will determine whether weapons are manufactured and shipped domestically or are imported.

While the Chinese-made weapons seen in overseas sales have undergone moderate changes, for the most part, they still remain outdated as near-replicas of Soviet weapons. Shinn, for example, points out that African countries possess a large number of Chinese-made weapons (mostly small armaments) dating to the 1960s and 1970s—weapons that imitate former Soviet technologies of the 1950s and 1960s ("Chinese Involvement in African Conflict Zones" 8-9). In its November 2010 online edition, *People's Daily*, the official Chinese Communist Party newspaper, notes America's increased attention to China's comprehensive weapons support for African countries that includes the copying and manufacturing of outdated Soviet-style armaments ("China Sells Old Weapons to Africa").

While developed countries would refuse to purchase older and outdated weapons, for many African nations, older weapons remain an advantage as these past technologies are simpler to use and repair (Luo). Chinese-made weapons are highly reliable and easy to operate in comparison to weapons made by other countries (Bitzinger, "Arms to Go" 84-85). Moreover, older, low-tech weapons technologies are more suited to meet the demands of developing countries that often lack military specialization or training (Bitzinger, "Arms to Go" 85). In recent years, African countries have also received more comprehensive weapons support from China that includes the copying and manufacturing of outdated and low-tech Soviet-style armaments ("China Sells Old Weapons to Africa").

On a continent of 57 countries, a total of 23 countries are currently active in purchasing a majority of their weapons and military equipment from China ("Africa"). Bitzinger points out that Chinese weapons are not only exceptionally easy to purchase due to the strong Chinese presence on the African continent, but are also an affordable alternative to weapons imported from Western powers. Therefore, African nations are able to purchase Chinese-made weapons and replace their more costly Western or Russian-imported counterparts ("China's Re-emergence as an Arms Dealer" 6). African nations buying Chinese weapons are not the only ones benefiting, as China continues to reap the economic and political benefits from such transactions (Kohli 1-2).

A Basic Survey of Chinese Weapons Sales to Africa

Chinese weapons exports to Africa originally began in 1955. These initial weapons transactions were more a reflection of Chinese political ideology than an embodiment of the Chinese desire for economic and financial gains (Bitzinger, "Arms to Go" 86). China perceived the opportunity to export weapons to Africa as a means of spreading socialism, a condition that took precedent to earning profits, ultimately resulting in a deficit from these early weapons sales to Africa (Michel and Beuret). Amongst all the African nations, China first established diplomatic relations with Egypt prior to signing numerous weapons and other trade agreements. Due to the length of Sino-Egypt relations and annual weapons deliveries, Egypt is thus the biggest buyer of Chinese weapons in Africa. China and Egypt signed their largest military and weapons treaty in February 1999, in which China agreed to provide 80 K-8 fighter planes for training, a deal worth $4.49 billion ("Weapons Gone Wild: Chinese Weapons Sales in Africa"). Comparatively speaking, while Egypt does not have huge domestic oil reserves, China and Egypt have developed their relations beyond basic weapons exchanges to the point that China has played a significant role in helping Egypt to develop its oil and natural gas industries, a trend that can be seen in China's relations with many other African nations (Voice of America).

Behind Chinese weapons sales to Africa lies the Chinese desire to further prevent Western imperialism and hegemony on the African continent. In the early years of China's diplomatic relations with African countries and weapons sales, China perceived the opportunity to deepen relationships with African states as a means to legitimize the newly-formed People's Republic of China at the international level (Rotberg). Growing closer to African countries was perceived as a means to prevent Taiwan (Republic of China) from becoming the representative of China (PRC) in international organizations (Meidan). Deeper relationships also provided a way for China to check the influence of Western countries in Africa and counter the possible future threat of Western nations on Sino-Africa relations (Meidan).

In addition to basic weapons transactions, China began a variety of lavish programs constructing infrastructure on African soil in hopes of further supporting the African desire for economic development. Rotberg detailed some of these infrastructure projects in his book, *China into Africa*, projects that range from building hospitals and educational institutions to constructing railroads and hydroelectric power plants in African nations. Not only do such projects foster domestic development in many African countries, but China's willingness to help Africa build much-needed infrastructure is a step towards further strengthening Sino-African relations (Rotberg). Furthermore, China has also supplied Africa with knowledge and technology in numerous fields to date, such as agriculture, forestry, natural resources and textiles (Rotberg). While China originally hoped that its involvement in African nations would provide a means to check Western development in the same regions, China has yet to recognize the negative side effects of their own presence on the African nations it has helped, an involvement that has fostered what Western media often labels a "neo-imperialist relationship" between China and African nations (Tan).

In 1978, as China embarked upon the new *gaige kaifang* ("Reform and Opening Up") policies under the leadership of Deng Xiaoping, the Chinese economy began to change at a rapid

pace, and military development was gradually deemed secondary in importance (Downs 31). However, with Iraq looking to China for weapons supply and armaments during the first Gulf War, China's military industry was quickly revived. In his research on China's weapons sales to the Third World, Bitzinger notes that China continued its traditional weapons sales to Egypt, but an increasing number of weapons and military equipment that arrived in Egypt were actually being secretly transported to Iraq and Iran for use in the Gulf War ("Chinese Arms Production to the Third World" 38). China was also able to supply fighter planes to Iran by first transferring these aircraft through North Korea (Bitzinger, "Arms to Go" 90). Despite the circuitous delivery routes of Chinese weapons sold during the Gulf War, China benefited from the conflict both by continuing to sell its weapons and by making a profit from the sales, thus contributing to the revival of China's military industry and domestic economic development (Bitzinger, "Chinese Arms Production to the Third World" 32).

Throughout the late 1970s and well into the 1980s, many third world countries were quickly becoming aware that using outdated weapons technologies and armaments was insufficient in contemporary military conflicts (Bitzinger, "Chinese Arms Production to the Third World" 36). This new awareness led to a sudden and abrupt decrease in Chinese weapons exports to developed countries during the early 1990s. At the time, a majority of African nations had grown especially close to Western powers and turned more frequently towards America and Europe to purchase weapons. Yet because of the developmental patterns of African nations in the 1990s, numerous African countries realized that the Western developmental model, a model that was trying to attach specific governance conditions to development, would not meet their own needs (Bezlova). For many African nations, China's fast, efficient, "no strings attached" bilateral development model was much more appealing (Broadman 38). Thus China and its lightening-speed economic development and strong military presence became the new choice for African countries in their search overseas for weapons dealers. In addition to providing an alternative model for domestic development, China was willing to provide no-strings-attached aid, more reasonably priced weapons and military training to African countries (Butts and Bankus 7-10).

At present, China continues to sell weapons to Africa not only because of an interest in continued, long-term access to natural resources, but also because armaments have become an important guiding direction in the development and maintenance of Sino-African relations (Butts and Bankus 11-13). China's Africa policies and direct involvement in African nations have been shaped by China's domestic long-term strategic interests and desire to continue its so-called "Peaceful Rise" to the status of a world power (He 28). The need for and importance of Sino-African relations is further solidified by the ongoing partnership between China, as one of the largest developing countries in the world, and Africa, a continent with a significant number of developing countries (Alden). Weapons-for-oil transactions not only benefit China's national interests but also have lasting reverberations for developing countries in Africa and elsewhere around the world.

Friendly relations between African states and China are well established and have developed further in recent years with increased economic and trade cooperation between the two. The relatively stable nature of Sino-African relations can be attributed to the importance placed upon the Five Principles of Peaceful Coexistence in the development and maintenance of Sino-

African relations. These principles are manifested in: mutual respect for territorial integrity and sovereignty, mutual non-aggression, non-interference in each other's internal affairs, equality and mutual benefit and peaceful coexistence (People's Republic of China, "China's African Policy"). Western foreign policy experts, such as Brautigam, Downs, Hanson and Hartung, argue that China's continued weapons sales to Africa are a direct contradiction to this foreign policy foundation because such sales inevitably lead to China's involvement in another country's internal affairs.

In understanding the extent of contemporary Chinese weapons sales to Africa, it is important to first examine and discuss the primary means through which Sino-African weapons transactions take place. Over the past few decades, China has pursued a very progressive and comprehensive foreign aid plan, particularly in Africa. Because China's low-interest, restriction-free loans "do not necessitate the market liberalization requirements that the International Monetary Fund (IMF) and World Bank stipulate," African countries are able to use Chinese loans toward more project-based objectives, such as building infrastructure or military development and arms acquisition (Tan). In 2010, for instance, Ghanaian President John Atta Mills visited China to sign a 20-year loan agreement worth $155 billion for use in construction of domestic infrastructure ("Ghana"). This hefty loan not only signifies China's perception of Ghana as an overseas investment target but also as a continued oil supplier. Moss and Rose of The Center for Global Development note that other African countries are no exception, as China continues to support other resource-rich countries with loans of all sizes in the hope that the African states will in turn use such money to invest in Chinese companies or purchase Chinese-made weapons. These loans (as well as the no-interest loans discussed below) are granted by China's Export-Import (ExIm) Bank, a state-owned bank charged with the responsibility of regulating Chinese loans to foreign countries and tracking the debt owed by other nations to China (Moss and Rose).

In 2006, the Chinese government published a white paper entitled "China's African Policy," a report that stressed the importance of Chinese corporations' expansion at the international level. This report recommended and supported continued investments and expansion of current Chinese state-owned enterprises in Africa, a policy move that encouraged continued loans and export credit from the Chinese government (People's Republic of China, "China's African Policy"). In conjunction with the release of this white paper, President Hu Jintao also announced that Chinese loans to Africa would gradually change to become interest-free loans by 2009, a promise that he ultimately did fulfill ("China Promises Billions in Aid"). China thus began to feed funding into African countries in other ways. While nowhere near complete, recent reports published by the Center for Global Development show Chinese aid to Africa in the following forms:

Angola	$2.16 billion in credit, a base rate likely to increase to $9.7 or $10.7 billion in the next five years
Ghana	$1.29 billion in zero-interest loans, including $600 million for construction of the Bui Dam on the Black Volta River
Mozambique	$2.48 billion in aid for various infrastructure programs (two dams, hydroelectric power stations etc.)

Nigeria	$1.72 billion in zero-interest loans for use in developing oil fields (fields which produce oil later exporter to China)
Republic of Congo, Sudan, Zimbabwe	Assorted zero-interest loans and credit

The financial aid provided by China's ExIm Bank to African states is based on the overall condition and economic development of the countries in question. For example, from 1997 to 2002 China's ExIm Bank supplied around $2.5 billion in zero-interest loans to Angola alone (Moss and Rose). In 2006, China National Petroleum Corporation (CNPC) signed an agreement with the government of Angola allowing Chinese access to and development of new oil fields, an agreement valued at approximately $1.56 billion (Kang 41). On top of this agreement, as a state-owned corporation, CNPC was able to encourage the government (and thus the ExIm bank) to continue to supply loans to Angola, bringing the total per annum amount of loans to Angola nearer to $3.35 billion (Kang 41). China's ExIm Bank has continued to financially support developing African nations like Angola because such development is advantageous to domestic Chinese interests in the long run.

While Chinese financial assistance to African nations may have ultimately resulted in a maldistribution of loans, the underlying incentives for granting such loans remain the same regardless of the amount and distribution of financial aid on the African continent. Monetary aid, as granted by China's ExIm Bank, gives African nations opportunities to diversify their economy and trade through economic development at the domestic and international level while also simultaneously promoting continued social development (Brautigam). Because such financial aid comes from China without any political, environmental or human rights prerequisites attached, the Chinese government expects that countries receiving support will abide by the "One China" principle, thereby acknowledging the People's Republic of China as the legitimate representative of all of China (Brautigam).

In response to the complex nature of loans and credit regulations, China has adopted easier terms for the sale of weapons to African countries. First, Sino-African bilateral transactions frequently include notions of "friendly pricing," a discount for African states on weapons in exchange for their continued recognition of China (PRC) and mutually beneficial diplomatic relations (Luo). The frequency of such discounts and the amount African states save on Chinese weapons purchases is unknown, however, as the Chinese government does not publish such information.

Second, China also remits the debt of African countries. Along with the publishing of China's white paper on African policy in 2006, the government also announced that, "China is ready to continue friendly consultation with some African countries to seek solution to, or reduction of, the debts they owe to China. [China] will urge the international community, developed countries in particular, to take more substantial action on the issue of debt reduction and relief for African nations" (People's Republic of China, "China's African Policy"). Within a year, at the annual meeting of the African Development Bank (ADB) Board of Directors, China announced that it would "relieve African countries of $12.2 billion worth of debt owed." Yet the reality of debt relief in Africa is perhaps much higher than what China initially stated. In 2010 at the United

Nations Summit on the Millennium Development Goals, Premier Wen Jiabao made known the fact that China has since remitted over $22 billion in African debt ("China Cuts African Debt").

Lastly, China has most recently pursued a "goods-for-goods" method in weapons sale transactions to Africa. This form of bartering is a business move by the PRC that has ultimately helped to somewhat alleviate the pressures of funding and military development faced by many developing African nations. "Goods-for-goods" transactions also began formally with the release of the white paper on African policy in 2006: "China will promote high-level military exchanges between the two sides and actively carry out military-related technological exchanges and cooperation. It will continue to help train African military personnel and support defense and army building of African countries for their own security" (People's Republic of China, "China's African Policy"). A moderate statement originating directly from the Chinese government, this component of the white paper insufficiently describes the extent and current state of China's military involvement and weapons sales on the African continent. China is truly only willing to provide loans, conduct weapons sales and deliver weapons to African states because such actions ensure continued goods-for-goods transactions and protection of China's biggest interest in Africa: energy.

China has relied heavily upon oil in its domestic development to date. In the mid-1990s, China was importing 11% of its oil per annum, with 46% of these imports coming from oil-rich Middle Eastern countries. By 2005, however, a noticeable change in oil imports is seen, as African oil exports to China increased to 31%, placing African oil second only to Middle Eastern oil (Downs 31). According to a 2010 report by the U.S. Energy Information Administration (EIA), China produces approximately 423 million barrels of oil per day domestically, while consuming just shy of twice this amount at around 835 million barrels per day ("China Energy Data, Statistics and Analysis"). With consumption levels far surpassing domestic production levels (Figure 1), China's high and consistent rate of oil consumption contributes to its status as the world's third largest importer of oil (behind the U.S. and Japan)

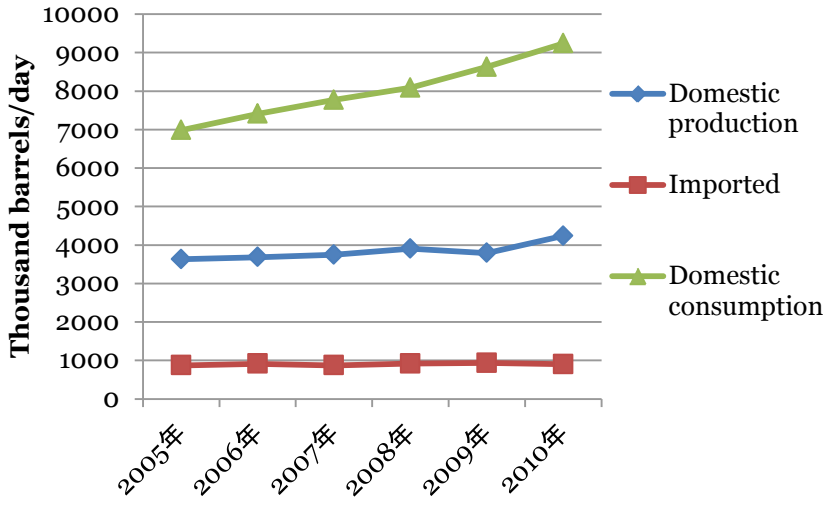

Figure 1. Chinese Oil Reserves. Source: "BP Statistical Review of World Energy."

In the wake of China's high-speed economic development and the popularization of cars in contemporary Chinese society, one of China's most immediate concerns is ensuring accessibility to overseas oil supplies. The Chinese government has continued, like other global powers, to look abroad for additional oil supplies and other forms of energy. African oil, with prices lower

than oil from the Middle East and a lower sulfur level that makes refinement easier, promises to help meet Chinese domestic oil demand (Hanson, "Vying for West Africa's Oil").

The African continent, with arguably the largest number of politically unstable regimes, civil wars, border wars, ethnic conflict, genocides, uprisings and rebellions in the world at present, is an ideal market for any weapons seller. Yet by opting to sell weapons to countries already plagued by domestic instability and conflict, China is further contributing to the instability. In recent years, Chinese weapons have emerged in the Somalia Civil War, Darfur conflict, Egyptian and Ugandan uprisings, Rwandan genocide, Nigerian revolution and other large-scale armed conflicts. One cannot deny that if China were to discontinue weapons sales to African countries, onlookers would likely see a decrease in the scope of conflicts on the African continent as a whole, since African nations would ultimately be forced to look elsewhere for new weapons suppliers (Michel and Beuret 152). Because few nations besides China produce the outdated weapons technologies African nations require, weapons sales to the African continent would thus decrease dramatically in the short-term while weapons manufacturers would scramble to come up with suitable armament options for the African market (Shinn, "Chinese Involvement in African Conflict Zones" 9).

For African countries lacking the financial capital necessary to purchase weapons, trading domestic oil reserves for Chinese weapons continues to be a very viable transaction method. Moreover, with low-technology weapons, little training is needed upon purchasing and African military forces (or rebels, in some cases) are able to get the weapons into the hands of their soldiers in a shorter period of time. As soon as the weapons leave China, the Chinese government argues that *how* the weapons are used in African countries is beyond their control. China firmly believes that weapons sales are by no means an interference with the domestic politics of any one African nation but rather just plain business (Brautigam). Regardless of how one evaluates the extent to which weapons sales reshape the domestic climate of African countries, continued weapons-for-oil transfers between China and African countries will influence the domestic development of African countries and their ability to build diplomatic relationships with other foreign powers.

The Weapons-for-Oil Dilemma

Africa is the largest oil-producing region in the world, with a majority of the oil reserves located either in Northern and Western African countries or offshore in these regions ("Africa"). Traditional oil-producing African nations include Nigeria, Angola, Libya, Congo, Algeria, Egypt and Tunisia. In the course of the past ten years, Equatorial Guinea, Sudan, Chad and Mauritania have all become important oil-producing nations as well. Oil exploration efforts have also increased exponentially in countries like Uganda, Guinea, Sierra Leone, and Liberia, thereby further contributing to the African continent's oil production levels. At the end of 2009, oil reserves in Africa totaled 1.28 trillion barrels, 150% higher than the 847 billion barrels estimated to be in African oil reserves ten years earlier ("BP Statistical Review of World Energy"). Predictions estimate that the coming 10 years will see the fastest development of the African oil industry to date, resulting in additional increases in the number of domestic African oil reserves ("Ten Years Later").

Statistics published by the U.S. Energy Information Administration in online analysis briefs indicate that China is currently the fastest growing energy-consuming country in the world. As early as 2003, China surpassed Japan to become the second largest consumer of oil. Ranked directly behind the United States, some experts predict that China will surpass the U.S. and rise to the title of largest oil consuming nation within the next 50 years (Wonacott 1). In the early stages of China's development, domestically produced and refined oil was sufficient in fulfilling Chinese demand. In recent years, however, with its demands for oil increasing at a rate of 13% annually, China was left with few alternatives but to look overseas for additional energy supplies ("China Energy Data, Statistics and Analysis").

To better ensure China's continued access to oil and other resources overseas, the Chinese government and energy-rich African countries reached a double-win consensus in the form of weapons-for-oil exchanges. Such exchanges ensure that African countries are never without arms for military use and China is never without oil to fuel its economy.

In the early 1990s, China and African nations formally began a cooperative relationship in the energy sector (Alden 63). With the arrival of the 21st century, and the continued growth of Sino-African economic relations, "Chinese and African cooperation in the energy sector is growing closer on a daily basis, expanding from oil trade to energy exploration and development" with many successful examples to date (Deng, "Sino-American Competition" 11-12). At present, China imports one-third of its oil annually from African countries, a mere 9% of all oil reserves on the African continent. The oil exported from Africa to China comes mainly from five countries: Angola, Equatorial Guinea, Nigeria, Congo and Sudan (Hanson, "China, Africa and Oil"). Sudan discovered domestic oil reserves as early as the 1970s but did not begin exporting for nearly twenty years. In hopes of developing these newly-discovered oil reserves, the Sudanese government quickly began to welcome overseas investment and offered loans and domestic assistance to countries that were interested in exploring and developing Sudanese oil fields. However, due to varying degrees of domestic conflict and political tensions, Sudanese infrastructure remained comparatively underdeveloped and posed a great risk to overseas investors looking to break into the Sudanese oil industry. For the Chinese, who saw Sudan both as a diplomatic ally and cooperative partner, it made sense to begin investing in development of the Sudanese oil industry. The Chinese involvement in aiding Sudan to explore domestic oil reserves marks the first time China began to help an African nation develop its oil industry and thus can be said to be an example of Chinese successes in this field. If China had not stepped in to provide the investments and logistical support necessary to build an oil industry from scratch, it is very likely that Sudan would not have been able to export any oil until the turn of the millennium (Lee and Shalmon).

At present, approximately 10% of the oil China imports annually comes from Sudan. Chinese investment in Sudanese oil fields is a unique method in and of itself; China not only buys oil extracted domestically (often via weapons-for-oil transactions) but also is responsible for the process of extracting and transporting oil. China National Petroleum Corporation (CNPC) decided in 1995 to participate in the development of Sudan's domestic oil industry, a move signaled by a loan granted by the Chinese government for CNPC use in Sudan. From that point forward, China agreed to supply Sudan with $2.5 billion in preferential loan agreements, and

CNPC signed a production sharing agreement with the Sudanese Energy Department to begin developing Block 6, one of the oil concessions in southeast Sudan. Within one year, CNPC won exclusive rights to the development of Blocks 1, 2 and 4 in the Muglad Basin of Southern Sudan as well as 40% ownership in the newly-formed Greater Nile Petroleum Operating Company (GNPOC). In addition, China built infrastructure necessary to the success of Sudan's newly formed oil industry, such as a 1,500 kilometer pipeline linking central Sudan with the Port of Sudan, where boats await oil for export to destinations around the globe (Rotberg). Sino-Sudanese energy cooperation has most recently taken on a new dimension when, at the start of 2011, the two nations began working on bioenergy in hopes of finding a substitute for traditional energy forms and encouraging Sudan to acquire more environmentally-friendly alternatives ("China-Sudan bio-energy project launched").

In order to ensure the security of Chinese investments and energy development programs in Sudan, China has since established a complete oil industry – including "extraction, refining, transport and retail" – a decision that has earned China the distinction as key player in the Sudanese domestic oil development (Deng, "Strategic Study" 52). However, China has also taken steps to further strengthen bilateral relations with Sudan in other areas by starting infrastructure programs to provide Sudan with hydroelectric power stations, airports, a domestic textile industry, and beyond. More importantly, China has continued weapons sales to Sudan and the Darfur region amidst China's growing dependency on the local oil resources. Despite China's investment in other areas of Sudan, it appears that China believes the best means for securing continued access to Sudanese oil is in developing stronger military relations with Sudan, especially through the medium of weapons sales. According to the UN Security Council Sudan Committee, 88% of the light arms currently in the Darfur region alone come from China. These weapons include, but are in no way limited to, Norinco-made 86S rifles (based on the Soviet-era AK-47), 122mm howitzer cannons, Type-59I 130mm cannons, 122mm missiles and 57mm antiaircraft guns (Kristof). China has supplemented light armaments with heavy weapons sales to Sudan, specifically Type-7M fighter planes, Y-8 transport aircraft, T-62 light tanks, and F-7/FC-1/J-6/J-7 fighter jets in recent years (Human Rights First; Figure 2). Lastly, China is currently making weapons in Sudan, but due to the covert nature of such manufacturing, research can substantiate only that these locally produced weapons include helicopters, ammunition, frigates and military transport vehicles (Human Rights First).

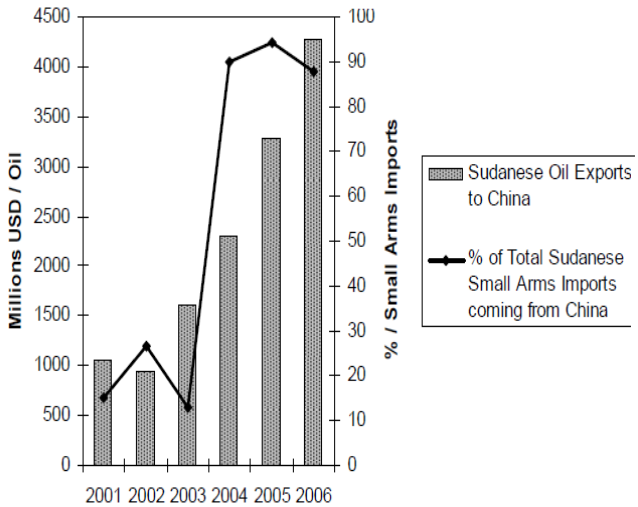

Figure 2. Sudanese Oil for Chinese Small Arms. Source: Human Rights First, U.N. Comtrade data.

China continues its main methods of selling the aforementioned weapons to Sudan, namely via loans and bartering. According to Human Rights First, China ExIm Bank has provided more than $1.27 billion in low- or zero-interest loans to Sudan over the past ten years. Normally, such large loans are reserved for impoverished and underdeveloped nations, but due to Chinese investments in and profits from Sudan's oil industry, Sudan is less dependent upon direct financial aid from China than other neighboring African countries. However, continuing loans are a means for China to safeguard its overseas interests, investments, and access to precious natural resources.

In recent years, Sudan has only been able to increase the amount of Chinese weapons it can purchase because of an increase in domestic profits from the oil industry. The Sudanese Minister of Industry, Dr. Awad Ahmed Al-Jaz, admitted that "Around 70% of domestic oil revenue is transferred directly to use by the Sudanese army," a statement suggesting that certain resources (including weapons) are likely be transferred directly to the army's hands and into the Darfur region (Human Rights First). One cannot overlook the fact that Sudan also has three weapons manufacturing plants opened and operated by the Chinese outside Khartoum (Human Rights First). China's move to produce weapons within Sudanese territory is a means for avoiding international sanctions which ban such sales to Sudan, while at the same time Sudan continues to supply China with oil in exchange for weapons.

Due to the extent of Chinese economic and military investments in Sudan, it is fair to conclude that the Sudanese government has been re-shaped by the Chinese presence. Therefore, even if the Darfur conflict is peacefully resolved, it is highly probable that China will still continue to sell weapons to Sudan in exchange for oil because this transaction method is time-tested and proven to be mutually beneficial. As the President of China's ExIm Bank, Li Ruogu, pointed out, "In the long term, regardless of the form Chinese investments to Sudan take, all such investments will be helpful in resolving the Darfur situation" (Bezlova). Here Li is referring to other Chinese aid to Sudan (such as infrastructure building programs) that will assist continued economic and political development, which in turn will ostensibly aid in resolving the Darfur conflict. Li's comment, however, blatantly overlooks the impact of Chinese weapons sales and transfers to the region. These points aside, Sudan will undoubtedly continue to receive Chinese investments and allow Chinese complete access to unrestricted development of the domestic oil industry. With many Sudanese currently viewing the Chinese as a people "of kind hearts… [and] a desire to remain separate from political problems," the Sudanese believe firmly that the Chinese investments are "a result of attractive business opportunities rather than a desire to reform the Sudanese political structure" (Goodman).

Without a doubt, continued Chinese weapons sales in exchange for Sudanese oil have triggered opposition at the international level. From a statistical perspective, 60% of oil reserves from the entire African continent are exported directly to China, a number that other nations have quickly begun to scrutinize (Hanson). Western superpowers – including the U.S., England, France and Russia – have openly criticized China's investments and weapons transfers on the African continent. Such a response is rather contradictory for many Western countries, since the U.S. and other militarily strong countries also sell weapons to developing African nations (Shanker). Due to the lack of set investment standards and transparency in the battle for energy

resources, it is difficult to account for any differences between the operations methods of state-owned Chinese corporations and their Western counterparts (Oliveria 293).

International organizations, most notably the United Nations Security Council, have responded even more strongly to continued weapons-for-oil exchanges between China and Sudan. In 2006, the UN Security Council confirmed not only that China was intervening in domestic African political problems but also that Chinese-made weapons were being actively used in domestic conflicts. In 2009, the UN Sudan Committee released a report at the Security Council regarding the use of Chinese-made weapons in Sudan. This report elaborated ways in which China was found to be violating an earlier UN arms embargo, namely in transferring weapons secretly through the capital into the Darfur region. The report from the UN Sudan Committee also proved the prevalence of Chinese-made weapons and their continued use in the Darfur conflict. The Chinese assistant foreign minister at the time, Li Zhaoxing, refuted such claims by saying that "China is not the only country selling weapons to Africa, [China] has no ability to control how the weapons are used after the sale is made. These weapons were likely to have been sold prior to the start of the Darfur conflict and we bear no responsibility for their arrival to the conflicted region" (United Nations, *Report on Resolution 1591*).

African countries receiving Chinese weapons and aid have yet another kind of reaction. Since China has actively invested in support of all types to at least 47 of the 57 countries on the African continent, many of these nations are all the more inclined to conduct high-level political dialogues with China. According to a 2010 public opinion survey conducted by Afrobarometer, a majority of Africans acknowledge the positive contributions China has made to the African continent, but remain suspicious of the political and economic impact of the Chinese government (Gadzala and Hanusch), an opinion that contrasts with the Sudanese sentiment previously mentioned. The sheer number of Chinese workers in many African cities exceeds the number of local laborers, exacerbating problems of unemployment plaguing many African countries. Chinese construction companies and contractors displace workers by the hundreds of thousands in stipulating quotas that employ predominantly Chinese nationals; elsewhere, Chinese products have monopolized local markets in outcompeting local manufacturers (Tan). In addition, many African citizens believe that China sees them as subordinates and thus is more prone to manipulate and bully the locals (Ampiah and Naidu 54). Yet Africans realize the extent to which their governments and economies have become dependent upon Chinese aid. Due to the extent of China's support to African countries – be it in the form of weapons, loans or other aid – discontinuing such aid would severely impact the development of countries across Africa.

Chinese weapons sales to Africa have also resonated within the Chinese government and popular discourse. China's continued weapons sales overseas are actually a violation of domestic export laws since such sales do not abide by China's domestic laws on weapons sales, "The Three Principles" (Ampiah and Naidu 173).

From a governmental perspective, Chinese involvement in Africa has also complicated the multidimensional nature of Sino-U.S. relations. First, China regards African countries as important partners in the process of Chinese development. The extent to which China has established partnerships with African nations serves as a model for other third-world countries

looking to develop closer relations with China. In the long term, due to Chinese demands for energy, the Chinese government will fulfill its promises and continue to support the development of African nations in as many ways as possible. The more of a Chinese presence there is on the African continent, the greater likelihood there is of competition between China and the U.S. over energy resources on the continent and domestic political sway. Second, China and Western nations have emerged as key competitors over African natural resources. In order to further ensure Chinese power at the international level, China will pay increasingly close attention to relationships between African nations and Western powers as well as any attempts by Western powers to further secure access to African oil and other resources. Third, current Chinese government officials have suggested that Sino-African relationships at present resemble a symbiotic colonizer-colony relationship (Alden). Chinese investments and weapons sales to African countries represent a business partnership between the two entities, but help in building infrastructure resembles actions an imperial power usually takes in developing a new colonial territory. Through such complex and multi-level relations, China has been able to replace Western influence and further assert its power and presence on the African continent.

To some extent, the emergence of a semi-colonial relationship between China and African countries is advantageous to the resolution of China's own domestic political problems, especially the so-called "Taiwan Problem." Taiwan is currently recognized by a mere four countries in Africa: Burkina Faso, Gambia, Swaziland and the islands of Sao Tome and Principe. The financial aid China is able to provide in Africa far surpasses Taiwan's own capabilities, a factor that has weighed heavily in the minds of African leadership as they adopt the "One China Policy" in favor of receiving larger aid packages from Mainland China.

China's investments and weapons sales to African countries have also resulted in two forms of response at the Chinese public level, namely liberal and conservative viewpoints. Liberals believe that China should pay no attention to the "China Threat Theory"[2] that is emerging from Western countries vis-à-vis China's relations with Africa and rather should continue to place domestic interests as top priority in guiding Chinese foreign affairs. The conservative viewpoint, on the other hand, argues that China needs to be most concerned with preserving its image at the international level and should thus decrease the amount of investments and energy ventures it is currently conducting in Africa (Huang, "China's Renewed Partnership" 305-310).

In short, Chinese weapons sales and financial aid in African countries are a highly debated matter. Regardless of what perspective one adopts in evaluating these weapons-for-oil transactions, it is clear that Chinese motives extend far beyond using weapons and other aid as a political bargaining chip. Based on the current state of Sino-African relations, it seems very likely that China desires to continue to develop its economic and political dominance on the African continent in hopes of strengthening Sino-African policies and broadening strategic weapons-for-oil transactions.

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[2] Proponents of the "China Threat Theory" see threats to the dominance of Western powers stemming from China on two levels, namely militarily and ideologically.

The Future of Chinese Weapons Sales to Africa

In recent years, Sino-African relations have entered a new stage as bilateral relations have progressed to the level of a strategic partnership. This new form of diplomacy in Sino-African relations is comprehensive and includes cooperation on issues of political, economic, cultural, security, and international importance. Especially worth noting is the progress China and Africa have made in cooperative efforts within the energy sector. Lacking sufficient domestic resources amidst an ever-growing population, China can only turn to imported oil and other imported energy resources to fulfill domestic demands resulting from continued economic growth and development (Li 16-19).

Due to China's sustained demand for energy and the resulting expansion of Sino-African trade over the past few years, China's weapons sales to African nations cannot but continue to increase in the coming years (Meidan 90-93). China and many African nations have developed a form of mutually dependent relations, a state unlikely to change in the near future. China Petroleum and Chemical Corporation (SINOPEC) has invested $7.4 billion in Sudan's oil industry to date, an investment that includes the construction of a transnational pipeline as well as ownership of 60% of Sudan's domestic oil fields (Perry 16) . Due to the Chinese government's push for domestic businesses to develop at the international level and China's continued investments in the African oil industry, Sudan and other oil-rich countries have been transformed from oil-importing to oil-exporting nations, including capabilities in accessing, refining, transporting and selling oil. Indeed, without China's support and financial investments, natural resources in many African countries are unlikely to have undergone such a comprehensive transformation in such a short period of time. China's interest in accessing precious energy resources and other overseas investment opportunities has proven to be rather advantageous to the economic development of African nations as well, since such interests often generate financial aid to jumpstart local economies and strengthen international trade.

In short, provided that China will continue to reap the benefits of investment and interaction with African nations, such investments are unlikely to stop. With weapons-for-oil transactions as a central component of Sino-African trade, China will not only continue to ensure its access to oil, but also continue to develop a presence in the domestic weapons market of many African countries. Moreover, with continuing political instability and conflict in Africa, the demand for cheap Chinese-made weapons is very likely to increase. With weapons-for-oil as the preferred transaction method by China and African countries alike, China will most certainly continue to provide a large number of low-technology and affordable weapons for African countries to use.

The future of Sino-African military relations, and thus sustained weapons-for-oil transfers, rests heavily upon three important factors. First, while China has invested and conducted weapons-for-oil transactions in a large number of African nations, China is unlikely to continue to isolate itself from military governments and internal conflicts of African countries. To protect its interests in the long term, China will become even more willing to invest in high-risk economic or political environments that plague many countries on the African continent, a risk that Western countries and corporations are frequently unwilling to take. Second, future forms of Chinese military support to Africa will not just involve weapons sales but are likely to also

include more zero-interest loans. Such loans from China's ExIm Bank will be used by the receiving African state to further develop its military or cultivate necessary infrastructure. Third, China will continue to develop the military training programs already in place on the African continent. Regardless of linguistic and cultural barriers, continued military training will further expand Sino-African military relations and ensure that African troops are learning to correctly operate the Chinese-made weapons they have acquired. With these three factors in mind, it is clear that military relations between China and Africa in the long run will result in increased benefits for both sides.

Due to the likelihood of continued Chinese weapons sales to developing African countries, other countries and international organizations must consider whether or not such transfers of weapons and energy are likely to result in conflicts at the international level. It is also important to consider the reaction of other world powers and international bodies, since the Chinese weapons sales are mainly targeted at countries in the midst of civil wars, ruled by military governments, or countries with numerous human rights violations. As a whole, European countries and the United States are the loudest voices of opposition to Chinese weapons sales in exchange for oil on the African continent. Yet such opposition carries little weight as China continues to pursue its domestic interests at the international level. China continues to overlook UN sanctions on weapons sales to Africa, while the international community merely voices opposition and utilizes tools of soft diplomatic power rather than pursuing tangible actions and policies in response to continued sanction violations. This passivity in responding to Chinese sanction violations suggests that perhaps Western powers also maintain an interest in weapons sales to Africa or African energy resources (Zhao 24). Regardless, if other weapons manufacturing countries, such as India, Russia, and the United States, embark on large scale weapons-for-oil transactions of their own in Africa, China's key position and investments will be directly threatened, a change that could ultimately lead to tensions between countries over rights to access energy.

Conclusion

Based on Africa's current level of dependence upon Chinese support, investments, and weapons sales, it is highly likely that a majority of African countries will continue to encourage Chinese development of the continent. China's steadily growing demands for oil, the profits that China reaps from weapons sales and the outcome of other Chinese investments in Africa truly do (and will continue to) enhance the mutually beneficial nature of Sino-African relations. The short-term costs of weapons transfers, infrastructure development and loans to African nations are negligible in comparison to the long-term profits China will gain through continued access to African oil and other natural resources. More importantly, China will continue to view itself and its actions abroad as those of a socialist leader rather than actions of a so-called "Merchant of Death." Although Chinese policies to sell weapons to African countries in exchange for oil receive much criticism from the international community, especially in light of recent events in Egypt and Libya, China remains unlikely to change its current trajectory. With other nations pulling away from unstable and war-torn African countries, especially as domestic conflicts arise, China is facing less competition from Western nations that also initially emerged on the

African continent in search of energy sources. Furthermore, even greater opportunities abound for a sustained Chinese presence on the African continent with ongoing foreign aid packages, infrastructure building, and other forms of goodwill constantly streaming from China into Africa. Thus if any resistance to Chinese expansion in Africa presents itself, it is most likely to come from the international community or Chinese domestic discontent rather than from the African nations that reap countless benefits of Chinese support.

Due to the likelihood of ongoing weapons-for-oil transfers and the continual development of Sino-African relations, the United States must also consider an appropriate response in order to ensure that its interests and relationships with China and African nations alike can be best maintained. The intimacy of Sino-African relations has resulted in much debate within America, a debate that focuses heavily upon how America should approach Sino-American relations with the addition of the new "Africa Factor" into the equation. In the eyes of many American politicians, a stable African continent composed of countries that advocate democracy and democratic rights is most advantageous to U.S.-Africa relations, a goal commonly reflected in U.S. policy towards African countries.

Along with the continued development of Sino-African relations, the United States should pursue three key goals. First, the U.S. should continue to monitor whether Chinese involvement on the African continent (especially weapons-for-oil transactions) is in the best interests of other countries around the world. Second, the U.S. can encourage greater transparency from China about its loans, weapons sales and other support given to oil-rich African nations. If the U.S. is able to understand the larger context of China's investments in Africa and domestic factors which have spurred China's search abroad for oil, the U.S. and China should ultimately be able to find more common interests on the African continent. Such shared interests could in turn become the basis for a cooperative relationship in Africa, a step in furthering the peaceful and stable development in Sino-US-Africa triangular relations. Third, U.S. government officials must pay more attention to the opinions and attitudes of Africans in reaction to the increased Chinese presence within their own societies. Through the pursuit of these three goals, the U.S. can help to ensure that its relationship with China remains stable and tensions between the two nations do not emerge on the African continent.

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Exploring the Adaptiveness of Moderate Dissociation in Response to Betrayal Trauma

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Abstract

Freyd's (1996) betrayal trauma theory posits that evolutionarily important attachment bonds make dissociation in response to trauma more likely when a relationship exists between victim and perpetrator. This dissociation, despite its immediate benefits in regards to attachment, is commonly thought to have harmful consequences over time. However, although negative mental and physical sequelae may result from chronic dissociation, it may also continue to serve a protective function in regards to attachment. This online study explores the relationship between dissociation, resiliency, betrayal trauma, and attachment using self-report questionnaires with a college student sample (400 participants, 68.7% female). We hypothesized that participants with moderate dissociation would be more resilient to childhood abuse and utilize multiple attachment strategies. Results revealed that higher dissociation was associated with poorer resiliency scores, although in a curvilinear analysis very high dissociative scores correlated with higher resiliency. Dissociation did not seem to be related to attachment; however, participants with a history of betrayal trauma were associated with more variability in attachment styles. These findings are particularly relevant because it could inform attachment theory and the effects of trauma on attachment.

Introduction

It is difficult to imagine how a person could possibly forget a traumatic experience, yet trauma-related amnesia and dissociation are well-documented phenomena (Herman, 1992). In an attempt to explain why forgetting and dissociation occur, Freyd (1996) developed betrayal trauma theory. According to this theory, the act of betrayal is the key factor influencing whether a person is likely to forget or dissociate from an event. In other words, betrayal trauma theory posits that forgetting and dissociation are more likely when a close relationship exists between the victim and the perpetrator. The importance of this relationship can be explained by attachment theory.

Attachment theory states that a biologically based bond exists between a child and her/his caretaker that ensures the protection and safety of the child (Bradbury & Karney, 2010). A child may forget abuse by a caretaker because remembering would endanger the child's relationship to, and thus its ability to trust and depend on, that individual (Freyd, 1996). In this way, amnesia of traumatic events allows the child to maintain an attachment to her/his caregiver. For humans, this is a matter of evolutionary importance (Bradbury & Karney, 2010). Although

Freyd's betrayal trauma theory explains why it could be adaptive for a child to forget or dissociate from traumatic memories, research exploring whether dissociation is an overall adaptive response to trauma is limited.

One problem inherent in viewing dissociation as an adaptive function is the fact that dissociation has typically been considered a significant risk factor for developing Post Traumatic Stress Disorder (PTSD). Using a meta-analysis, Ozer, Best, Lipsey, and Weiss (2003) found that dissociation was the single most predictive variable for developing PTSD when compared to six other risk factors including: previous psychological adjustment, family history of psychopathology, perceived life threat during trauma, prior trauma history, level of social support following the trauma, and dissociation in relation to a person's emotional response. Dissociation, described at times as self-hypnosis, is protective in that it allows a person to mentally escape a trauma (Terr, 2003). However, when employed outside a traumatic context, as is common with people who dissociate, inaccurate analyses of events and the world often occur. Therefore, dissociation can be seen as adaptive specifically within the context of trauma but is maladaptive once the danger is past (Herman, 1992). While dissociation is considered a significant risk factor for PTSD it is still possible that it could help protect against other psychological effects of trauma.

Traumatic events are more than merely frightening or life-threatening—they are life-altering in that they have the power to destroy a person's basic concept of how the world is supposed to operate. One of the most lingering effects of trauma is rooted in the fact that trauma destroys "the victim's fundamental assumptions about the safety of the world, the positive value of the self, and the meaningful order of creation" (Herman, 1992). The trauma cannot be assimilated within the victim's inner schema, or template of the self and its relationship to the world (Herman, 1992). This leads to a loss of faith and connection to others and society. For example, one question brought about by a traumatic war experience—how could God allow young children to die (Herman, 1992)?—is a question that could not be logically answered within the framework of the victim's current belief system. In order to accept and accommodate such trauma, these views must be altered and may subsequently change a person for the rest of her/his life (Herman, 1992). Trauma, when it occurs in adulthood, affects a person's inner schema of how the world operates; however, as children have not yet formed these schemata, the consequences can be very different.

Young children are entirely dependent on their caretakers, which makes abuse by a caretaker or trusted other particularly detrimental because a child must continue to trust and maintain a relationship with that person. Therefore, when confronted with abuse of this type a child is forced to "find a way to preserve a sense of trust in people who are untrustworthy, safety in a situation that is unsafe [and] power in a situation of helplessness" (Herman, 1992). A child must live in a world that is, by the nature of her/his situation of dependency, completely contradictory. In order to do this, children resort to psychological defenses which preserve their primary attachments to their caretakers regardless of their abuse (Herman, 1992). For example, a child may blame him/herself for the abuse or see him/herself as a "bad child" in order to free the caretaker of responsibility. According to the child, the caretaker is not evil, and it is the child who deserves punishment (Herman, 1992). This and other cognitive strategies function to

minimize the trauma and preserve the attachment bond. It is this sort of cognitive monitoring or dissociation that can also lead a child to forget traumatic events. Although the immediate benefits from cognitive strategies of this type are readily acknowledged, it has yet to be determined whether these coping strategies remain beneficial throughout life, or even if children who dissociate from childhood abuse have a better attachment to their caretaker than those who do not.

To determine if dissociation protects attachment relationships, attachment theory must first be discussed. Attachment theory recognizes four different attachment styles: secure, preoccupied, dismissing, and disorganized (Purnell, 2010). These styles can be broken down in terms of beliefs about the value of the self and the other. This in turn determines levels of anxiety and avoidance concerning attachment relationships. Secure attachment is characterized by a positive value of the self and other, and is thus considered low in terms of avoidance and anxiety (Bradbury et al. 2010). A preoccupied attachment style, involving high anxiety but low avoidance, is rooted in a negative value of the self but positive value of the other. A person with this type of attachment style is often preoccupied that he/she is unworthy of the attachment figure's love. Individuals with a dismissing style believe in the positive value of the self but negative value of the other—leading these people to have low anxiety but high avoidance. Dismissing individuals tend not to regard relationships as beneficial. Finally, disorganized attachments are characterized by high anxiety and high avoidance and are formed by a negative value of the self and other.

Of all of these styles, a secure attachment with a caretaker is considered ideal. Secure attachment is associated with better outcomes and adjustment throughout the life of the child (Bradbury & Karney, 2010). Contrarily, a disorganized attachment is associated with children who have experienced trauma (Purnell, 2010). This style is tied to many negative outcomes and poor adjustment, as well as to dissociative tendencies (Purnell, 2010). Therefore, the argument that children who forget caretaker-induced trauma are able to maintain a secure attachment style is a difficult one to make. However, in regards to the context of childhood abuse, a disorganized attachment style may be associated with resiliency.

Although a disorganized attachment style has been associated with negative adjustment, it may be unfairly singled out, as dismissing and preoccupying styles are also associated with poor outcomes. Disorganized attachment is generally separated from the other three models because it is considered an 'unmindful' or irrational attachment strategy. This is because it combines aspects of both dismissing and preoccupying styles (Liotti, 2004). Therefore, disorganized attachment has been considered a collapse of the attachment system and is associated with extreme trauma and/or neglect (Purnell, 2010). However, this assertion has been questioned. In the dynamic maturation model of attachment, Crittenden (2000) theorized that attachment styles are always deliberate, and although disorganized styles may seem to mix contradictory strategies, with extremely unpredictable caregivers this may in fact be an appropriate response (Purnell, 2010).

The idea of disorganized attachment as an adaptive strategy is also supported by evidence involving studies of previously diagnosed disorganized infants in middle childhood. These

children, categorized as disorganized in infancy, received a diagnosis of disorganized controlling in middle childhood (Liotti, 2004). Appearing to be normal, competent individuals, these children differed only in that they were often inappropriately controlling of their caretaker or inappropriately caring. This can be viewed as a strategy to accommodate an incompetent caretaker by taking control of the situation (Liotti, 2004). These slightly older children seemed to employ a reasonable strategy for dealing with their situation. In fact, the disorganized aspect of their assessment did not manifest until they were confronted with a separation anxiety test (Liotti, 2004). Although typically associated with negative outcomes, in this situation, a disorganized attachment can be seen as a logical response to an incompetent caretaker in the moment. It remains to be seen if this response is beneficial over time.

The advantage to a disorganized attachment and the related dissociation may lie with the ability to compartmentalize trauma both in relation to knowledge of the abuse and how attachment relationships are affected. Freyd (1996) cited an account of a victim of incest who dissociated from her abuse by separating herself into a day child and a night child. The night child was aware of the abuse but the day child was oblivious beyond the fact that she knew she was afraid of the night. This separation or compartmentalization is argued to be a cognitive strategy that allows the child to function normally during the day. It is also a strategy that can be viewed in relation to a child's attachment. John Bowlby, a psychoanalyst who studied child development, developed an attachment theory which states that children form internal working models or schemas of their relationship to their primary caretaker and that these models form the lens through which all other deep attachments are perceived (Bradbury et al., 2010). Following this model, a child who dissociates from abuse perpetrated by a caretaker may have a disorganized attachment style. However, because dissociation may allow the child to compartmentalize the trauma, this disorganized attachment may be an internal working model through which the child interprets some relationships but not others. In other words, the actual disorganized nature of the attachment style, as in the earlier example, could allow the child to respond appropriately to the caregiver during the day versus during the night. This is supported by recent criticism of attachment theory that states attachment styles can change and that infants can have multiple attachment styles depending on a particular caretaker (Del Guidice, 2009). Dissociation, because it can allow a child to compartmentalize trauma, can be viewed as an adaptive strategy over time.

Compartmentalization of trauma could allow an individual to demonstrate resiliency in the face of trauma, but resiliency needs to be defined, as well as the distinction between recovery and resiliency. A study by Bonanno (2005) found that resiliency to trauma was not necessarily the same thing as recovery. Bonanno found that resiliency is associated with an individual's ability to operate at normal levels with low levels of psychological disturbance. Recovery was associated with higher immediate negative psychological symptoms that significantly interfered with an individual's functioning, before gradually declining to normal levels (Bonanno, 2005). Therefore, a child that dissociates from abuse, although not effectively dealing with it, may be cognitively compartmentalizing the trauma by pushing it out of consciousness—allowing him/her to maintain a normal or near-normal level of functioning. This is supported by cases like that of Ross Cheit, who recovered memories of childhood abuse when he was 36 years old

and a professor at Brown University (Freyd, 1996). Based on such evidences, dissociation that allows a victim to compartmentalize trauma and maintain high levels of functioning could be highly adaptive, even if it is also associated with negative sequelae.

Dissociative disorders tied to child abuse (Herman, 1992) are not beneficial once a child is free of the situation that forced him/her to employ such extreme cognitive strategies. This is because people with a history of abuse often have difficulty accurately assessing dangers in the world around them. However, dissociative tendencies lie on a continuum (Bernstein & Putnam 1986). In other words, people with mild to moderate dissociative tendencies may be more resilient when faced with repeated childhood trauma than people who experience either high levels of dissociation or little to no dissociation. In regards to childhood trauma involving attachment bonds that are so essential, this could very well be the case. It is therefore possible that, firstly, people with moderate dissociative tendencies are more resilient to childhood abuse than those with very high or very low tendencies, and that, secondly, people with moderate dissociative tendencies can be expected to utilize multiple attachment strategies. A hyperbolic relationship is predicted for the relationship between dissociation and resiliency, with a certain level of dissociation being beneficial. This dissociation, in turn, is expected to be associated with greater variability in attachment styles.

Method

Participants

Participants signed up for the study through the University of Oregon's online system for study management, which enables students in the psychology and linguistics departments to receive class credit for participating in research. The ethnicity of the participants therefore reflects the ethnicity of the university student population. Participants were 1.5% American Indian/Alaska Native, 9.5% Asian, 1.5% Native Hawaiian or Other Pacific Islander, 2% Black or African American, 79.3% Caucasian, and 5.5% other. 0.8% of participants declined to self-identify ethnicity. Participation in the study required fluency in written English but there were no other restrictions. In total, 514 people participated in the study and complete data collection included 400 participants in the analysis. . Of these, the average age was 19.9 years with 68.8% female participants and 31.3% male participants. This ratio represents a somewhat biased sample, with women being more likely to experience a betrayal trauma than men (DePrince & Freyd, 2004).

Materials

Four different measures were included in the study. To begin, the Brief Betrayal Trauma Survey (BBTS; Goldberg & Freyd, 2006) was used to measure trauma exposure. The BBTS measures both betrayal trauma (if a person has been abused by a trusted other) and non-betrayal trauma. The BBTS includes two sets of 14 identical questions, targeting experiences before and after the age 18. A sample item includes "You were made to have some form of sexual contact, such as touching or penetration, by someone with whom you were very close (such as a parent or lover)." Participants then responded based on how many times the above event

occurred, and response options for frequency questions were as follows: Never, 1 or 2 times, and More than that.

The second survey used was the Dissociative Experience Scale (DES II; Bernstein & Putnam, 1986), which measures dissociation tendencies. This scale is a 28-item measure that examines the percentage of time a person spends in a dissociative state. An example item includes, "Some people have the experience of finding themselves dressed in clothes that they don't remember putting on." Participants are then asked to "choose a number to show what percentage of the time this happens to you" with options ranging in 10% increments from 0% of the time to 100% of the time.

The Inventory of Parents and Peer Attachment (IPPA; Armsden & Greenberg, 1987), was the third measure used. It is a useful measure because it assesses attachment in more than one domain. It is an assessment of an individual's attachment to his/her parents and an individual's attachment to her/his peers. The measure is comprised of 53 items, 28 pertaining to attachment to parents and 25 that assess attachment to peers. The scale can be further broken down along three domains with items measuring trust, communication, and alienation for both parents and peers. An example of a trust item regarding parents would be, "My parents accept me as I am," while an example of an alienation item for peers would be, "I get upset a lot more than my friends know about." Participants then chose between five options: Almost always or always true, Often true, Sometimes true, Seldom true, and Almost never or never true.

The last measure presented to participants was the Trauma Resilience Scale (TRS; Madsen, 2010). The TRS is a 48-item measure of resiliency along four domains, including the ability to generate and maintain supportive relationships, optimism or hopefulness, spirituality, and problem solving skills. The scale includes 13 spirituality items, 13 relationships items, 12 optimism items, and 10 problem solving ability items. A sample question for optimism includes, "Most people say that I have a hopeful outlook on life." Participants then chose from seven options: Almost never true of me, Rarely true of me to, Frequently not rue of me, Sometimes not true of me/Sometimes true of me, Frequently true of me, Very often true of me, and Almost always true of me.

Procedure

Participants were recruited online via Sona Systems and the study was administered online. The study could therefore take place at any location where internet access was available. The surveys were always presented in the same order, beginning with the BBTS followed by the DESII, then the IPPA, and finally the TRS. As all surveys were self-report questionnaires, a controlled environment was not deemed necessary. The average amount of time that it took to complete the entire study was 35.23 minutes.

Results

Descriptive Statistics

See Table 1 and Table 1.1 for detailed descriptive statistics. Table 1 shows the total percentage and numbers of participants with either high betrayal trauma histories or low/no

betrayal trauma histories. This grouping was determined by placing all participants that reported a betrayal trauma in the betrayal trauma group while all other participants were placed in the low/no betrayal trauma group. In addition, 42.3% or 169 participants reported interpersonal violence (physical or sexual assault which includes both betrayal and non-betrayal trauma) while 57.8% or 231 participants did not. These numbers are similar to those found by DePrince and Freyd (2004) in a college sample that reported interpersonal violence rates of 45%. Moreover, they found that these rates were lower than a community sample that had rates of interpersonal violence at 79%.

Table 1.1 shows the means, standard deviations, and minimum and maximum scores for the Trauma Resilience Scale (TRS), the Inventory of Parents and Peer attachment just including parents (IPPA Parents), the Inventory of Parents and Peer attachments just including peers (IPPA Peers) and the Dissociative Experience Scale II (DES II). For the TRS, the highest possible score (i.e. a score of seven for all questions) is 336 while the lowest possible score (i.e., a score of one for all questions) is 48. For the IPPA parents the highest score possible is 140 with the lowest score possible being 28. For peers the highest possible score is 125 while the lowest is 25. In regards to the DESII, the measure includes 28 items, and the highest score possible for each item is 10 and the lowest possible score is zero.

Table 1. *Breakdown of trauma history into High and Low/No betrayal trauma groups.*

BBTS	**High Betrayal**	**Low Betrayal/No Betrayal**
	33.2% (129 participants)	67.8% (271 participants)

Table 1.1. *Descriptive statistics for Trauma Resilience Scale (TRS), Inventory of Parents and Peer Attachments (IPPA), and the Dissociative Experience Scale II (DESII).*

Measure	Mean	Standard Deviation	Minimum	Maximum
TRS	236.35	39.91	122.00	332.00
IPPA Parents	94.10	10.28	59.00	115.00
IPPA Peers	89.88	9.28	50.00	113.00
DESII	11.85%	11.15	0.00%	58.00%

Questions about dissociation

A linear regression model and independent t-test was used to determine the effects of dissociation on trauma resiliency. The TRS score was the dependent variable and the DES score was the independent variable. Higher dissociative scores were significantly predictive of lower resiliency, $F(1, 400) = 21.05$, $p < 0.05$, $r^2 = 0.05$. That is, people with higher dissociative scores were less resilient to trauma. Betrayal trauma was also found to predict dissociation. Results of a one-way ANOVA with the DES as the dependent variable and the independent variable as trauma category (high betrayal trauma or no and/or low betrayal trauma) indicated that high betrayal trauma predicts higher dissociation scores, $F(1, 397) = 9.58$ $p < 0.05$, $r^2 = 0.02$ compared to low or no betrayal trauma.

In a test of the first hypothesis, which questions the relationship between dissociation and resiliency, a quadratic regression model was used to examine a curvilinear relationship between resiliency and dissociation (see Figure 1). The dependent variable was the TRS score and the independent variables were DES score and the DES score squared. Higher resiliency was found in people with very high dissociative scores compared to those with moderate dissociative scores, $F(1, 400) = 11.10$, $p < 0.05$, $r^2 = 0.03$. This is contrary to the prediction that a benefit to dissociation would be found in people with moderate dissociative scores as opposed to very high scores. Moreover, results were also significant when this analysis was repeated controlling for betrayal trauma, $F(1,400) = 5.65$, $p < 0.05$ and $r^2 = 0.01$, and for a non-betrayal trauma, $F(1, 400) = 6.26$, $p < 0.05$, $r^2 = 0.02$.

Questions about betrayal trauma and resiliency

Results of a one-way ANOVA, where the dependent variable was also the TRS score and the independent variable was high betrayal trauma or no and/or low betrayal trauma, indicated that betrayal trauma predicted lower trauma resiliency, $F(1, 397) = 6.91$, $p < 0.05$, $r^2 = 0.02$. A linear regression model where the dependent variable was the TRS score and the independent variable was a continuous variable measuring number of betrayal traumas, also indicated that higher betrayal trauma scores predicted lower trauma resiliency, $F(1, 400) = 9.98$, $p < 0.05$, $r^2 = 0.02$. A linear regression model using number of non-betrayal traumas instead of betrayal traumas also revealed that higher numbers of trauma predicted lower resiliency, however the effect size was smaller, $F(1, 400) = 5.54$, $p < 0.05$ and $r^2 = 0.01$.

Questions about attachment

A linear regression model with the dependent variable being the difference scores between parents and peers and the DES score being the independent variable was used to test the second hypothesis. Results found that people with higher dissociative tendencies were not more likely to have greater differences between parent and peer attachments, $F(1, 400) = 0.67$, $p > 0.05$, $r^2 = 0.00$. However, in a one-way ANOVA, individuals that experienced betrayal trauma were more likely to have a greater difference in attachment scores in regards to parents and peers than individuals that did not experience betrayal trauma, $F(2, 400) = 10.70$, $p < 0.05$, $r^2 = 0.03$.

A one-way ANOVA with a dependent variable of IPPA scores for parents and an independent variable of either high betrayal trauma or no and/or low betrayal trauma was also conducted. Betrayal trauma was found to predict lower attachment scores to parents $F(1, 397) = 10.56$, $p < 0.05$, and $r^2 = 0.03$. However, a repetition of this analysis using IPPA scores for peers, found that betrayal trauma did not predict attachment to peers, $F(1, 397) = 0.08$, $p > 0.05$ with $r^2 = 0.00$. Interestingly, a linear regression model with the TRS as the dependent variable and difference scores between parents and peers as the independent variable found that higher difference scores in attachments to parents and peers predicted lower resiliency $F(1, 400) = 6.18$, $p < 0.05$, and $r^2 = 0.02$.

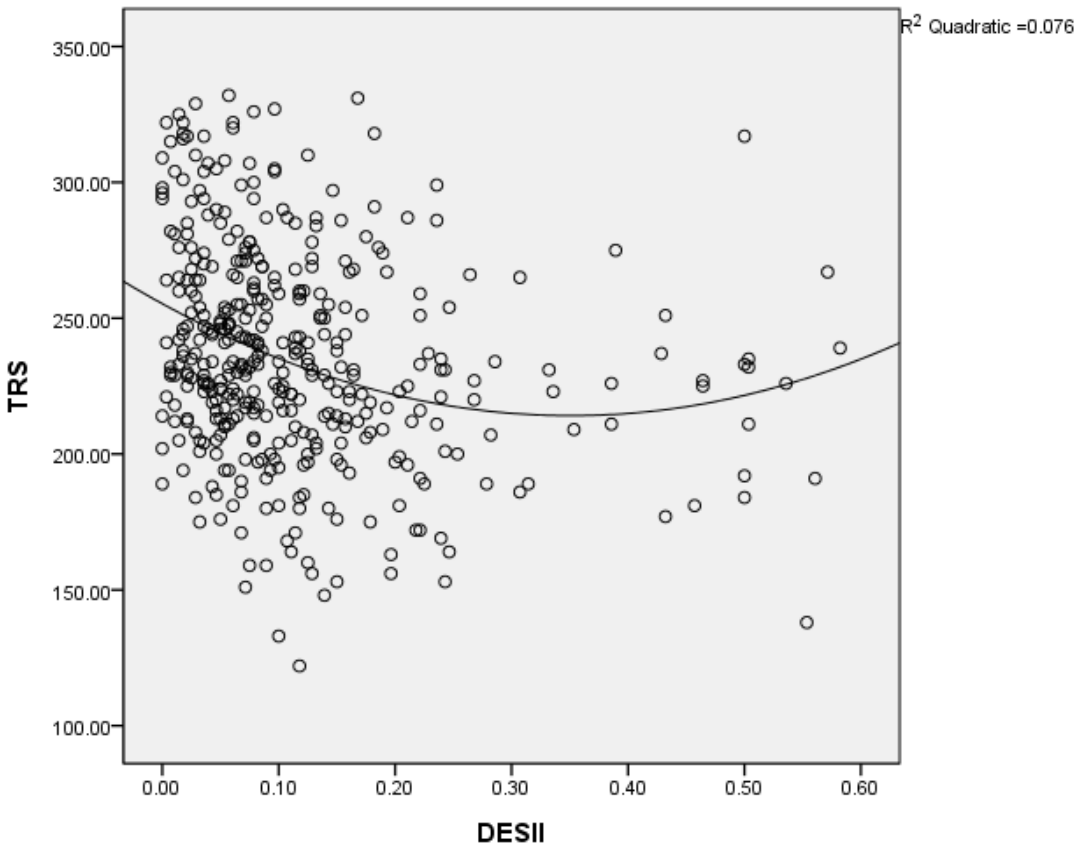

Figure 1. Interaction between TRS score and DESII scores.

Discussion

We hypothesized that moderate dissociative scores would be associated with higher resiliency for participants with a history of betrayal trauma. Moreover, we expected that higher dissociative scores would correlate with lower resiliency in a simple regression model across participants. In regards to this second expectation, the results indicated that higher dissociative scores were significantly associated with lower resiliency. These data support previous research correlating dissociation with negative effects (Freyd, 1996; Loitti, 2004; Crittenden, 2000). A statistically significant effect was also found for the hypothesized curvilinear relationship between resiliency and dissociation in a quadratic regression model. However, this relationship is opposite to our prediction. Findings did not support the idea that people with moderate dissociative tendencies were more resilient to childhood abuse than other groups. Surprisingly, high resiliency scores were associated with participants who had very low or very high dissociative scores but were not associated with participants with moderate dissociative scores. Trying to replicate these findings would be beneficial, as dissociation was only related to higher resilience when scores were at a level that could indicate a possible diagnosis of a dissociative disorder (Briere et al., 2005). This is counter-intuitive, as dissociative disorders are generally thought to be found in people who are the least resilient (Briere et al., 2005). In the current study, dissociation does not seem to relate to higher resiliency, except in very extreme cases.

Regarding the hypothesis concerning differences between parent and peer attachments, participants who experienced betrayal trauma were, as predicted, more likely to demonstrate greater variability in attachment. In a one-way ANOVA, individuals who experienced betrayal trauma were more likely to have greater difference scores in attachment with regards to parents and peers than individuals that did not experience betrayal trauma. Attachment differing between parents and peers is particularly interesting because it indicates that primary relationships may not uniformly affect attachments—at least among people with a history of betrayal trauma. However, this trend does not seem to be related to dissociation as was expected from the literature. A linear regression model indicated that people with higher dissociative tendencies were not more likely to utilize multiple attachment strategies. Therefore, while data supported the idea that people with a history of betrayal would be more likely to have higher difference scores between parents and peers, dissociation did not seem to be involved. Thus, how dissociation protects a child's attachment to his/her primary caretaker is still unclear. Variability in attachment styles in participants with a history of betrayal trauma is still a noteworthy finding.

The finding that participants who experienced betrayal trauma were more likely to have higher differences in attachment scores to parents and peers is noteworthy because it may potentially inform attachment theory. Attachment theory posits that primary relationships formed in childhood structure the models through which all other later relationships are perceived (Bradbury & Karney, 2010). Moreover, considerable research has found that attachment styles may be moderately to highly constant across time. Scharfe (2003), in a review, found that approximately 60%-70% of participants maintained attachment styles and 30%-40% reported change in attachment style when measured across various time periods. However, little research has examined reasons for change in adult attachment styles, although some support has been found for an association between change in attachment style and change of relationship status (Scharfe & Cole, 2006). This finding in relation to betrayal trauma may provide evidence that betrayal trauma history may be another variable that is associated with a higher likelihood of differing attachment styles across different relationships.

These results are also relevant because they may provide additional insight in Bowlby's (1982) hypotheses regarding trauma and attachment. Although Bowlby (1982) expected that attachments would remain relatively stable throughout time, he also discussed changes in attachment. In particular, Bowlby theorized that changes in attachment styles may occur in reaction to traumatic events or experiences (Scharfe, 2003). Bowlby also hypothesized that this response may be adaptive because it would allow an organism to adjust to its environment (Scharfe, 2003). Therefore, the results of this study, which indicate that differences in attachment styles may be more variable between parents and peers when betrayal trauma is involved, could support this idea. These findings indicate that although betrayal trauma predicted lower attachment scores to parents, it did not predict lower attachment scores to peers, as measured by the IPPA. This further supports more fluidity in regards to attachment styles and betrayal trauma. The differential attachment scores between parents and peers warrant additional research. This is positive information for people who have experienced betrayal trauma as it supports the idea that different relationship outcomes are possible—at

least in relation to peers. Furthermore, it is worth exploring what factors could have led to the results that were found and what role dissociation may have in protecting attachments to primary caretakers.

As discussed above, in the context of childhood abuse, dissociation can be seen as a beneficial response to trauma because it allows a child to maintain an attachment bond to his/her primary caretaker (Freyd, 1996). Unfortunately, childhood abuse may also lead to a disorganized attachment style, which is associated with poorer overall outcomes in adulthood (Purnell, 2010). It has also been suggested that a disorganized attachment is still a type of attachment and may, in fact, be an effective strategy for dealing with unpredictable caregivers in situations that are not ideal (Purnell, 2010). Consequently, a disorganized attachment can be seen to be both hazardous and helpful in dealing with an unpredictable caretaker. Because dissociation did not appear to be related to attachment styles, these results indicate that an important distinction may have to be made. By helping a child maintain his/her attachment, dissociation may protect an emotional bond with a caretaker, which is beneficial for survival. However, a disorganized style of attachment, characterized by high levels of dissociation and conflicting strategies (Purnell, 2010), may be a hindrance to future development because it could lead an individual to incorrectly interpret his/her environment. One possible limitation in this study is the fact that although attachment was assessed for both parents and peers, emotional feelings, such as love, were not separately examined. The idea of attachment styles and emotional involvement being inequivalent is supported by previous research (Stein, Fonagy, Fultz, & Target, 2005).

Attachment styles reflect how people view themselves and others, and how they expect other people to respond in a relationship. They do not reflect how connected people feel or how much love a person feels for another. A study by Stein et al. (2005) examined this idea, specifically looking at whether positive feelings for a relationship were being confounded with secure attachment items and vice-versa for insecure attachment styles. The authors found that positive feelings for a relationship (i.e. feeling accepted, enjoying being with a person) were not necessarily indicative of a secure or insecure attachment (Stein et al., 2005). Attachment styles can therefore be seen as internal working models that determine what type of bond an individual may have with another—not the strength of the bond. Therefore, it could be that dissociation is beneficial in that it protects not attachment styles, but rather emotional attachment or attachment strength.

It is therefore possible that cognitive strategies that children sometimes employ, such as self-blame or dissociation, could help preserve feelings of attachment rather than attachment style. Moreover, in a study exploring why women do not choose to leave abusive relationships, Ferraro & Johnson (1983) found that lack of options coupled with emotional attachment and loyalty led women to ignore or rationalize the abuse. Evidently, these women, although probably not securely attached to their partner, still experienced a strong bond. In regards to children who are completely dependent on their primary caretakers, these emotional attachments may be what is preserved when children are ignorant to betrayal. This highlights a key limitation in this study, as the IPPA measures attachment but does not determine an attachment style, nor does it include any measures that consider feelings of love and connection. The IPPA examines three

dimensions of attachment—trust, communication, and alienation—domains that do not clearly examine feelings of love and connection nor do they concretely determine an attachment style. Undoubtedly, feelings of attachment and attachment style would be interesting to examine, as it would provide a more complete picture.

Although dissociation does not seem to benefit attachment relationships as earlier proposed, it may be that the benefit of dissociation resides in protecting feelings of love and connection. To determine if this is the case, it could be interesting to repeat this experiment with some changes. For example, a measure like the Experiences in Close Relationships Inventory Revised (EDR-R; Fraley, Waller, & Brennan, 2005) could be included to measure attachment styles (Sibley, Fischer, & Liu, 2005), and the Rubin Love and Liking Scales (LOV & LIK: Rubin, 1970) could be used to measure emotional attachments (Bailey & Nava, 1989). A replication of this experiment including these scales could provide a more accurate description of this phenomenon by comparing the relation between attachment style and dissociation with the relation between feelings of love and dissociation. Participants with a history of betrayal trauma with higher dissociative scores could be expected to have higher scores on the LOV and LIK scales than those with lower dissociative scores, and attachment styles may not always match up with LOV and LIK scores.

Undoubtedly, dissociation is a common response to traumatic events and Freyd's (1996) betrayal trauma theory explains why such an extreme and outwardly counterintuitive response would occur. How dissociation protects relationships is still not entirely clear. Although the research presented here does not provide clear answers concerning this question, it does provide additional relevant information and brings up issues that require further consideration. The results indicating that higher variability in attachment between parents and peers is associated with betrayal trauma is particularly relevant because it could provide additional information on attachment theory and the effects of trauma on attachment. Moreover, the fact that dissociation did not seem to correlate with this higher variability indicates that dissociation may not affect attachment style. However, this does not mean that dissociation does not protect some element of attachment. Future research should examine whether dissociation could allow a child to protect his/her emotional bond to his/her primary caregiver.

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Signaling for Attention: Mobility and Student Performance in United Way's Promise Neighborhoods

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Abstract

From a middle school student's perspective, the worst part about transferring schools is the need to make new friends. But is that the only negative impact of mobility on students? In this paper, we use a fixed effects linear least-squares statistical regression model to explore the relationship between student academic performance and student mobility in the Bethel School District in Eugene, Oregon. Our client, United Way of Lane County, has struggled with student mobility as the organization refines its new Promise Neighborhoods project, aimed at distressed neighborhoods in Lane County. Student mobility may limit United Way's ability to improve the educational and developmental outcomes of students. We use voter registration data to estimate total mobility in Lane County and in the Promise Neighborhoods. We also use Bethel School District student transfer codes and statewide state test scores as data. Due to the structure of our data, we cannot draw a definitive conclusion regarding the direction of causality between mobility and learning. However, we can say with confidence that, at a minimum, there is a significant relationship between disruption to learning and high levels of mobility – a good starting point for United Way as they continue to explore mobility and refine its Promise Neighborhoods project.

Introduction

United Way is a national non-profit organization that collects and allocates donations to benefit communities in need. Lane County's United Way headquarters is in Eugene, Oregon; this office focuses on serving Lane County. United Way has three primary goals centered on education, income, and health. These goals include preparing children to succeed in school and life, moving families from poverty to financial stability, and ensuring people have access to basic health care. United Way's ability to achieve these goals depends on its ability to understand, prioritize, and best respond to the needs of Lane County residents.

United Way's positive impacts on Lane County residents are a function of its projects and programs. The Promise Neighborhoods, a new project developed by United Way, will develop a continuum of "cradle through college and career solutions" to improve the educational and developmental outcomes of children living in Lane County's most distressed areas (United Way, 2010). Inspired by the success of the Harlem Children's Zone, the Promise Neighborhoods project aims at providing a comprehensive support system for families and children at risk.

Student mobility – a student's physical change of address – is an important topic in education and in United Way's research agenda. Holly Mar Conte, United Way of Lane County's Associate Director of Education, explained that student mobility may become a significant topic on United Way's agenda as they continue to develop a formal strategy for the Promise Neighborhoods (Mar Conte, 2011). Mobility has been a discussion topic in United Way meetings, but little is known about its effect on students or the surrounding community.

Mobility raises two primary concerns for United Way and its goals in the Promise Neighborhoods. First, mobility poses financial impacts. United Way has finite financial resources, and they must make decisions on how best to effectively allocate their money in order to produce efficient and effective programs that generate positive outcomes. The movement of students in and out of the Promise Neighborhoods could threaten to reduce the effectiveness of a dollar spent in the Promise Neighborhoods relative to a dollar spent in a less mobile area. Second, mobility may affect student performance. There is evidence that student mobility has a negative effect on school performance (Beatty, 2010). If there is a high mobility rate in the Promise Neighborhoods, then students who are more mobile will be less likely to be "ready to learn" – a possible phenomenon and major concern in the Promise Neighborhoods that United Way seeks to address.

Based on these concerns, we investigated three main questions: what are the specific mobility rates – within, between, to, and from – in the Promise Neighborhoods? Should United Way be concerned? What kind of data should United Way collect for future studies? We predict mobility has a significant effect on student test scores. In the next section, we will outline our background research on the topic.

Background Information

United Way's Promise Neighborhood is based loosely on the Harlem Children's Zone (HCZ). The project began in the early 1990s, when HCZ organized a wide range of support services for a single block in Harlem, a neighborhood in New York City. The goal was to address typical problems facing poor families: violent crime, chronic health issues, failing schools, and inadequate housing. The single block program eventually expanded to cover a 24-block area. United Way created the Promise Neighborhoods program to mirror the successes of HCZ. The Promise Neighborhoods will focus on two small neighborhoods in the Bethel and Springfield School Districts and will primarily target childhood development and education. The vision is to develop a full continuum of support for children in these areas. Currently, around 82 percent of children in the Promise Neighborhoods entering kindergarten do not meet the early literacy benchmark, determined by United Way as its indicator of "learning readiness" (Mar Conte, 2011). This benchmark includes simple skills such as knowing how to hold a book or recognizing letters of the alphabet.

Specifically, within the district lines of Fairfield and Malabon Elementary schools, which span the Bethel Promise Neighborhood, 92% and 88% of kindergarten students are below the benchmark set by United Way (Mar Conte, 2011). Brattain and Maple Elementary Schools in the Springfield Promise Neighborhood have 76% and 71% of students under United Way's

benchmark, respectively (Mar Conte, 2011). We believe these statistics may exhibit sampling error and bias; United Way only surveyed a handful of kindergarteners, resulting in a small sample size. Nonetheless, they are indicators of the poor academic performance common to the Promise Neighborhoods. Figures 1 and 2 show boundary maps of the Promise Neighborhoods. Figure 1 shows the Springfield Promise Neighborhood, and Figure 2 the Bethel Promise Neighborhood.

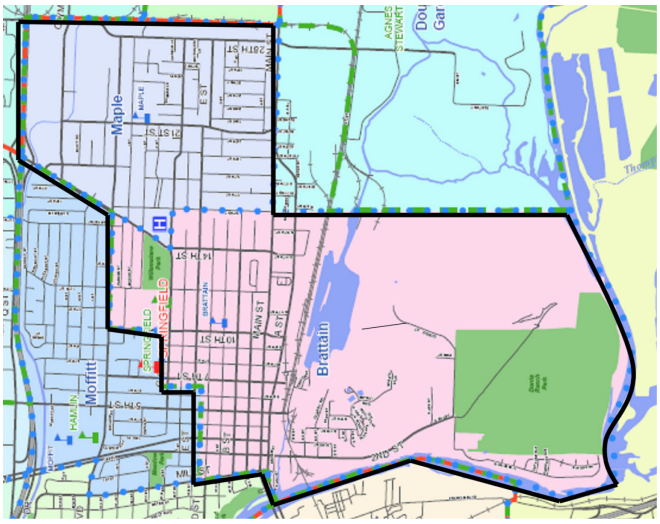

Figure 1. Map of Springfield Promise Neighborhood.

Figure 2. Map of Bethel School District, showing the Bethel Promise Neighborhood.

Having consulted the current scholarly literature on student mobility and performance, we find that the relationship between student mobility and test performance is somewhat unclear. This may be due to the geographic scope of the various analyses. When studies draw from a national database, the effects of mobility and performance dissipate. That is, the more heterogeneous the sample, the harder it is to isolate the effects of mobility on student performance. For example, at a recent conference on student mobility, Lee and Burkam hypothesized that demographic characteristics of children and their families impact mobility (Lee and Burkam, 2009). They used national data to determine mobility rates by race, socioeconomic status, and gender. Lee and Burkam observed that the effects of mobility seem to be small, and that mobility appears less intrusive as an overall effect when one considers an entire population (Lee and Burkam, 2009). On the other hand, effects of mobility and student performance strengthen within local projects or studies that draw from more a homogeneous sample. For example, the Chicago Longitudinal Study finds that one move costs students about two months' worth of achievement, and that students who move three or more times are five to six months behind their peers (Kerbow, 1996). Since this study is based on data gathered in Chicago, consequences of mobility on student performance are more evident.

When scholars do find statistically significant effects, they conclude that mobility has a negative impact on student performance. Highly mobile children perform worse on academic achievement tests than their peers (Beatty, 2010). Repeated mobility – when a student changes schools more than once in a short time period - has a consistently negative effect and its magnitude increases with the frequency of moves (Beatty, 2010). Reynolds and his colleagues found a significant relationship between mobility and both lower school achievement and dropout rates (Beatty, 2010). Additionally, he finds that both early mobility and mobility during high school have the greatest impact on these measures of student achievement.

Yet the literature is likewise unclear on mobility's relationship to specific academic skills. Many scholars suggest students who move repeatedly often fall further behind those of their peers in reading skills (Smith, Fien, and Paine, 2008). At the same conference on student mobility, Lee and Burkam highlighted research indicating that moving had a lesser effect on mathematics performance than on reading performance (Lee and Burkam, 2009). However, at the same conference, Hannaway argued that moving produces consistently negative effects on mathematics scores and marginal effects on reading scores (Beatty, 2010).

According to Beatty (2010), housing problems are the main factor behind students' moves. Forty-three percent of households with children had at least one significant housing problem in 2007 (Beatty, 2010). These problems include housing that is physically inadequate or overcrowded, as well as housing that costs more than 30% of the resident's income. Persistent housing and poverty problems can motivate populations with children to move, which translates into high mobility rates. Forty-two percent of fourth-grade students from poor families changed schools in the last two years, compared with just 26% in non-poor families (Smith et al., 2008). A change in a family's physical residence – for whatever reason – is the main reason why students change schools (Beatty, 2010).

Changing schools involves navigating a complex set of issues that span student learning, classroom instruction, and school organization (Kerbow, 1996). Student mobility not only affects the individual student who moves, but also affects the non-mobile children in the school. Evidence implies that when schools experience high rates of mobility, achievement levels decrease across the board. Lash and Kirpatrick (1990) state that high student mobility rates can also disrupt the learning environment in the classroom and throughout the school This suggests a negative spillover effect: the negative impacts of mobility reside in a school, even after the student has moved (Beatty, 2010). In other words, mobile students also influence the performance of non-mobile students. Student mobility is therefore likely to interfere with instruction, academic skills that build over time, and social networks.

Scholars reiterate that findings on mobility and performance are likely to be conservative (Beatty, 2010). The negative impacts of mobility are actually more pronounced than past studies have shown. Current researchers largely emphasize local studies over national studies to help examine the specific effects of mobility on student performance within a unique area. Thus, piecing together individual studies with robust findings can help shed more light on the dynamics of student mobility.

Data Organization

We utilized two data sets to calculate and interpret the effects of mobility. The first data set was obtained from Lane County Elections and contained the entire population of registered voters in Lane County in 2006 and 2008. This data set included around 400,000 observations and allows for the calculation of mobility rates with respect to age groups and geographical areas for the entire population of registered Lane County voters. Mobility rates were also calculated for the Promise Neighborhoods. It is acknowledged that this data set excludes those who are too young to vote – most of whom are the students and children United Way seeks to serve. However, this data set provides the most extensive information on address changes available. Since most children live with a parent or guardian, an adult moving may give insight into how frequently families move.

Our second data set included withdrawal codes for each student changing classrooms in the Bethel School District. While we focused on physical moves with the Lane County voter data set, we were uncertain whether a withdrawal from a school meant that a student had physically moved and changed addressed. Therefore, we used the Bethel data set to calculate measures of educational disruption. This variable served as a rough proxy for mobility.

Lane County Election Data

Using Lane County voter data, we were able to calculate the mobility rates of registered voters in Lane County. The data set included 394,577 observations across two years with names, precinct numbers, full addresses, and unique voter identification numbers. Using the unique voter identification numbers, we could match observations across the two time periods with very high accuracy. We then determined three possible categories for mobility for each 2006-registered voter over the two-year period:

1. Outbound Mobility: a registered voter had an observation in 2006 but no observation in 2008, and thus was presumed to have moved out of the county. Even if a citizen does not vote, that citizen remains in the data.
2. Churning: a registered voter had an observation in 2006 at a given address and an observation in 2008 at a different address.
3. No Move: a registered voter had an observation in 2006 and 2008 but at the same address.

The Lane County election data had a few yet important drawbacks that decreased the reliability of the estimates. First, there is a possibility that not every individual who moved in Lane County reregistered at his or her new address. Second, outbound mobility includes deaths. Registered voters who pass away are automatically removed from the database within two to three weeks after death. If an individual registered prior to the 2006 cutoff and died between 2006 and the 2008 cutoff, the individual is coded as outbound mobility. Our inability to distinguish between death and outbound mobility causes our calculation to have an upward bias. Conversely, individuals who have lower incomes are more mobile and are less likely to register to vote. According to U.S. Census data on elections and registration, there is a negative relationship between total family income decreases and the percentage of families registering to vote (Figure 3). Furthermore, there is a negative correlation between the duration of time spent in a household and the percentage of families registering to vote (Figure 4). This results in selection bias, causing the mobility rate calculations to be understated. We assert that this downward bias is likely far stronger that the upward bias resulting from voter death.

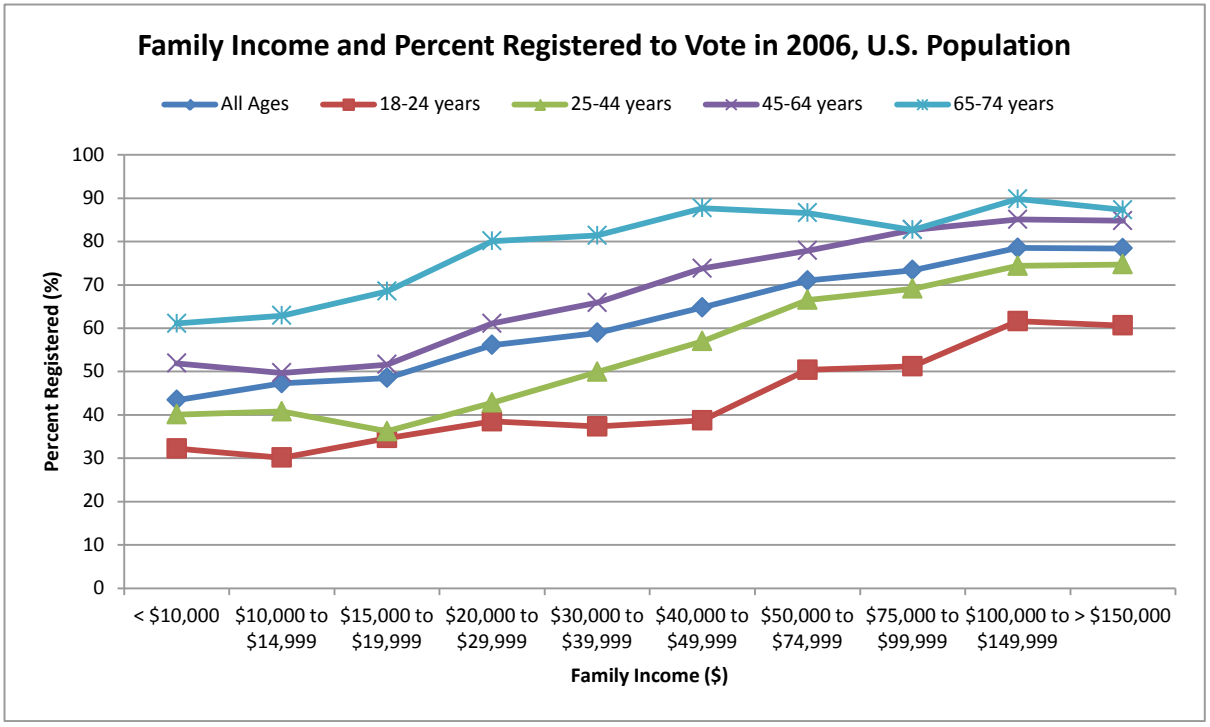

Figure 3. Family income and percent registered to vote in 2006, U.S. population.

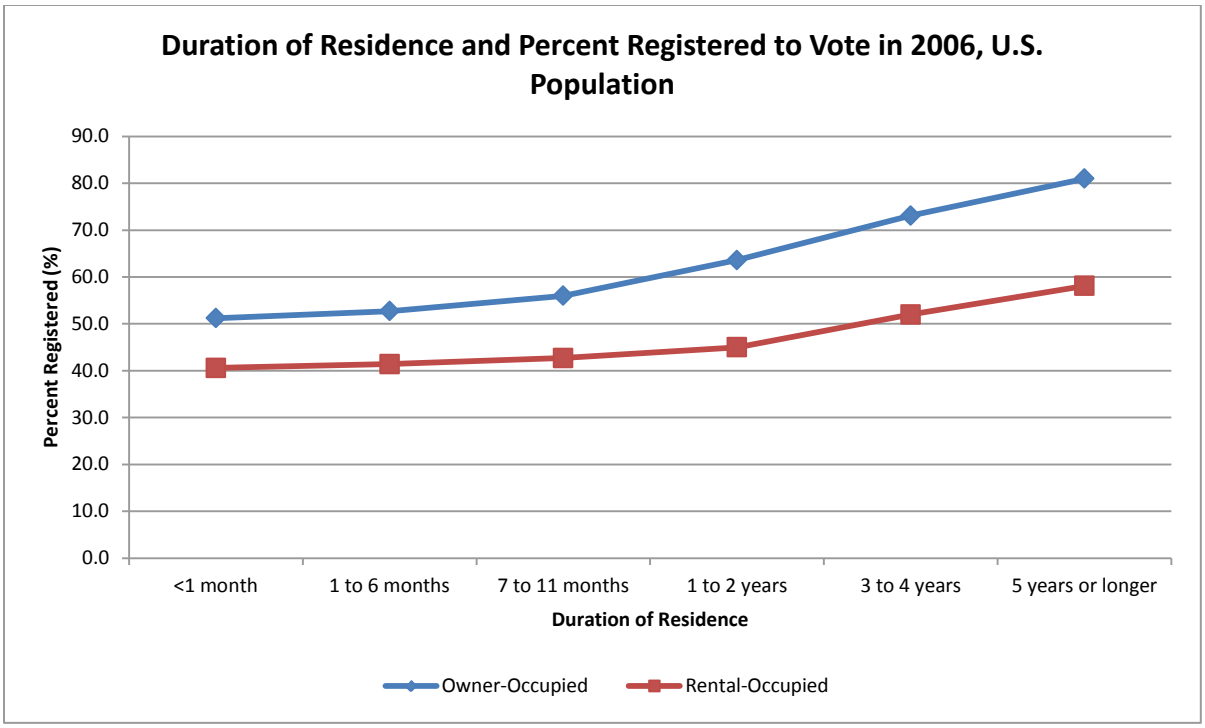

Figure 4. Duration of residence and percent registered to vote in 2006, U.S. population.

Bethel School District Data

We received data that coded each student withdrawal from a particular school in the Bethel School District from 2002-2008. Each code listed a specific reason for why the student left the classroom or school. We created four general groups for the withdrawal codes: private, public, out, and dropout. Observations with codes that did not fit into these categories were dropped from the data set, since they provided no ultimate insight into student mobility, and thus were outside the scope of our project.

1. *PRIVATE*: The student moved from a Bethel School to a private institution in the same area. Over the seven years of data, this code was never used.
2. *PUBLIC*: The student changed schools within the Bethel School District.
3. *OUT*: The student moved out of the Bethel School District.
4. *DROPOUT*: The student was not reported to be attending a new school, or the student simply had stopped attending for a variety of reasons, excluding reported health-related circumstances.

Unfortunately, problems with this data set forced us to transform the data set from an individual unbalanced panel (a data set where the number of years (t) and the number of schools (n) are unequal) into a school-wide balanced panel (a data set where t and n are equal). If the student withdrew from a school and was coded as a move within district, there was no new school code. Therefore, it was unclear where a student who moved within the district actually settled. We also lacked individual demographic information attached to individual students' moves. This would have been important to justify an argument for a direction of causality:

particularly whether mobility affected test scores, or vice versa. Finally, the panel data set was very unbalanced: some students would have multiple observations in certain years and none in other years. With such a low number of total observations, we were concerned whether we would have enough degrees of freedom to pursue meaningful statistical results.

To organize the data into a balanced panel, each individual observation was grouped into school-wide percentages. For example, we determined the percentage of moves within the Bethel School district in school *i* at time *t*. Oregon Department of Education statistics were then used to attach demographic information to each school in each respective period. The total number of students who exhibited the trait at a specific school (*i*) was divided by the total enrollment that specific year (*t*). This produced a percent in terms of the entire student body. We included the variables:

$BLACK_{i,t}$: Percentage of students who self-identify as Black in school *i* at time *t*.

$INDIAN_{i,t}$: Percentage of students who self-identify as Native American/Indian in school *i* at time *t*.

$ASIAN_{i,t}$: Percentage of students who self-identify as Asian in school *i* at time *t*.

$HISPANIC_{i,t}$: Percentage of students who self-identify as Hispanic in school *i* at *t*.

$RATIO_{i,t}$: Ratio of students to teachers at school *i* and time *t*.

$MEAN_{i,t}$: Percentage of students who met or exceeded the state math benchmark at school *i* at time *t* added to the percentage of students who met or exceeded the state reading benchmark at school *i* at time *t* divided by two. This variable compiles math and reading scores, and acts as a proxy for average student performance.

$FREE_{i,t}$: Percentage of students who qualify for free lunch at school *i* and time *t*.

We chose variables that would make it easy for United Way to target specific populations. For example, if we found that Black – or any other ethnicity – caused a decrease in student performance more than other variables, United Way could direct their resources towards a certain ethnicity. United Way could take this approach with other variables, such as students eligible for free lunch.

After preliminary analysis, Kalapuya High School, an alternative high school in the Bethel School District, was dropped from further analysis because its data contained a high number of outliers. Furthermore, we dropped the 2002 time period since not all schools had demographic data available in that year. After organization, the data were in the form of an unbalanced panel data set, with $t = 6$ and $n = 10$; thus, in total, there were 60 observations.

Analysis

The total mobility in Lane County had a maximum of 66% of individuals aged 24-26 old, and a minimum of 20% in the 51-60 year-old age group. In the Promise Neighborhoods, mobility had a maximum of 65% in the 21-23 year-old category, and a minimum of 24% in the 51-50 year-old age group.

In all of Lane County, the outbound mobility reached a maximum at 43% in the 21-23 year-old age group, and a minimum of 10% in the 51-60 year-old age group. Outbound mobility in the Promise Neighborhoods reached a maximum at 42% in the 21-23 year-old age group, and a minimum of 12% in the 51-60 year-old age group.

Mobility within Lane County peaked at 24% in the 27-29 year-old age group, and dropped to a minimum of 8% in the 61+ year-old age group. Mobility within the Promise Neighborhoods had a maximum of 28% in the 27-29 year-old age group and a minimum of 9% for individuals over 61.

Figures 5 and 6 display mobility rates of registered voters in Lane County (Figure 5) and in the Promise Neighborhoods (Figure 6).

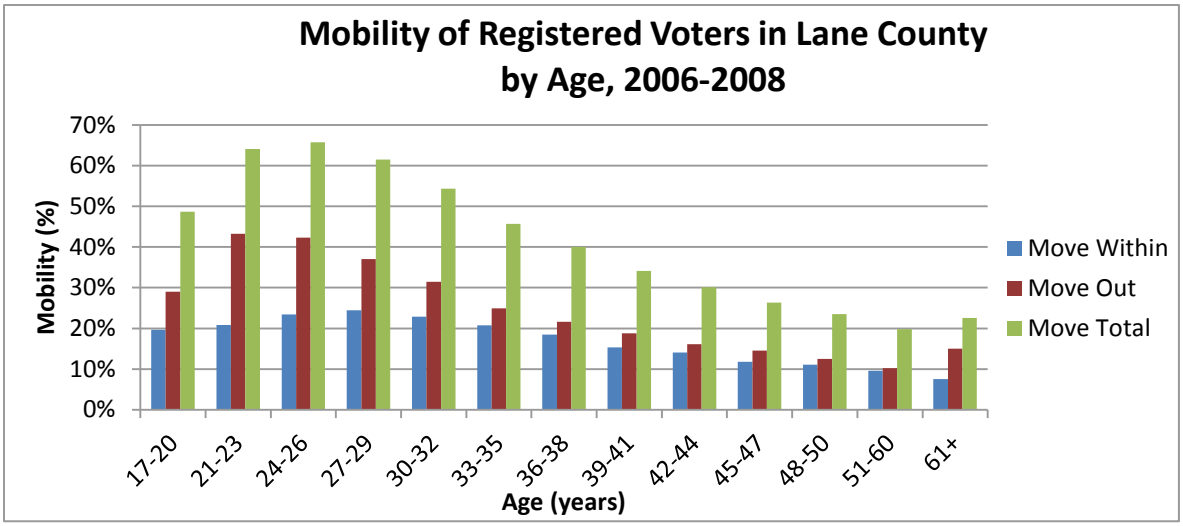

Figure 5. Mobility of registered voters in Lane County by age, 2006-2008.

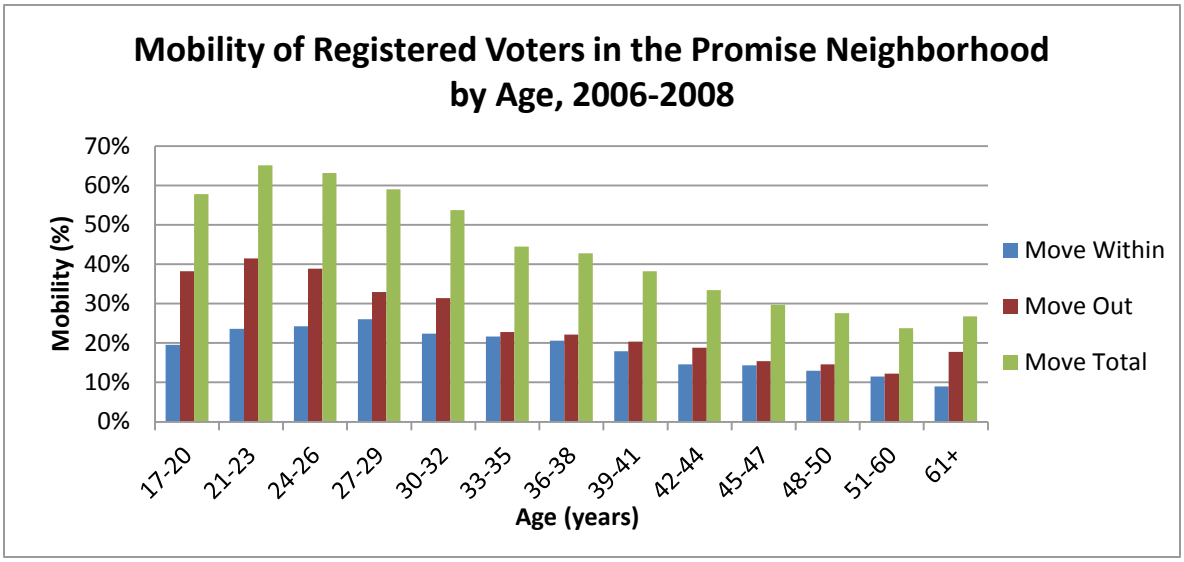

Figure 6. Mobility of registered voters in the Promise Neighborhood by age, 2006-2008.

A multi-variable regression model was used to determine the effects of a list of explanatory variables on a dependent variable. This mathematical technique determines the effects of each explanatory variable and provides a unit-specific numerical estimate of how much that variable affects the dependent variable, controlling for other explanatory variables.

The ability to determine causality between mobility and student performance failed when data was transformed from individual to school-wide percentages. With demographic data attached to each individual who moved, it would have been possible to know which variables were having a unique effect on the dependent variable. Privacy concerns and time constraints kept us from acquiring this type of data. The variables included represent percentages across each school and year. Nonetheless, it is still possible to use the model to test the demographic variables' influence at the school-wide level.

The final two regressions (Tables 1 and 2) use mobility both as a dependent variable and as an explanatory variable. The first empirical model of the Bethel School District data used student performance as the dependent variable. The literature suggests that the casual relationship spanned from student mobility to student performance, and we wanted to explore whether this hypothesis was accurate in the Bethel School District. As noted in the background information, scholars disagree on whether mobility more strongly affects reading or math scores. Considering this disagreement, each school's reading and math test scores were averaged and used as the total performance variable. The second empirical model with the Bethel School District data instead used student mobility as the dependent variable in order to check the direction of causation.

Table 1: Regression model testing the hypothesis that average test scores influenced mobility within the district.

Dependent Variable: PUBLIC

Variable	Coefficient	Std. Error	t-Statistic	Prob.
C	0.052877	0.021617	2.446122	0.0193
RATIO	0.000127	0.000219	0.578006	0.5668
FREE	0.008063	0.008310	0.970264	0.3382
BLACK	-0.072553	0.065588	-1.106188	0.2758
HISPANIC	-0.003308	0.023157	-0.142861	0.8872
ASIAN	-0.009165	0.039450	-0.232314	0.8176
INDIAN	-0.063285	0.059769	-1.058843	0.2965
MEAN	-0.130037	0.055853	-2.328203	0.0255
SQMEAN	0.078516	0.036695	2.139685	0.0390

Cross-section fixed (dummy variables)
Period fixed (dummy variables)

R-squared	0.754803	Mean dependent var	0.002202
Adjusted R-squared	0.609010	S.D. dependent var	0.004081
S.E. of regression	0.002552	Akaike info criterion	-8.820988
Sum squared resid	0.000241	Schwarz criterion	-8.018156
Log likelihood	287.6296	Hannan-Quinn criter.	-8.506956
F-statistic	5.177221	Durbin-Watson stat	2.233768
Prob(F-statistic)	0.000006		

Table 2: Regression model testing the hypothesis that moving within the district influenced average test scores

Dependent Variable: MEAN(1)

Variable	Coefficient	Std. Error	t-Statistic	Prob.
C	0.736662	0.134980	5.457585	0.0000
RATIO	0.007377	0.005238	1.408427	0.1700
FREE	-0.115141	0.146709	-0.784828	0.4391
BLACK	-3.068347	1.625216	-1.887963	0.0694
HISPANIC	0.452954	0.587683	0.770745	0.4473
ASIAN	0.047676	1.193381	0.039950	0.9684
INDIAN	-1.237077	1.241203	-0.996676	0.3275
PUBLIC(1)	-6.866180	3.005855	-2.284268	0.0301
PUBLIC	-6.953119	2.957995	-2.350619	0.0260

Cross-section fixed (dummy variables)
Period fixed (dummy variables)

R-squared	0.916583	Mean dependent var	0.769200
Adjusted R-squared	0.854021	S.D. dependent var	0.132432
S.E. of regression	0.050598	Akaike info criterion	-2.829610
Sum squared resid	0.071686	Schwarz criterion	-1.988319
Log likelihood	92.74024	Hannan-Quinn criter.	-2.509241
F-statistic	14.65066	Durbin-Watson stat	1.627852
Prob(F-statistic)	0.000000		

For both models, linear least squares multi-variable regression with period and cross-sectional fixed effects were used. Since the data was non-random, it was necessary to control for unobserved heterogeneity in the data when the heterogeneity was constant over time. Furthermore, it was necessary to control for differences in the data when anomalies were constant across schools. The fixed effects technique accomplishes both of these goals. In order to cope with heteroskedasticity, or unequal variance in the error terms, White's diagonal robust standard errors was used. This form is the most general technique, and robust to all forms of heteroskedasticity, especially when N and T are small and roughly the same size. To measure goodness of fit, R-squared was referenced and the Akaike Information Criterion was minimized. The significance of each explanatory variable was gauged at the 5% level.

Our first regression (Table 1) used $PUBLIC_{i,t}$ as the dependent variable and tested the hypothesis that average test scores influenced mobility within the district. We regressed all the demographic explanatory variables on $PUBLIC_{i,t}$. To investigate whether there was any non-linear relationship, we also included a non-linear term, $SQMEAN_{i,t}$ which was the square of the $MEAN_{i,t}$ variable. Apart from the constant, the only statistically significant variables ($p < 0.05$) were $SQMEAN_{i,t}$ and $MEAN_{i,t}$.

Our second regression (Table 2) used $MEAN_{i,t}$ as the dependent variable and tested the hypothesis that moving within the district influenced average test scores. We regressed all the demographic explanatory variables, including $PUBLIC_{i,t}$, on a dependent variable that was one period ahead: $MEAN_{i,t}$ (+1). This was done to determine whether mobility ($PUBLIC_{i,t}$) or any of

the other explanatory variables had an effect on mean test scores starting in the next school year. We also included a contemporaneous $PUBLIC_{i,t}$ (+1) to determine if mobility had an effect on mean test scores in the same period. Apart from the constant, the only statistically significant variables ($p < 0.05$) were $PUBLIC_{i,t}$ and $PUBLIC_{i,t}$ (+1).

Results

Several mobility trends are present in both the Promise Neighborhoods and in Lane County. We observed high rates of outbound mobility in younger age demographics, with that trend decreasing steadily after age 30. Additionally, outbound mobility reaches its maximum at 21-23, whereas mobility within reaches its maximum in the 27-29 year old age group. This data suggest that young parents, in age groups anywhere between 20-30 years old, may move more frequently than older parents.

The first regression showed that, across all schools and periods, a school's average test scores had a significant effect on the same school's mobility rate. This linear relationship was negative; a 1% increase in a school's average test scores was associated with a 0.13% decrease in the same school's mobility rate. However, this non-linear relationship was positive; a 1% increase in the square of a school's average test scores was associated with a 0.07% increase in the same school's mobility rate.

Figure 7 contains a graph representing the change in average test scores, holding all other variables in the regression constant.

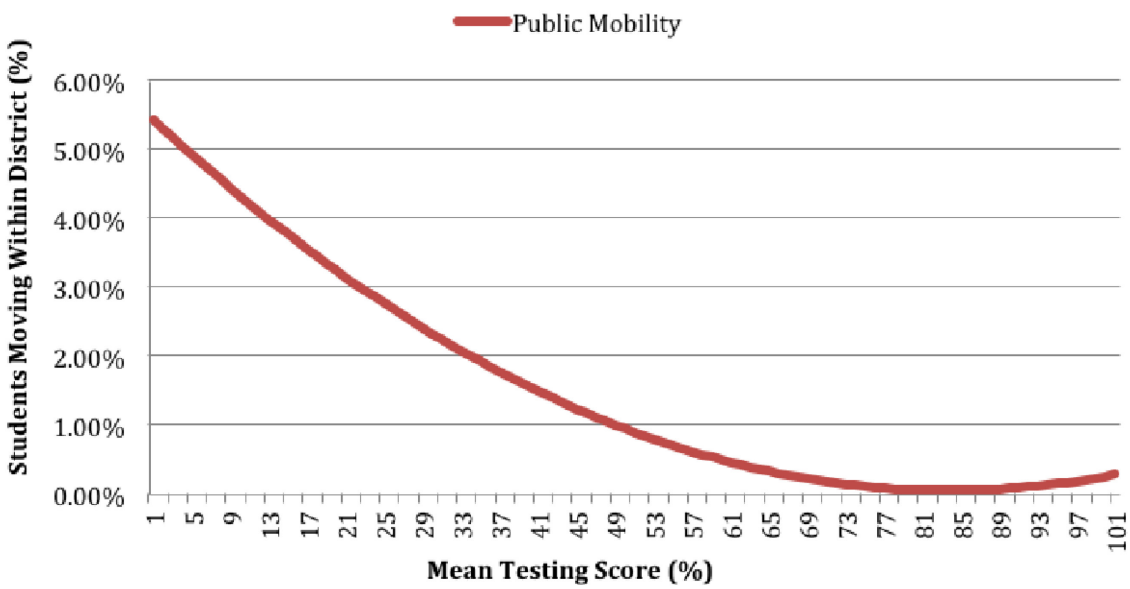

Figure 7. Estimated mobility within Bethel School District.

The negative slope of the line indicates that there is a negative relationship between a school's average test scores and its mobility rate. Furthermore, this relationship exhibits decreasing marginal returns, meaning the curve's slope decreases as average test scores increase. In other words, a percentage change in average test scores is associated with a

decreasing percentage change on the mobility rate as average test scores increase. At a low mean test score, mobility will change more; at a high mean test score, mobility will change less. Hypothetically, if United Way were to spend money to increase test scores with the hope of decreasing the percentage of student moving within the Bethel School District, it would get the most effective use of its recourses by targeting schools with lower average test scores. That said, it cannot be implied that average test scores influences mobility, but rather that the relationship between test scores and mobility exhibits decreasingly marginal returns.

The second regression found that, across all schools and periods, mobility had a detrimental effect on a school's average test scores. Considering the lagged $PUBLIC_{i,t}$, the previous period's mobility rate turned out to have a negative effect on student performance in the following period. A 1% change in the percentage of students moving within the district in one year at a particular school is associated with a 6.95% decrease in average test scores the following year at the same school. Looking at the contemporaneous variable, $PUBLIC_{i,t}$ (+1), moving in the same period had a negative effect on student performance in the same period. A 1% change in the percentage of students moving within the district in one year at a particular school is associated with a 6.86% decrease in average test scores in the same year at the same school. The coefficient values are very similar (-6.95 versus -6.86), suggesting the residual effects of student mobility are closely related to the contemporaneous effects of mobility at the same school.

Due to the generalization of our data, it can only be inferred that mobility, though significant in both regression models, is only a powerful negative relationship. In the first model, it was estimated that average student performance in a certain school has a negative effect on mobility, with the relationship exhibiting decreasing marginal returns. In the second model, it was estimated that student mobility has a negative effect on average student performance in a school, both in the previous period and the contemporaneous period. Both of the models show that there is a strong negative relationship between mobility and student performance. However, the panel data that was used in the regressions were school-wide averages instead of individual observations; therefore, it cannot be concluded that either the act of a student moving causes a decrease of test scores, or that low test scores will cause one to move. It is not known if the students who moved belonged to a certain ethnicity, income level, or had low or high state test scores. Furthermore, if a student moved, we did not know where the student physically moved.

A perfect data set would include a data point for each mobile student, indicating his or her specific demographic data. It would also include that individual's state testing scores. Analysts would also need to code which school the student left, and where in the district a student moved.

Conclusions

The pure effects of mobility and student performance cannot be determined and it was concluded that mobility was merely a signal. The study found that schools with a higher percentage of mobile students have lower average test scores, but it is not known, demographically, what kind of students these mobile students are. Furthermore, we do not know for certain why a school's average student performance decreases. The following are a few

hypotheses that might explain the findings surrounding the relationship between mobility and student performance

- High-performing students are moving out of a school, lowering the school's average test scores.
- Spillover effects: mobile students are negatively affecting non-mobile students' performance at school.
- Low-performing students move to a new school, causing the new school's average to decrease faster than increases in student performance.

Overall, a school's mobility rate signals that either mobile students – or possibly an unobserved demographic group that our data could not measure within the mobile population – need attention. The results suggest that, in the Bethel School District, there is a negative relationship between a school's mobility rate and the respective school's average student performance. United Way must address mobile populations, as the results show they are associated with lower student performance. Although the causal relationship will require further research, we conclude that mobility should be measured and explored to maximize outcomes generated by United Way's Promise Neighborhoods program.

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